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DEALING WITH TECHNICAL PROBLEMS
RELATING TO THE PRODUCTS, PROCESSES AND INVESTIGATIONS OF
N.V. PHILIPS' GLOEILAMPENFABRIEKEN

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A TRANSPORTABLE TELEVISION INSTALLATION

by J. VAN DER MARK. 621.397.7

A description is given of a transportable apparatus suitable for the demonstration of television broadcasting and reception. The installation is housed in two automobile trailers, and is completely equipped for picture and sound reproduction of studio and outdoor scenes as well as of films.

In order to be able to give demonstrations of television at any desired place, the Philips Laboratory has built a portable apparatus, some particulars of which are given here.

The installation, which makes possible the transmission of studio and outdoor scenes as well as of films, is housed in two automobile trailers, each one having a floor surface of about 6 by 2 m. A plan of these two trailers is given in *fig. 1*. In the first is

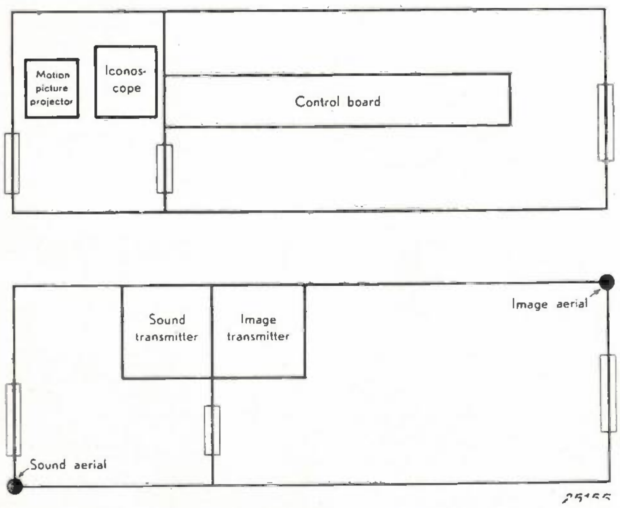

Fig. 1. Plan of the two automobile trailers for the portable television installation.

found the apparatus for the generator of the auxiliary signals and direct voltages necessary for the working of the iconoscope of the recording camera, in addition the apparatus for the conversion of the picture signals from the iconoscope and of the sound signals from the microphone or film, and finally the apparatus for monitoring the picture and sound obtained (*fig. 2*). Two kinds of pictures may be obtained: in a small fire-proof cabin (in the back-

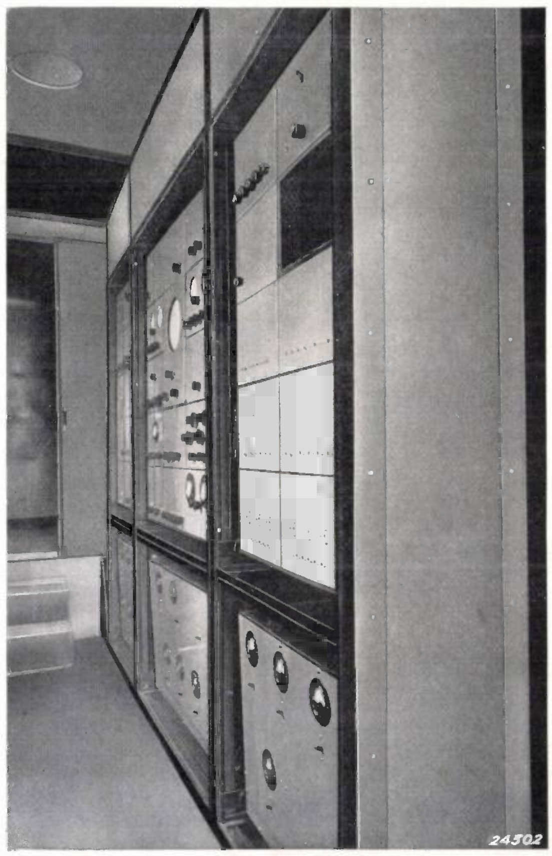

Fig. 2. Operation and control board. In the centre of the middle panel may be seen the front of the cathode-ray tube which shows the control picture. By means of the small cathode-ray tubes to the left and right of the big one, the picture voltages may be checked, in the first place in the form in which they are present in the cable to the transmitter car, and in the second place in the form present in the signal sent out from the transmitter, so that the correct functioning of the transmitter can also be supervised at this point in the first car.

Fig. 3. The iconoscope camera for studio and outdoor work. The base of the camera is provided with a steering gear working on the front wheels, which can be operated by the left hand of the camera man. With his right hand he can direct the camera upon the object with a steering rod. The camera is mounted on its base in such a way that it rotates easily around horizontal and vertical axes. By twisting the handle of the steering rod the lens may be moved forward or back in order to focus the image on the mosaic of the iconoscope.

ground of fig. 2) there is an apparatus for the scanning of films by means of an iconoscope camera (*fig. 3*) which is connected with the trailer by a cable. *Fig. 4* shows the camera in use for outdoor scenes.

The signals from this first trailer are conducted through a cable to the second trailer, in which there are two small transmitters for broadcasting the picture (*fig. 6*) and the accompanying sound. For each transmitter there is an aerial at the top of a demountable mast about 10 m high. *Fig. 5* shows the aerials mounted on the television car.

The installation is suitable for the broadcasting of 25 pictures per second, with 405 or 567 lines per complete picture, while interlaced scanning is employed. (If 567 lines are used, a frequency spectrum must be dealt with which extends from about 50 cycles per second to about 5×10^6 cycles per second, for 405 lines the necessary frequency spectrum extends only to $2.5 \cdot 10^6$ cycles per second.

Since the circuit of a television transmitter has already been described in Philips techn. Rev. 1 p. 16 and 325, 1936, we shall not go into it again. Various refinements have been introduced, but the principle has remained the same.

A small studio set-up is also carried, which consists of an easily demountable framework of steel tubing (*fig. 7*) upon which five water-cooled super high pressure mercury lamps, each of 1 kW, can be attached wherever necessary. These have a closed water-cooling system.

During transport the studio set-up as well as the camera and the aerials can be carried in the transmitter car. All sensitive parts of the installation are sprung, so that they may withstand jarring.

The whole apparatus is fed with alternating current, no generators or accumulators need thus be transported to supply the necessary electrical energy, which must be taken from the local mains. The whole thus forms a complete low-powered television transmitting station. In giving demonstrations several receivers may be set up in the neighbourhood in order to exhibit the picture received "through the ether"; there is, however,

Fig. 4. The television camera in use.

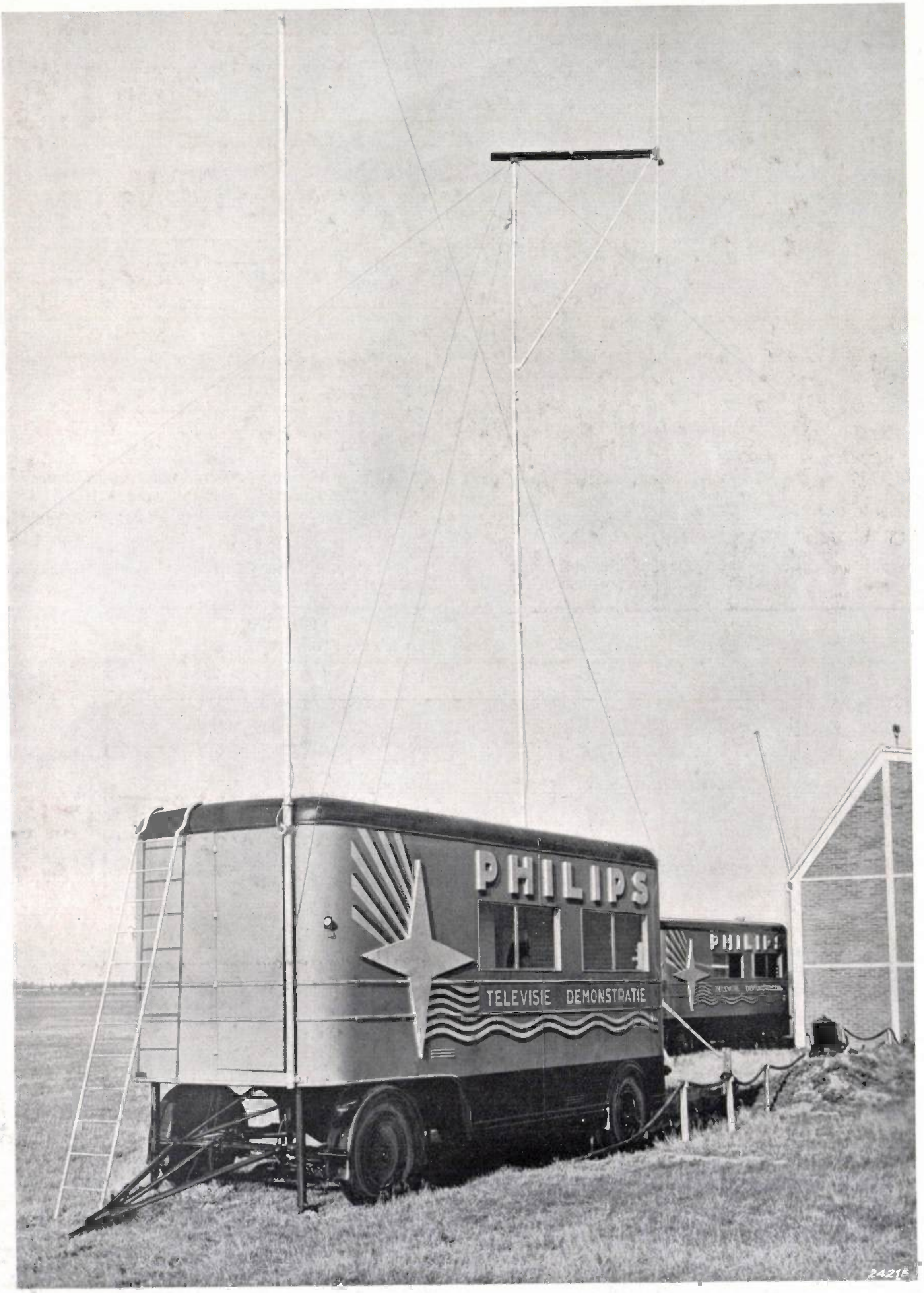

Fig. 5. In the foreground is the transmitter car with the two aluminium aerial masts set up. The aerial proper consists of a half-wave aerial fed at the middle, and is at a distance of a quarter wave length from the actual mast. In the background is the other car. The cable connecting the two cars is plainly visible.

the possibility of transmitting the picture by means of a cable directly to several demonstration receivers, in case local conditions make the use of a radio link impossible.

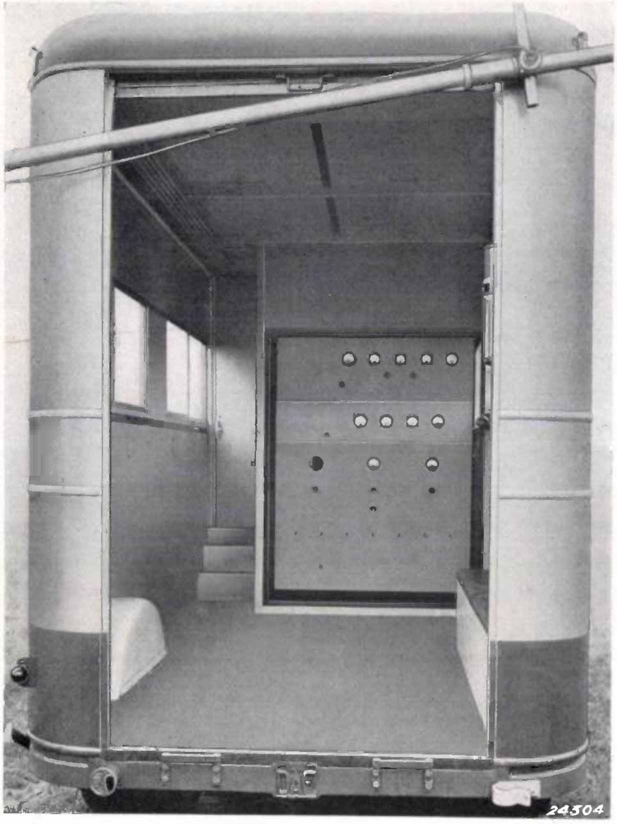

Fig. 6. View of the interior of the transmitter car with the picture transmitter. The sound transmitter is in another compartment of this car, reached by a door visible to the left of the picture transmitter. At the top of the photograph, in front of the shutter which closes the end of the car, may be seen part of the picture aerial demounted for transport. During transport the free space in this car is used for carrying the camera and the studio installation.

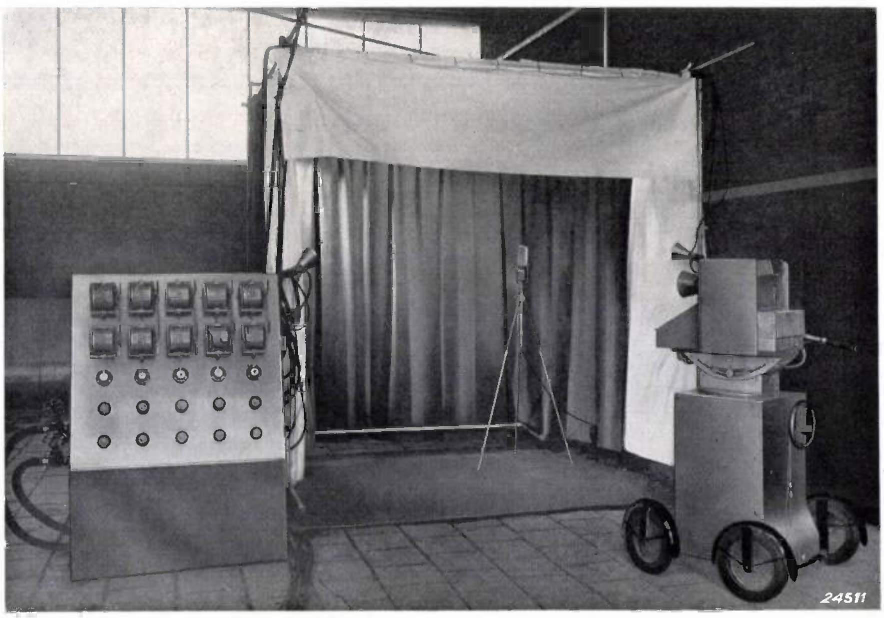

Fig. 7. The studio installation consists of an easily demountable framework of steel tubing, upon which, at any place, super high pressure mercury lamps may be fastened. This framework can be covered with cloth and forms in this way a small studio of a very simple type.
The cabinet contains the apparatus for the supply and operation of each lamp separately, and for the distribution of the cooling water for the lamps. The circulation and cooling of the water is effected by means of a small pump and a radiator.

THE USE OF ULTRAVIOLET RADIATION IN INDUSTRIAL LUMINESCENCE RESEARCH

by A. VAN WIJK.

544.63 : 621.327.31

In the case of many manufactured products abnormalities can easily be recognized by their luminescence. Because of this the observation of luminescence has provided valuable as a check in many industries such as the textile industry, the paper industry, the paint industry, the rubber industry, food industries and laundries. The ultra violet radiation of mercury lamps is used for exciting luminescence, the visible rays being absorbed by a filter. Two pieces of apparatus for luminescence analysis are described and several examples of their application are given.

Luminescence

The impression of colour which a mixed radiation makes on the eye depends upon the relative intensity of the wave lengths in the visible region occurring in that radiation. When a "white" mixed radiation falls on a substance, the reflected radiation may be coloured, that is, in the reflected light the relative distribution of the rays is modified in the visible region.

In many cases the observed change may be adequately explained by noting that each of the incident visible wave lengths is weakened in the reflection by a specific amount depending on the wave length and the substance, so that certain wave length regions make a relatively smaller contribution to the total light reflected.

The influence on the radiation by the substance irradiated is, however, not always confined to a more or less weakened reflection of each of the incident wave lengths individually. With an incident radiation of suitable wave length many substances have the property of emitting a portion of the absorbed energy with an altered, generally longer wave length [1]). If an appreciable length of time elapses between the absorption of energy and its radiation, one usually speaks of phosphorescence, otherwise of fluorescence. The two phenomena collectively are called luminescence. The phenomenon of luminescence may lead to differences between the mixed radiation incident on a substance and that reflected from it, which are essentially unlike the differences which occur in the absence of luminescence. When this property is present wave lengths may occur which are missing in the spectral composition of the incident radiation. The fact that colour changes due to luminescence are not usually observed in daily life, is not due to the fact that substances which exhibit the phenomenon are rare. On the contrary, a great number of substances in our surroundings have this property when they are irradiated with suitable wave lengths.

Nor does the cause of this failure in observation lie as a rule in the lack of these wave lengths in the light used for irradiation, but in the domination of light reflected normally, that is, without change in wave length. This is especially true when "white" light such as daylight or electric light is used. If the incident mixture of radiation is not white but coloured, the conditions for observation of colour components due to luminescence are more favourable, provided the luminescence gives wave lengths which are missing in the incident light.

Luminescence is, however, most striking when a mixture is used for irradiation which contains no visible light. Normal reflection then gives no visible component, so that there is no obstacle to the observation of any luminescence resulting from the irradiation with invisible rays. For the purpose of exciting luminescence in this way only radiation with a wave length shorter than that of visible light may be used, since the wave length increases upon luminescence. The first to be considered is, therefore, ultra violet light, which continues the spectrum at the short wave end of the visible region. Further the ability of X-rays to cause luminescence in substances is quite well known. Frequent use is made of this property in rendering X-rays visible, and in reinforcing the photographic action of X-rays [2]). The use of X-rays for testing materials by luminescence, however, is out of the question for economic and other reasons (danger).

The character of the luminescence, that is the composition of the luminescence light, and

1) This law, called Stokes' law, is related to the fact that the energy of one light quantum is smaller at longer wave lengths. After losing a portion of its energy in reflection, a light quantum can change to a longer wave length, but not to a shorter one.

2) See Philips techn. Rev. 2, 314, 1937.

even the very occurrence of the phenomenon upon irradiation with certain wave lengths, are specific properties of the substance irradiated, and as such are useful in the recognition and examination of that substance. In many cases the result of the method of manufacture or of a preliminary treatment manifests itself in the luminescence. The value of luminescence observation for material testing and process control is being recognized more and more in the most diverse cases in all kinds of industries.

Since the observation is very rapid and does not injure the substance examined or interfere with the process of manufacture, and furthermore since only very small quantities of the substance are necessary, while the results obtained in many cases are remarkably good, it is not surprising that luminescence analysis, as the process is now called, is steadily gaining ground.

Ultra violet sources of radiation

An important condition for the general application of luminescence analysis in industry is the existence of a simple apparatus, suitable for industrial purposes, which supplies a sufficiently intense ultra violet radiation with the exclusion of visible light. In the following we shall give a description of two such pieces of apparatus each of which has its particular range of application.

In both cases the necessary ultra violet light is supplied by quartz mercury lamps. The visible light emitted by these lamps is absorbed by a suitable filter. The filter is made of a special kind of glass "blackened" with nickel oxide.

Table I gives the transmission of two types of glass coloured with NiO for a number of mercury lines. In both cases the transmission for visible light is practically zero (except for the extreme red, where they both transmit again; since mercury lamps give only very little red radiation, this transmission of the red is usually of no great importance).

Table I.

Transmission of two types of "NiO"-glass. The objects tested were bulbs of equal thickness.

λ (Å) 1 Å = 0.1 mμ	Transmission in %	
	Type 1	Type 2
4047	2	2
3655	71	65
3342	68	27
3130	43	3
3022	28	—
2967	10	—

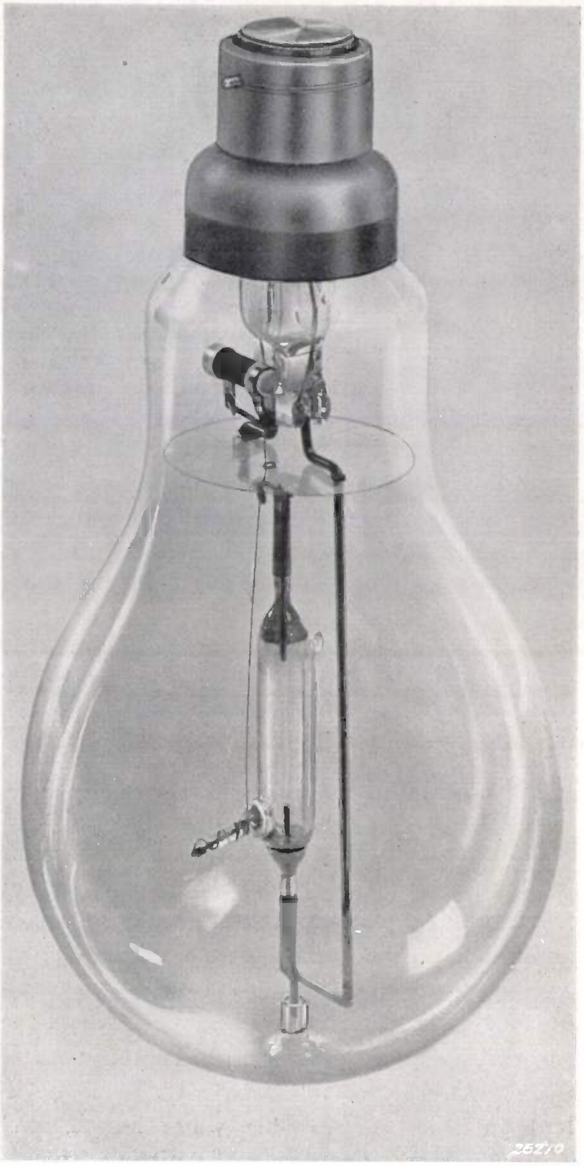

Fig. 1. The mercury lamp 80 watts, 3000 lumens.

The maximum of transmission for both kinds of glass lies at about 365 mμ, where the strongest line of the whole mercury spectrum is situated. On the other hand the two glasses show a great difference at shorter wave lengths. With type 1 the transmission of the wave lengths 313 mμ and 297 mμ is still considerable; with type 2 the transmission is here practically zero. This has been brought about by an extra addition to the molten glass in order to prevent the occurrence of conjunctivitis. Conjunctivitis (snow blindness) is a painful inflamation of the conjunctiva of the eye which can be caused by ultra violet radiation (mountain sunlight). All ultra violet light, however, is not equally active in causing this complaint; the shorter the wave length the greater the action [3]).

[3]) See Philips techn. Rev. 2, 21, 1937.

Radiation with a wave length greater than about 313 mμ no longer causes conjunctivitis even in large doses; it is, however, effective in causing luminescence, so that without important sacrifices in the desired activity of the radiation, the safety of the user may be insured.

The most obvious method of obtaining a filtered beam is to build the lamp into a cabinet and to provide the cabinet with an opening which may be closed by a window of the filter glass. It is clear, however, that in this way only part of the radiation emitted by the lamp is available for use, unless the dimensions of the window are made very great. To this there are technical objections such as risk of cracking of the glass which has to be fairly thick. In the Philips apparatus a different method has been chosen which makes the total radiation of the lamp available by surrounding the lamp with an outer bulb of the filter glass.

As a source of ultra violet radiation the high pressure mercury lamp "Philora" 80 watt (*fig. 1*) and the "Biosol" lamp (*fig. 2*) are used. Both of these lamps have been described in this periodical [4]). The high pressure mercury lamp of very small dimensions is always surrounded by an outer bulb. By making this bulb of the glass containing nickel oxide of type 2 above, a luminescence 80 watt lamp giving 3 000 lumens is obtained which may be connected directly to the 220 V alternating current mains *via* a small choke coil. By placing the lamp in a reflector made of material suitable for the desired application, the greatest possible efficiency can be obtained. Aluminium is a very suitable material for the reflector. *Table II* gives the intensity of the radiation emitted at a distance of 40 cm from the lamp in a direction perpendicular to the radiation element. The input to the lamp is 80 watts.

This lamp is suitable for use in cases where small samples are to be examined one at a time.

For industrial purposes where larger surfaces must be intensely irradiated it is better to use the "Biosol" lamp which gives a higher intensity and has a reflector specially designed to give a wide beam. The "Biosol"lamp, type A, is intended for use with a tubular filter which absorbs ultra violet radiation of wave lengths shorter than 280 mμ. When the lamp is to be used for luminescence analysis this filter must be replaced by a similarly-shaped filter of NiO glass of type 2. The luminescence lamp so obtained is about 3 times as

Table II.

Intensity of the radiation of the HPW-80 Watt lamp at a distance of 40 cm in a direction perpendicular to the radiating element.

λ (Å) 1 Å = 0.1 mμ	Intensity in ergs/sec/cm^2
4047	13
3655	1050
3342	60
3130	32

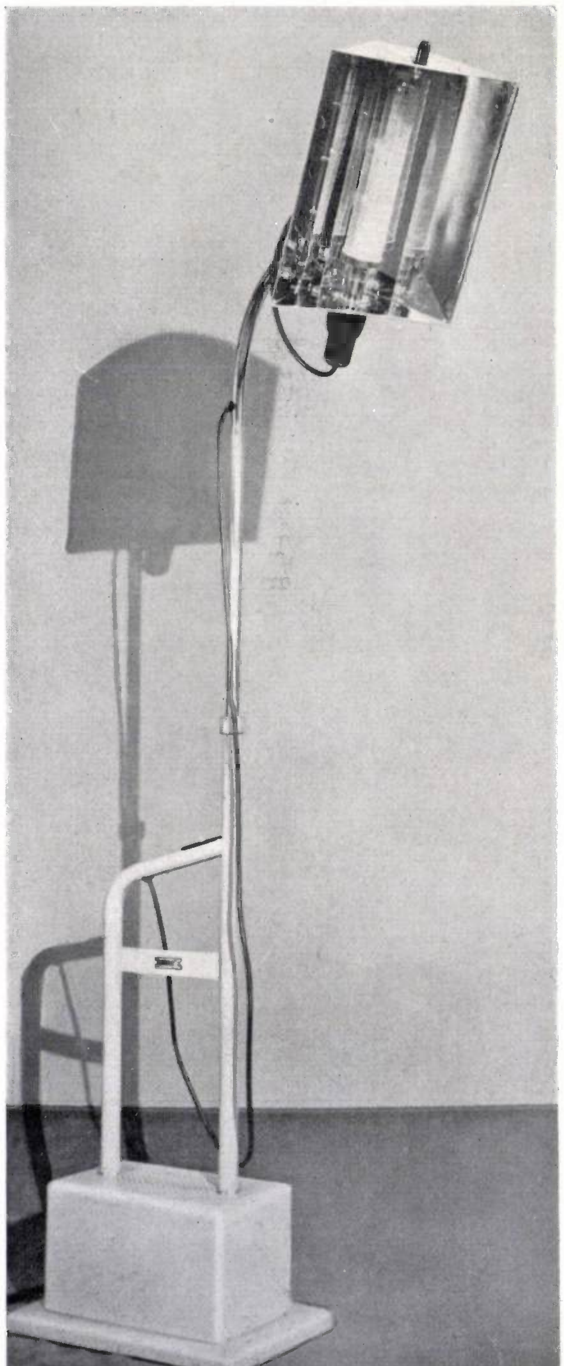

Fig. 2. Philips "Biosol" apparatus type A.

4) The mercury lamp 80 watts, 3000 lumens, Philips techn. Rev. 1, 129, 1936; the "Biosol" lamp, Philips techn. Rev. 2, 18, 1937.

strong as the 80 watt lamp. The chromium-plated reflector increases the intensity in the working direction by an additional factor of about 2.5. At a distance of 50 cm the dimensions of the cross section of the beam are about $50 \cdot 80$ cm^2.

The higher intensity of the "Biosol" lamp necessitates a higher consumption of energy, namely 250 watts. The lamp HPW-125 occupies a position between the "Biosol" A with a NiO filter and the 80 watt lamp. This lamp is similar to the latter but it has an input of 125 watts.

Applications

Several examples of the application of these lamps are given in the following, from which it may be seen that the possibility of irradiating large areas is of practical importance.

Inspection of eggs: As was shown by van Oyen and Molanus the age of hens' eggs may be ascertained at a glance by means of the fluorescence of the shell. A fresh egg fluoresces red and an old egg blue. For checking large numbers of eggs, for instance at an egg market, luminescence analysis is the best method; it, is however, necessary that whole racks can be inspected at once.

Control of laundry: If hard water is used which contains lime, spots of insoluble calcium soap appear in the material. When these spots are fresh they are invisible under normal lighting conditions, while under the luminescence lamp they glow with a bluish brightness. For the inspection of large quantities of laundry the irradiation of large surfaces is desirable.

Invisible marking of linen: A clever application of the phenomenon of luminescence is the following. Instead of disfiguring numbers or other visible marks, the laundry stamps the incoming goods with a number in invisible ink of a kind which luminesces strongly. The ink is fast to washing, so that after being washed the goods may be sorted in the usual way by the light of a luminescence lamp.

Control of coloured textiles: Another interesting application of luminescence investigation is supplied by the textile industry. The printing of materials is often done with soluble colourless reduction products of dyes, the leuco-bases. Only in the following step, oxidation, is the insoluble adsorbed dye itself obtained. Printing faults can, in general, be seen only after the oxidation; they can then, however, no longer be corrected. Practically all the leuco-dyes fluoresce strongly upon irradiation with ultra violet light, so that the printed material may easily be inspected before the oxidation process with the help of the luminescence lamp. In this case also the observation of a large surface is desirable; in many cases, however, the simpler lamp is sufficient, since the luminescing power of the leuco-dyes may be very strong, and an extremely strong irradiation is then not required to make them visible.

It is impossible in the space of this article to give a description of the observations which can be made with apparatus for luminescence analysis on the diverse substances which enter into all kinds of industries. Such a description could not be given without reservations, moreover, because the observations are liable to be found incorrect in certain points upon closer examination. The extremely sensitive method of luminescence analysis has, in common with other sensitive methods, the property of being easily disturbed. In many cases the luminescing power of a given substance is due only to extremely small quantities of impurities, or such impurities may modify the luminescence. Sometimes such impurities will be essential for the use to which the substance is put, sometimes they are completely useless. In the latter case it would of course be incorrect to condemn the substance on the basis of the presence or absence of luminescence, or of luminescence of a changed character.

As examples of industries in which experience has shown the value of luminescence analysis in skilled hands, the following may be mentioned: textile industry, paper industry, paint industry, rubber industry, food industry, laundries.

In addition there is room for application in goods inspection, the detection of all kinds of falsifications as of checques, postage stamps, banknotes, paintings, etc.

It may be mentioned that in many cases the use of an extra filter between the eye of the observer and the substance being examined may prove valuable. This filter serves to absorb the ultra violet radiation reflected from the substance, as well perhaps as the mercury line with a wave length of 405 mμ which is transmitted by NiO glass.

It is incorrect to say that ultra violet radiation is entirely unobservable by the eye. Long ultra violet wave lengths give a certain impression, if the radiation is sufficiently intense. With respect to the limiting wave length there are individual differences, but the mercury line of 365 mμ used chiefly to excite luminescence is observable by practically everyone.

The light impression obtained upon looking directly into the luminescence lamp may be ascribed chiefly to the ultra violet radiation, with the exception of a small amount of red. By introducing a filter as described above, which absorbs radiation with a wave length shorter than about 410 mμ, but transmits longer wave lengths, a truer luminescence colour is observed. Many non-luminescing materials give the illusion of greyish blue luminescence without the filter. The filter can be best introduced directly before the eyes in the form of spectacles of a special glass. Noviol-O or Noviol-A glass of the Corning Glass Works is suitable for this purpose.

The luminescence apparatus may be employed not only for analysis but also for obtaining special lighting effects in shop windows or on the stage. In such cases, paints or textiles, of which there are many, are chosen for their strong luminescence.

Another important application of luminescence lamps may be in the making visible of objects (they may be painted with a luminescent paint) by means of a source of invisible light. For example in cases where there is danger of glare from sources of visible light, this application deserves serious consideration.

"PHILITE" AS A STRUCTURAL MATERIAL

by L. L. C. POLIS.

679.562

Several kinds of "Philite" are discussed with special emphasis upon their application in technology. In the first place their physical properties are described, and the conditions are discussed which must be fulfilled if the substance is to be used as a structural material, while finally various possibilities for the application of this material are indicated.

Introduction

In September 1936 a survey was given in this periodical [1]) of the most important kinds of "Philite", their properties and several possibilities of their application. Before this material can be put into practical use it must become as familiar to those using it as the materials which are now common, such as metals, wood, stone, cement, etc. For this purpose it is at first necessary that the different mechanical, electrical, chemical and thermal properties of "Philite" become generally known.

A table, accompanying this article, has been compiled in which the most important of these values are given. At the same time it may be seen from the table that very many different kinds of "Philite" are made, and that their physical and chemical properties are so widely divergent that it will usually not be difficult to find among all the different kinds a material which satisfies the requirements in any given case.

Since "Philite" must be pressed in a mould, it is possible to manufacture many kinds of articles in this way. If the shape of the article to be made cannot be pressed directly, or if it is one which would demand too great a complexity of the mould, "Philite" will be found to have the additional advantage of being easily worked.

Construction with "Philite"

Because of the high cost of the necessary mould, "Philite" is especially suitable for articles which have to be made in large numbers. Nevertheless, the financial returns on the mould, even from a small number of articles, rises with the increase in complexity of the product, that is with the increase in the cost of making the same product in another way, for example by machining or casting.

For complicated small articles (up to about 100 mm in diameter) (*fig. 1*) the manufacture of 500 pieces may be enough to make a mould pay for itself.

[1]) R. Houwink, Properties and applications of artificial resin products, Philips techn. Rev. 1, 257, 1936.

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Kind of "Philite"	Tensile Strength	Resistance to Concussion	Resistance to Pressure	Modulus of Elasticity	Elongation	Surface and internal resistance in ohms (VDE) [1]		Resistance of 1 cm cube	Breakdown potential [2]	tan δ [4]	Dielectric constant [4]
	kg/cm²	kgcm/cm²	kg/cm²	kg/cm²	%	before tropical test	after tropical test	ohm cm/cm²	kV/mm	10^{-4}	—
S	350	7	2000	80 000	0.5	10^{10}	10^{9}	app. 10^{9}	8	500	4 to 6
O	300	6,5	2000	80 000	0.4	10^{11}	10^{10}	app. 10^{10}	8	400	4 to 6
I	250	5.5	1200	130 000	0.2	10^{10}	10^{9}	app. 10^{9}	1 to 5	> 600	6 to 8
K	350	7	1800	78 000	0.4	10^{11}	10^{10}	app. 10^{10}	5 to 10	app. 600	4 to 6
P	250	6	—	40 000	0.5	$> 10^{12}$	$> 10^{12}$	10^{13} at 25°C	20 to 30	2 to 5	2.4
160 D	300	5.5	—	60 000	0.5	10^{11}	10^{11}	10^{13} at 75°C	—	65	4 to 6
T	300	10	1500	78 000	0.4	10^{8}	—	—	—	> 600	4 to 6
Transphilite	300	5	—	60 000	0.5	10^{11}	10^{11}	10^{13} at 50°C	30 at 15°C 20 at 50°C	150	4 to 6
High tension "Philite"	300	6.5	2000	70 000	0.4	10^{11}	10^{10}	10^{13} at 50°C	30 at 15°C 20 at 50°C	app. 400	4 to 6
Philitext	750	50	2000	35 000 to 90 000	—	10^{9}	10^{6}	—	15 [3]	> 500	4 to 6
Philitax	750	30	1500	60 000 to 100 000	—	10^{10}	10^{8}	—	25 [3]	> 500	4 to 6
Planite	750	30	1500	30 000 to 100 000	—	10^{8}	10^{3}	—	app. 10	> 500	4 to 6

1) Measured according to V.D.E. specification 0302.
2) Alternating voltage peak value (not effective value).
3) For thicknesses greater than 5 mm this becomes 10 kV/mm.
4) Measured at a wavelength of 200 m, that is a frequency of 1.5 mega cycles per second.
5) Measured between 20 and 80°C.

Specific gravity	Linear coefficient of expansion [5])	Thermal Stability		Chemical resistance. All the kinds of "Philite" mentioned in this table are stable to weak acids, weak bases, alcohol, acetone and oil.						REMARKS	Kind of "Philite"
		Long exposure	Short exposure	Water	Strong acids	Strong alkali	Petrol	Benzine	Ether		
—	10^{-6}	°C	°C								
1.35	40	130	200	yes[8])	no	no	yes	yes	yes	Common "Philite"	S
1.35	40	130	200	yes[8])	no	no	yes	yes	yes	Common "Philite", better electrical quality	O
1.85	20	215	220	yes	no	no	yes	yes	yes	Resistant to higher temperature and more resistant to moisture	I_2
1.5	40	90	120	yes[8])	no	no	yes	yes	yes	Suitable for light colours	K
1.04	100	60[6])	60	yes	yes	yes	no	no	no	High frequency "Philite"	P
1.5	—	215	220	yes	no	no	yes	yes	yes	High frequency "Philite", resistant to higher temperature	160 D
1.35	30	95	160	yes[8])	no	no	yes	yes	yes	Quality with higher resistance to concussion (organic fibre filling)	T
1.25	60	200	220	yes	yes[9])	no	yes	yes	yes	Transparent; also good electrical and chemical quality	Transphilite
1.35	40	130	200	yes[8])	no	no	yes	yes	yes	Very high breakdown potential. Not suitable for outdoor use	High tension Philite
1.35	// layers 10 ⊥ „ 80	95	150	yes[8])	no	no	yes	yes	yes	Sheet material, for mechanical purposes only	Philitext
1.35	// layers 10 ⊥ „ 90	130	150	yes[8])	no	no	yes	yes	yes	Sheet material, for electrical uses only	Philitax
1.35	// layers 15 ⊥ „ 50	125	125	yes[8])	no	no	yes	yes	yes	Sheet material especially for mechanical purposes	Planite

[6]) Softening point determined by Martens' method.
[7]) In some cases this value is only reached after heat treatment.
[8]) Upon long contact with water, the water is absorbed.
[9]) Not attacked by hydrochloric acid (sp. gr. 1.19) and phosphoric acid (sp. gr. 1.7) and at 100°C also resistant to a mixture of hydrogen fluoride and sulphuric acid (sp. gr. 1.22).

After the kind of "Philite" has been chosen on the basis of the data given here (cf. table), the shape must be determined with reference to the method of manufacture of the material. In the article quoted above, a diagram of a mould is given, from which may be seen the principle upon which the determination of the shape of the article rests.

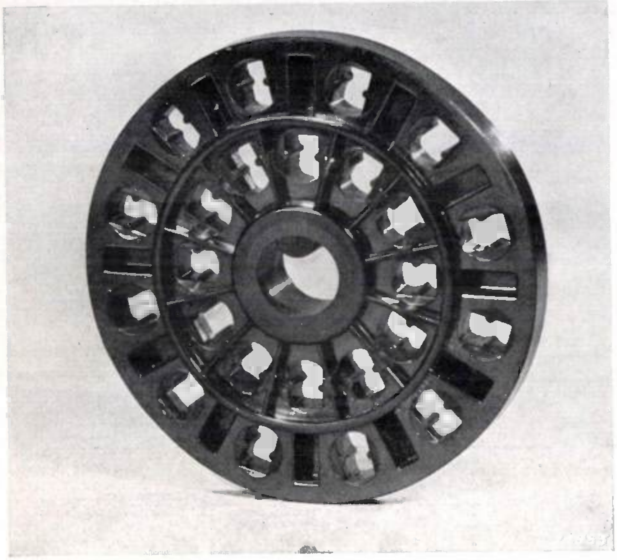

Fig. 1. Disc for switch shaft.

If it is desired to keep the mould and the process of manufacture as simple as possible, and thus as cheap as possible, care must be taken that the product is easily removable from the mould. Lateral projections or depressions are therefore excluded. On the other hand full use must be made of the valuable property of the material of temporary plasticity. Therefore supports, reinforcing and fastening knobs and ribs as well as parts for fastening such as nuts and bolts may be incorporated directly in the mould, so that they are pressed into the product without further expense.

This last-mentioned possibility, consistently exploited, is one of the most important advantages obtained from the use of the material. It means the elimination of loose clips for fastening, and the saving of time in assembling the apparatus of which the moulding is a part. *Figs 2*and *3* give good illustrations of this. In the example of fig. 2 advantage has also been taken of other properties of the material such as its attractive appearance, since this cover forms part of the body of a vacuum cleaner. Use is also made of the fact that the material is a good insulator, since the terminals which serve as connections for the current are pressed directly into the material. In this single example, therefore, advantage has been taken of the mechanical, aesthetic and electrical properties of the material. In *figs 4* and *5* parts of X-ray apparatus are shown in cross section as examples of the shapes which can be pressed from "Philite".

In the construction of articles from "Philite" there are several other points which demand attention. We shall discuss these points one at a time in the following.

The material does not become thin and fluid in the press but remains a tough plastic mass. The shape must therefore be chosen so that the mass need not flow around sharp angles, but so that the whole is stream-lined (*fig. 6*). Fig. 2 also gives a very good example of this. When this principle is not applied, unnecessarily high pressure must be used in pressing, and the risk remains that there will be incompletely filled spaces, called air spots.

Stresses occur in all products which are manufactured under a heat treatment. Since in the case of Philite these stresses cannot be removed by subsequent heating as is the case with glass, special consideration must be given in designing the form to provision for:

a) the greatest possible symmetry in the shape,

b) as near as possible equality in the thickness of the walls,

c) gradual transitions to thinner walls when local thickening cannot be avoided,

d) no sharp angles, and radii of curvature always as great as possible.

Fig. 2. Cover of a vacuum cleaner.

The mould must be divided in the correct way. It is very important that the cross section where the upper and lower moulds join should

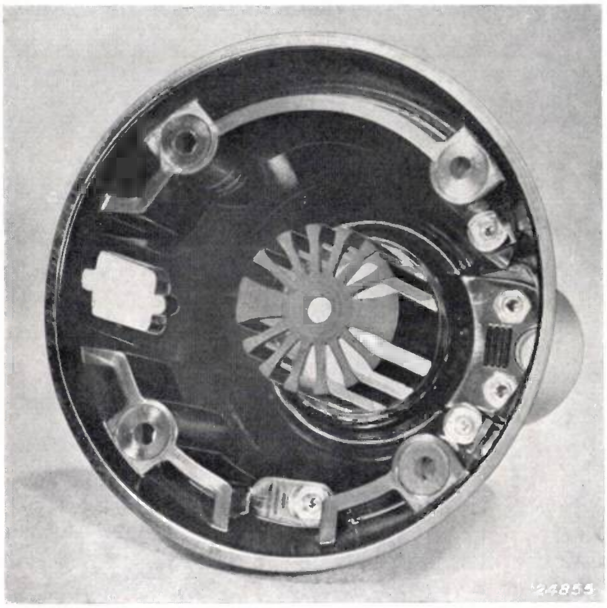

Fig. 3. Cover containing terminals of a vacuum cleaner.

lie at the highest point of the article to be pressed, in order to prevent the enclosure of air which is no longer able to escape and therefore causes air spots in the moulding (*fig.* 7). The edge *A*, which forms a groove in the upper mould is filled with air. During the closing of the press this air must be forced through the press mass. Since this takes place at the very end of the process, that is, when the mould is practically closed, the press material

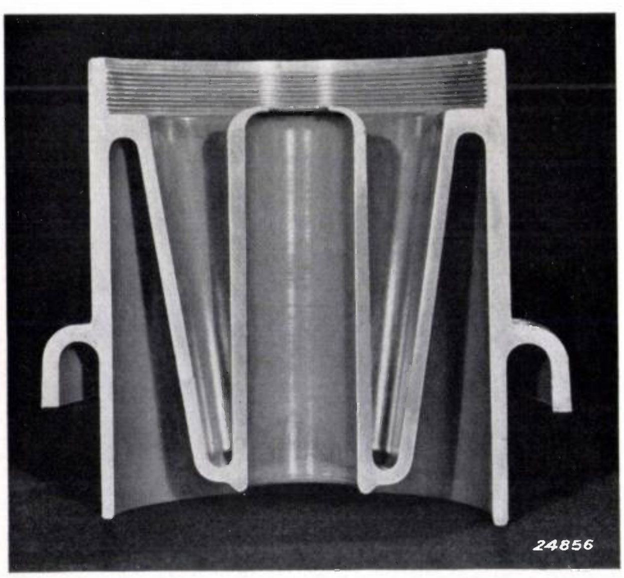

Fig. 4. Cross section of part of an X-ray apparatus.

has a greater tendency to escape through the opening between the two halves of the matrix than to fill the uppermost edge. The air is therefore unable to escape from this groove. In the construction of the mould provisions may be made against the escape of the material; these provisions however also hinder the escape of air. In the case illustrated the flange must be at the very top, or if an edge is necessary (fitting rim) it must be kept as short as possible (several millimetres at the most).

"Philite" is a poor heat conductor, so

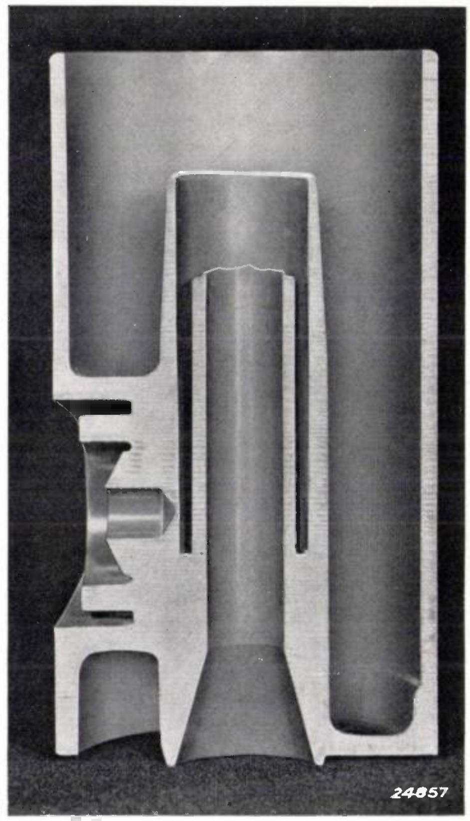

Fig. 5. Cross section of part of an X-ray apparatus.

that the thickness of the material has a very great influence on the time necessary for the material to harden in the heated mould. The material at the core must also be subjected to the high temperature for a sufficiently long time in order to complete the hardening process. This longer time for pressing is of course a factor in the cost price. Normal thicknesses of wall are from 3 to 6 mm; 10 mm must be considered as a maximum.

Deep round holes with a small diameter, so that the depth is greater than twice the diameter of the hole, or any deep narrow holes must be avoided.

When one considers that a pressure of from 300 to 500 kg/cm² must be applied in the mould, it will be clear that the long thin projections of the mould which must form the holes in question would be subjected to enormous forces, and that bending or breaking might easily occur.

The metal parts for fastening, which are to be pressed in, may consist of eyelet-holes, bolts or nuts. There is little to be said about the first two; they must be provided with a head which is enclosed in the mass and which must be so constructed as to prevent turning and pulling out. The head is usually milled or made of hexagonal material to prevent turning, and is provided with a circular groove to prevent its being pulled out. A metal nut which is to be pressed in must be provided with the same guarantees against turning and pulling out. It is however also possible to press the screw thread directly into the "Philite", or as in metals to tap it afterwards. Although the pressed thread is better in quality than the tapped one, the pressing of the thread involves certain risks to the mould which one usually prefers to avoid by tapping. The shaping of the threaded hole must be done with the aid of loose pins which are extracted with the moulding, and then screwed out of the latter and reinserted into the mould. It may happen that these pins are not properly fixed in the mould, or, if they belong in the upper half, that they fall out, and seriously damage the expensive mould upon pressing. The method of tapping the thread will be used preferably when it is to serve for permanent fastening, for instance for the fastening of parts in a Philite radio cabinet.

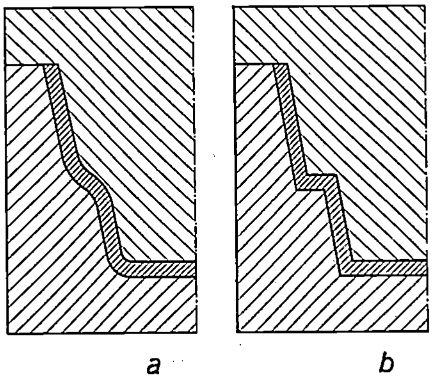

Fig. 6. Construction with "Philite" must be "flowing" as in *a*. No sharp edges as in *b*.

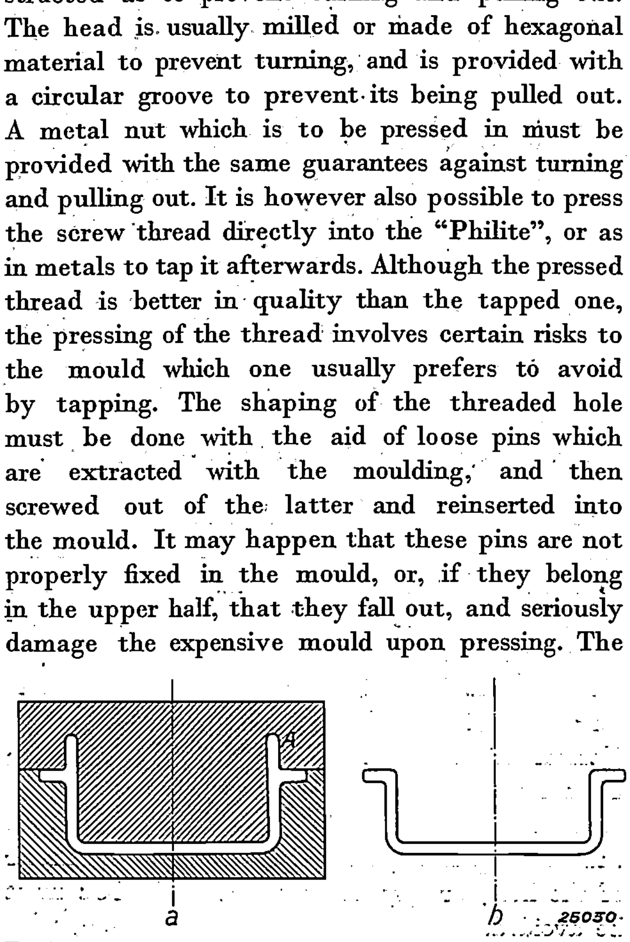

Fig. 7. The flange in *a* is incorrectly placed, so that the "Philite" has difficulty in filling rim *A*. The flange must be as high as possible as in *b*.

One condition for success is, moreover, that the length of the threaded part of the nut should be at least equal to twice the diameter of the bolt. Since the hole must be drilled before tapping, and the tap must have a few millimetres of play, and since also there must be sufficient material below the hole to prevent the drill breaking through (about 3 mm), care must be taken in the design that the projections in which the thread must be tapped are not too small (*fig. 8*).

Lateral projections or depressions can only be given to "Philite" mouldings at the cost of special provisions in the mould and in the method of pressing. Fig. 5 shows the cross section of a moulding with depression on one side. *Fig. 9* shows that spools or bobbins with several flanges can also be pressed from "Philite". The construction of these complicated forms must, however, be studied in each case individually, and can only be carried out with the closest cooperation of the mould maker. It necessitates a rather important increase in cost because of the more expensive mould and the longer time for pressing arising from the complicated manipulations between two successive pressings.

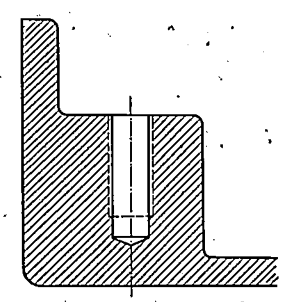

Fig. 8. Sketch of a tap hole in a "Philite" boss.

As has already been mentioned, because of the necessity of an expensive mould "Philite" is especially suitable for mass production. For this purpose it is advisable to try to make the design so that it consists of components which occur several times in the construction; and in addition to construct it in such a way that the same structural elements may be used for different purposes. Good examples of this are the switch shaft disc in fig. 1 and the condenser or insulator elements (*fig. 10*). The latter form structural elements of the Philips high tension apparatus as may be seen from *fig. 11*. Since they are provided at the top and bottom with a screw thread, they may simply be screwed together and form columns 6 to 7 m high, while by means of exchangeable parts in the mould

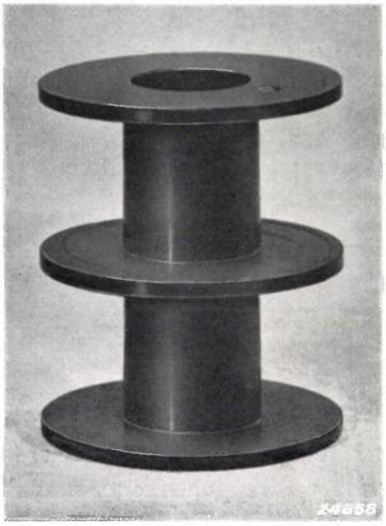

Fig. 9. Cylinder with three flanges.

these containers can be used as condenser pots, supporting and conduit insulators.

Finally we shall discuss still another application of "Philite". Since the material ("Philite" T) has a coefficient of friction of only about 1/10 like good bearings metal with oil lubrication it is very suitable for the moulding of bushes. When we consider the fact that the maker of bushes has up to now worked exclusively with materials which were good heat conductors, it will be obvious that certain pre-

Fig. 10. Containers for condensers or insulators.

cautions must be taken in changing over to "Philite" having a poor heat conduction.

The first success with bushes made of artificial resins was achieved in the steel rolling mills in Germany, where they were used for slowly turning shafts with shaft pressures of 200 kg/cm^2. In connection with the poor heat conduction of the material, bushes of artifical resin must in the first place be provided with a good cooling system by circulation lubrication or with a special water-cooling system. A temperature higher than 80°C is not permissible in the bearing. Special attention

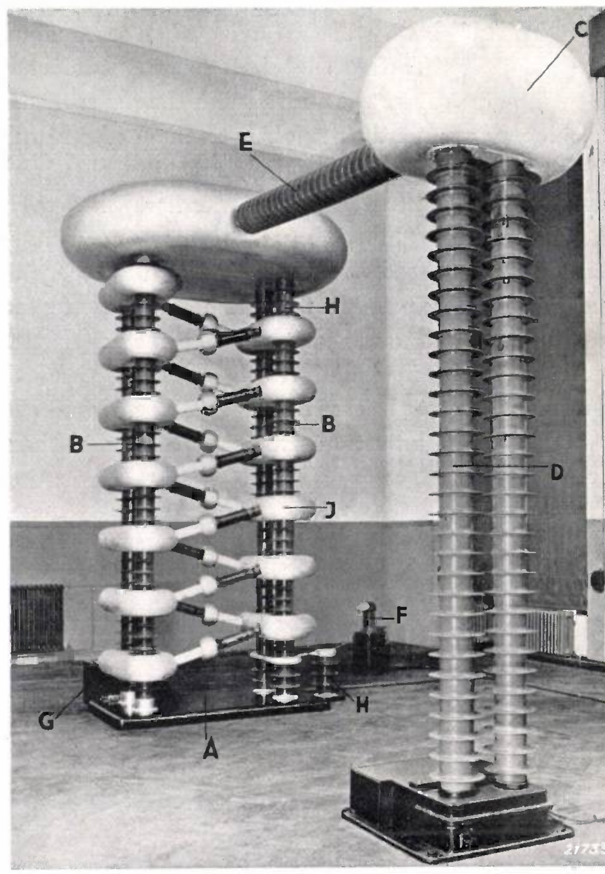

Fig. 11. View of high tension installation (described in Philips techn. Rev. 2, 161, 1937), in which the columns *B* and *D* are composed of the "Philite" elements shown in fig. 10.

is called to the possibility of lubricating these bearings with water, so that in their use in under-water motors and pumps a simplification of the construction is obtained. In view of corrosion and oxidation in the use of water it is, however, better to lubricate with oil or emulsions of oil and water. Furthermore it is desirable to make the bush in the form of a relatively thin shell of "Philite" enclosed in a metal housing, so that the heat can be more easily conducted away.

With metal bearings a clearance of 0.2 per cent is usually allowed between axle and bearing, with bushes pressed from artificial resin a minimum of 0.25 per cent must be taken (in electric motors with small air gap), but where it is permissible the clearance should not be lower than 0.3 per cent.

The usual quantity used to characterize the load on the bush is the product pv of the specific pressure p in kg/cm^2 and the peripheral velocity v in m/sec. With normal lubrication with oil, water or an emulsion this product may be raised to 90 (p up to 40 kg/cm^2). With pressure lubrication it is possible to go as far as $pv = 250$ with the same maximum pressure.

A "Philite" bush will carbonize when it becomes too hot, and therefore the shaft bearing surface is not damaged. Since the shafts with the parts of the machine permanently attached to them often form an expensive part of the whole machine, this is an important advantage. On the other hand, however, the bushes of "Philite" must be more accurately centred than those made of metal. In the case of the latter it is possible for the metal to be forced out upon local overheating, so that a better centring is obtained and the bearings continue to run smoothly. In the case of bushes made of artificial resins, local combustion takes place under these circumstances.

Since products made of artificial resins have increased so rapidly in importance, the workmen who know how to handle this material have obtained their knowledge up to now only by practical experience, that is to say, in those industries in which the material plays such an important part as electrical insulation-material. From the foregoing it will be seen that the material may also be used extensively in other industries. Since, however, in these other industries experienced workmen are still lacking, the execution of the designs will in many cases have to be carried out by the manufacturer of the material. In order to reach the correct solution of a given problem it is necessary that the designer be fully informed about the function of the part in the machine or apparatus, and about all the influences to which it will be subjected. By giving adequate information to the designer disappointments may be avoided for which the material is often incorrectly held responsible. A study of the table will show that the properties of "Philite" are not unlimited, and that therefore its use does not mean the solution of every problem. A critical choice for a practical use is, as with all materials, essential; and when such choice has been made it will be found that with the proper use of the material many problems can be satisfactorily solved.

THE EXPERIMENTAL SHORT WAVE BROADCASTING STATION PCJ

by P. J. H. A. NORDLOHNE.

621.396.712.029.58

A description is given of the way in which the transmitter P.C.J. has been successively rebuilt for the purpose of increasing its output. Details are given of the recent enlargement of the transmitter. This station transmits at frequencies of 9.59 and 15.22 megacycles per sec, and gives an unmodulated carrier wave with an output of about 60 kW. Two triodes of type TA 20/250 are used in the final amplifier stage. There is a discussion of the difficulties which are encountered in the use of very high frequencies (above 2 megacycles per sec) due to the large size of these valves. In addition a description is given of the measures which can be taken to ensure that the valves behave the same in a short wave transmitter as in a long wave transmitter.

Introduction

In the beginning of 1937 the experimental short wave transmitter P.C.J. was again transferred from Eindhoven to the grounds of the Phohi (Philips Omroep Holland Indië) broadcasting station in Huizen. The transmitter was built in the Philips Laboratory in Eindhoven in 1926, and the first experimental transmissions on a wave length of about 31.5 m ($9.5 \cdot 10^6$ cycles per sec), which have since become historical, took place from that laboratory.

The original transmitter was controlled by a quartz crystal. The last stage contained a transmitter valve of the type TA 12/20 000 K; in this last stage the modulation was obtained by means of water-cooled modulator valves type MA 12/15 000. The aerial consisted at that time of a single wire 40 m long slanting upwards.

This transmitter was soon found to cause such interferences in the laboratory that it had to be moved. The choice then fell upon the factory grounds of the Nederlandsche Seintoestellenfabriek in Hilversum, where there was also room available for a more effective aerial installation. After several years of experimentation there, the broadcasting was temporarily stopped.

Meanwhile in Eindhoven a design had been developed of constructing a new transmitter with four water-cooled transmitter valves in the last stage. In addition the transmitter was to be made suitable for 15.22 and 9.59 megacycles per sec, that is, for both the available frequencies.

Furthermore sectional construction was to be employed. For the first stages, which, among other functions, must multiply the frequency of the quartz crystal, a transmitter was used which had been installed in the Carlton Hotel in Amsterdam from November 1930 to October 1931, and had served as an experimental broadcasting station on about 40 megacycles per sec [1]). This transmitter was supplied exclusively with alternating current, and could immediately be connected with 50 cycles supply mains. The transmitter could be rebuilt in a fairly simple way for the frequencies of 15.22 and 9.59 megacycles per sec, giving an output of about 1 kW. In this way the transmitter became the first section of the new transmitter P.C.J. Between this section and the final stage there had to be an intermediate stage with two smaller water-cooled valves (type TA 10/5000 K) where modulation was obtained. The new P.C.J transmitter with three sections has broadcast experimental programs regularly since September 1934, at first only at 15.22 megacycles per sec, later at 9.59 megacycles also. It was set up about 6 km from Eindhoven on the railway line to Utrecht. The aerial consisted of a single horizontal dipole aerial which extended approximately in a North-South direction at about 22 m above the ground. Toward the end of 1936 it was decided to increase the output of the transmitter P.C.J. by adding a stage with two water-cooled valves of the type TA 20/250. These are valves each of which can give 60 kW carrier wave energy at the lower frequencies (less than about 3 megacycles per sec), with the possibility of 100 per cent modulation when used as class B amplifiers. At a frequency of 15.22 megacycles per sec about one half of this output could be counted on. This increase in output was to take place when the transmitter was transferred to the grounds of the Phohi transmitter in Huizen, where there was sufficient space available for setting up directional aerials, while many other improvements also were introduced.

Before passing on to the description of the

[1]) P. J. H. A. Nordlohne, Experimental Radio Broadcast on a Wave Length of 7.85 m in Amsterdam. T. Ned. Radio Genootsch. 6, 1, 1932 (Dutch).

enlarged transmitter we shall give a few particulars of the transmitter before it was enlarged.

First section

This section contains a driver stage which is controlled by a piëzo crystal having socalled „zero temperature coefficient" cut and its frequency thus remains fairly constant without the necessity of a thermostat. The multiplication of the frequency takes place in several steps, followed by two amplifier stages. If the mains voltage were to change slightly this section would need considerable readjustment because it consists of 7 stages. This difficulty was avoided by the introduction of an induction regulator, so that the supply voltage of the whole section can be adjusted in one operation.

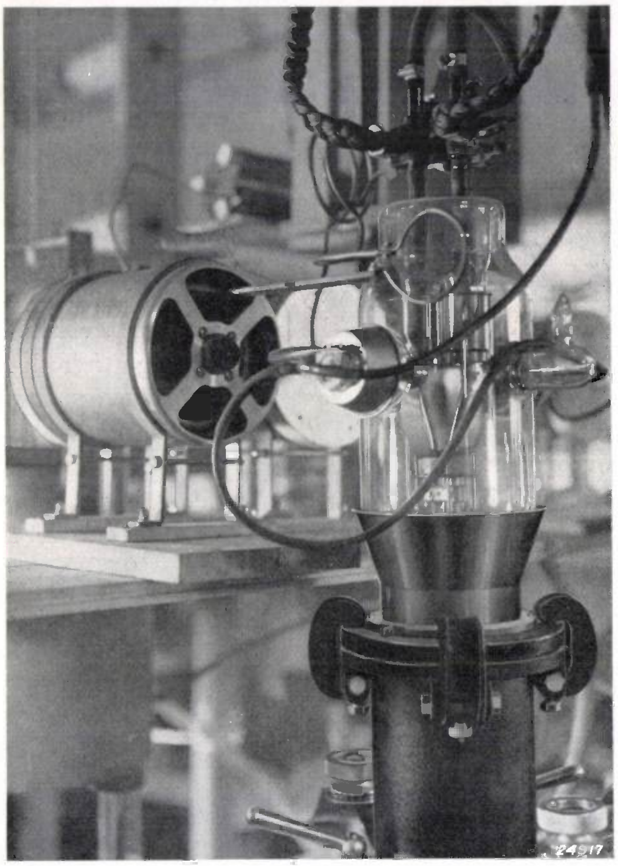

Fig. 1. The water-cooled output valve TA 20/250.

Second section

As already mentioned the second section contains the modulator stage. The grid and anode circuits are carefully shielded from each other. In order to avoid parasitic oscillations, the grid output terminals are connected directly with a pair of resistances, each of which is shunted by a coil of several turns. These coils form a high impedance for a possible high frequency parasitic oscillation but only a very low impedance at the frequencies used, viz. 15.22 and 9.59 megacycles per sec. By choosing suitable dimensions for the chokes between grid and filament, low parasitic frequencies are avoided.

The anode voltage to this stage is supplied by three gas-filled rectifier valves D.C.G. 5/2500.

Third section

The grid and anode parts are shielded from each other as much as possible in this stage also, and the whole section is very compactly constructed. Since the four triodes (TA 12/20 000 K) used in this section have an upper and a lower glass bulb, the mechanical layout of this section was obviously not very simple. The valves are inserted vertically in holes in a shielding plate with the grid below and the anode above, so that outside the valve the grid circuit is shielded from the anode circuit. The filament leads pass through the upper bulb and those of the grid through the lower one.

The constructional difficulty lies mainly in the fact that the valves must be able to be changed quickly. Moreover, there is the chance of undesired couplings due to the relatively strong stray fields of the neutralising condensers. In the P.C.J. transmitter therefore condensers are used which consist of a flat box, as the first electrode, containing a sliding plate which serves as second electrode. The whole is set up vertically in the grid compartment. The upper part of the box is at about the height of the horizontal shield. The box is connected with the grid of the neighbouring pair of valves, while the inner plate which is shielded by the box is connected with the anodes of the opposite pair of valves. This requires rather long connecting strips which are in the anode compartment.

Although no resistances are introduced into the grid and anode circuits, with this arrangement no parasitic frequencies occur over a wide range on both sides of the tuning point of the anode condenser and with aerial coupled. A resistance is introduced in parallel with the grid circuit, and forms the main load on the second section. It also provides that the excitation of the last stage has a linear relation to the depth of modulation, and is thus independent of the grid current load, which varies with the depth of modulation. With this arrangement of the third stage no extra measures are necessary to stabilize the valves.

The negative grid voltage is supplied by a special apparatus (R.S.A.) consisting of a six-phase rectifier with three rectifier valves, type No. 1768. This apparatus is heavily loaded in order to keep the voltage constant during modulation. The R.S.A. is provided with a liberally designed filter to prevent hum.

The anodes are fed by six gas-filled rectifier valves, type D.C.G. 10/15, which contain a grid and are connected in the ordinary six-phase circuit. By means of a three-phase induction regulator, whose stator and rotor windings are completely separated electrically, the phases of the grid voltages of this rectifier are regulated with respect to those of the anodes, so that the voltage can be varied very smoothly between zero and the desired value.

We shall now go on to describe the enlarged transmitter.

The 60 kW stage [2])

The first three sections of the P.C.J. transmitter can give about 12 kW carrier wave energy, which can be modulated for 100 per cent. The energy could of course be increased, as is sometimes done, by allowing several 12 kW transmitters to supply energy to the same aerial. The frequency is then determined by a single piëzo-electric crystal. Five such transmitters could in this way deliver about 60 kW. There would then be twenty triodes (12/20 000 K) in use, as well as many other triodes and diodes.

The more valves in use, the oftener "technical hitches" will occur, since every valve has a limited lifetime. Nevertheless the use of several similar transmitters in parallel may in some cases be justified. For an experimental transmitting station such as P.C.J., however, this parallel arrangement has no particular attractions. As an experiment it is interesting to find out whether this increased energy can be obtained with two triodes of higher power, on the one hand from the point of view of transmitting valve technique, and on the other hand from that of transmitter technique.

For this reason a fourth stage was added to P.C.J. which is provided with two triodes type TA 20/250 (*fig. 1*) which were developed for frequencies of about 3 megacycles per sec. The inter-electrode capacities have such a value that at a voltage of 15 kV and a frequency of 15 megacycles per sec the triodes became too hot, when in preliminary tests each lamp was allowed to develop continuously 100 kW in a self-oscillating circuit. The capacitive currents are then of the order of 100 A and raise the grid connections to red heat. In spite of watercooling all the temperatures were too high, including those of grid and filament supports. At an anode voltage of 15 kV and with 30 kW carrier wave energy per valve, the valves were found not to be overloaded providing that the above-mentioned water-cooling was introduced, and that the glass bulbs were shielded by a metal anode cap and metal grid caps. These caps prevent the occurrence of strong electric fields near the glass, and especially near the chrome-iron to glass joints.

For the use of the transmitter in a broadcasting station, experiments had to be carried out on stability and neutralising. Difficulties were expected because of the very steep slope of the characteristic of the valves and the high stray capacities with respect to earth, which tend to disturb the symmetry of the circuit. These difficulties were found to exist in practice. As a matter of course an attempt was made to build the transmitter as symmetrical as possible with respect to earth. The whole is built on a large copper ground plate, which may have the disadvantage of being a coupling element, but which also has the advantage of symmetry and constancy with respect to earth. Further the water for cooling the anode, as in the second and third sections, is supplied through the middle of the anode tuning coil, which consists here of a short Lecher system of two double copper tubes, two tubes for the ingoing water and the other tubes soldered to the first for the outgoing water (*fig. 2*). The rubber inlet and outlet tubes for the cooling water are thus attached at a point which is neutral with respect to high frequency, so that, provided they are set up at some distance, they have no influence whatsoever on the stability and the tuning. A sliding bridge provides for accurate tuning, while an extra coil between the two anodes serves for rough tuning. The anode voltage is also

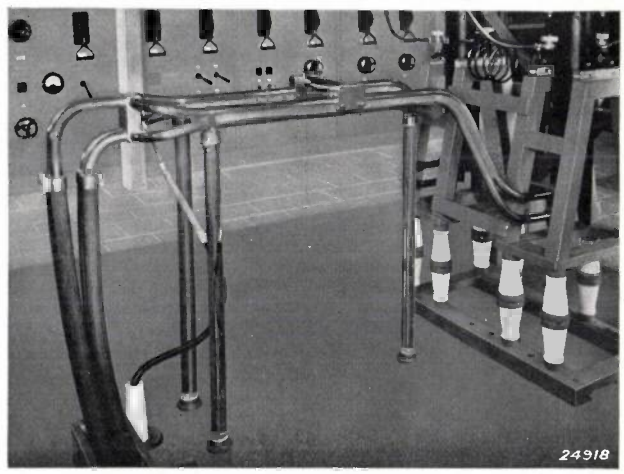

Fig. 2. Lecher system of two double copper tubes which also serve for the water-cooling.

[2]) Various details of the circuit of the 60 kW stage and of the aerial system are due to K. Posthumus.

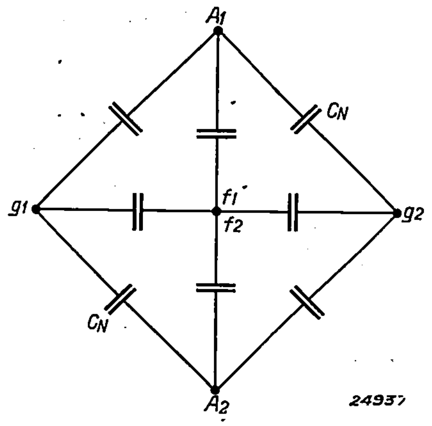

Fig. 3. Bridge circuit with only capacitive branches. A_1 and A_2 anodes, g_1 and g_2 grids, f_1 and f_2 filaments, C_N neutrodyne condensers.

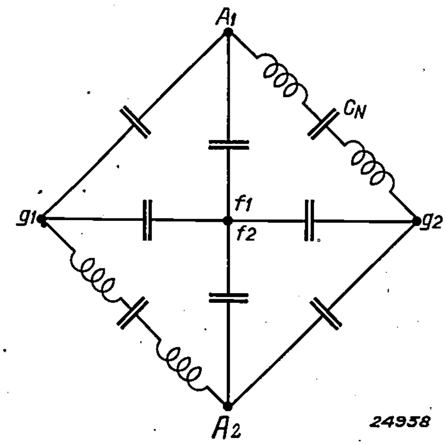

Fig. 4. Bridge circuit in which the equilibrium is disturbed by the presence of self-inductances.

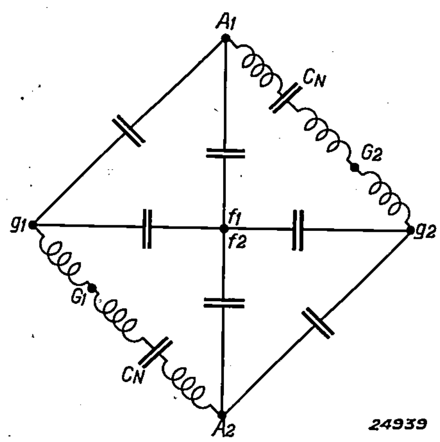

Fig. 5. Bridge circuit in which G_1 and G_2 represent the grid terminals.

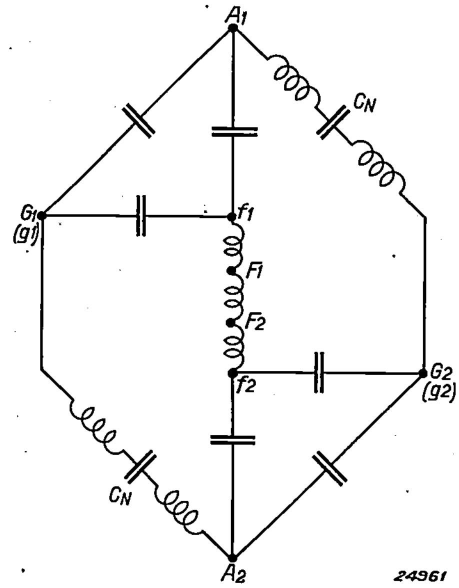

Fig. 6. Bridge circuit in which F_1 and F_2 represent the filament terminals.

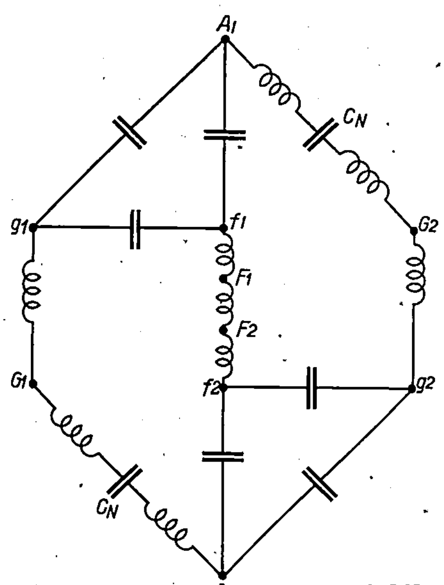

Fig. 7. Bridge circuit in which all possible disturbing self-inductances have been considered.

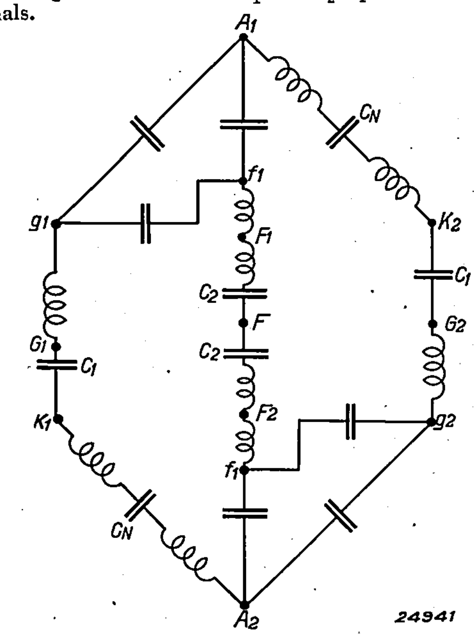

Fig. 8. Bridge circuit in which the filaments and grids are connected with the points F, K_1 and K_2 through condensers C_1 and C_2 which are tuned with the self-inductances of the connections.

supplied at the end of the closed Lecher system, which is the neutral point.

Two water-cooled loading resistances are introduced between the grids of the two valves. Each consists of a nickel iron wire which is wound non-inductively and is placed in a glass tube with inlet and outlet for water. The two resistances are in series, so that the connection between them carries no high frequency voltage. Nickel iron was chosen, because, due to its high permeability, there is a considerable skin-effect. Thus a high resistance can be obtained with a relatively short length of wire. In these resistances about 12 kW from the third stage is dissipated. For this third stage the resistances must form a much greater load than that caused by the variable grid currents of the fourth stage, so that a linear relation is as nearly as possible ensured between the high frequency voltage at the grids of the fourth stage and the depth of modulation.

On attempting to neutrodyne, the first difficulties appeared, and were connected with the use of triodes of such large dimensions with relatively high frequencies. Neutrodyning serves to keep the reaction small, that is, it serves as far as possible to prevent a high frequency current in the anode circuit of a valve from inducing an alternating voltage between grid and filament of the valve, or in the grid circuit connected with it. In general the second condition is fulfilled if the first is satisfied, because the grid circuit is connected directly with the grid and filament. In large transmitter valves and at very high frequencies, however, the inductive reactance of the connecting wires between the valve electrodes and the grid circuit may no longer be neglected, so that different voltages may prevail in the grid circuit from those between grid and filament. In order to demonstrate the solution of this difficulty we shall begin with the fundamental circuit diagram of the two transmitter valves given in *fig. 9*. We shall for the time being neglect the condensers C_1 and C_2 and we are then concerned with the substitute diagram of *fig. 3*. As soon as there are self-inductances in the branches of the neutrodyne condensers C_N, the equilibrium is disturbed (*fig. 4*), but can be reestablished with a different value of the neutrodyne condenser for one frequency. If, however, grid g and/or filament f are separated by self-inductances from their corresponding connecting terminals G and F, as is shown successively in *figs. 5*, *6* and *7*, then by changing C_N it is impossible to prevent the appearance of potential difference between G_1 and F_1 (or G_2 and F_2) or between g_1 and f_1 (or g_2 and f_2) when alternating anode currents are flowing.

In order to meet this difficulty, condensers C_1 and C_2 are added in series with each of the two

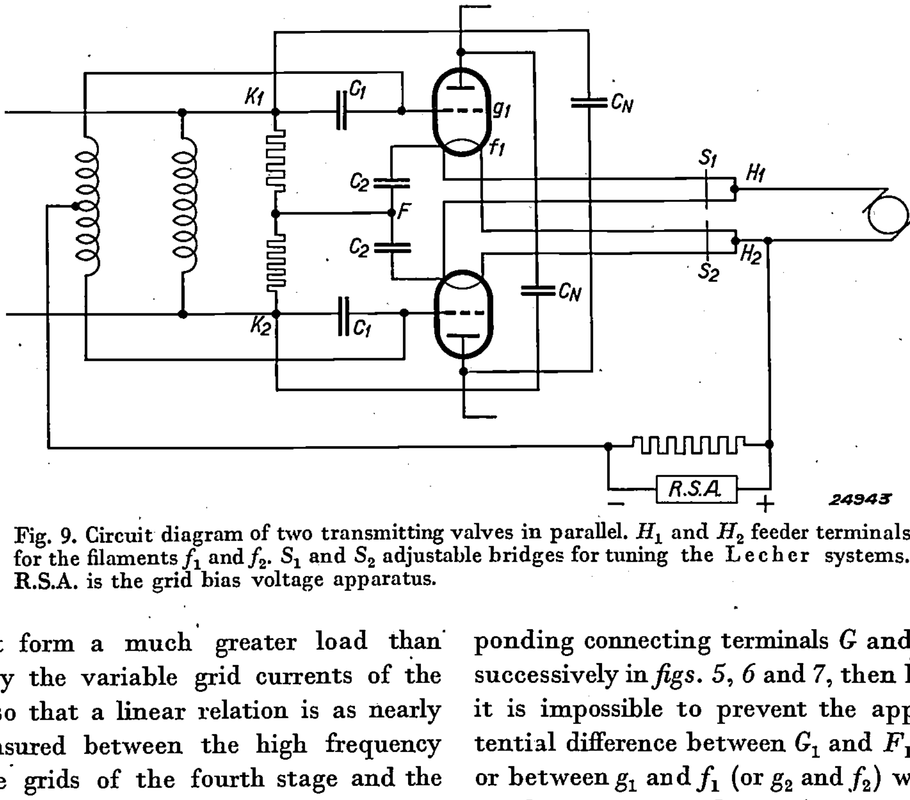

Fig. 9. Circuit diagram of two transmitting valves in parallel. H_1 and H_2 feeder terminals for the filaments f_1 and f_2. S_1 and S_2 adjustable bridges for tuning the Lecher systems. R.S.A. is the grid bias voltage apparatus.

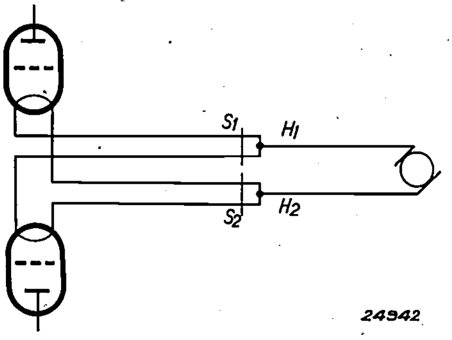

Fig. 10. Detail of the circuit diagram showing the filament supply.

grid connections and with each of the filament connections. These condensers are in series resonance with the self-inductances of the electrode connections for the frequency used (*fig. 8*) so that three points K_1, K_2 and F are thereby obtained outside the valves, which have the same high frequency voltage as the two grids and the filaments respectively. The two filaments are then at the same potential, i.e. at rest from the standpoint of high frequency, when the circuit is symmetrical, which can be brought about in a simple way using a two gang condenser.

For a single frequency, therefore, fig. 8 can be replaced by fig. 4. At this frequency the circuit shows no reaction.

All other connections are now made at the points K_1, K_2 and F thus established: grid tuning coil, neutrodyne condensers, loading resistances and high frequency grid feeders (from the third stage, cf. *fig. 9*).

Since the filament connection terminals F_1 and F_2 are at high-frequency voltage, the source of current for the filaments cannot be connected directly to these points, but the filaments must be fed through two **Lecher** systems, which are tuned to a quarter wave length by means of S_1 and S_2 (*fig. 10*) and therefore form an infinite impedance for the frequency in question. The tuning by S_1 and S_2 is not critical. It is enough if the filaments are joined to each other by means of a constant high impedance. H_1 and H_2 are neutral points at which the machine for feeding the

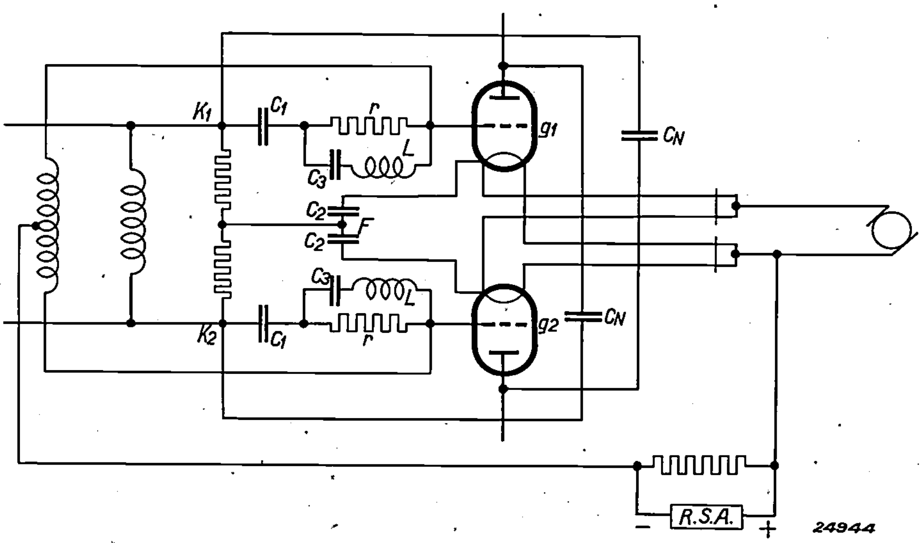

Fig. 11. Circuit diagram of the transmitting valves, in which self-inductances (L) and capacities (C_3), tuned to the working frequency, are introduced between grid (g_1 and g_2 respectively) and grid terminal (K_1 and K_2 respectively).

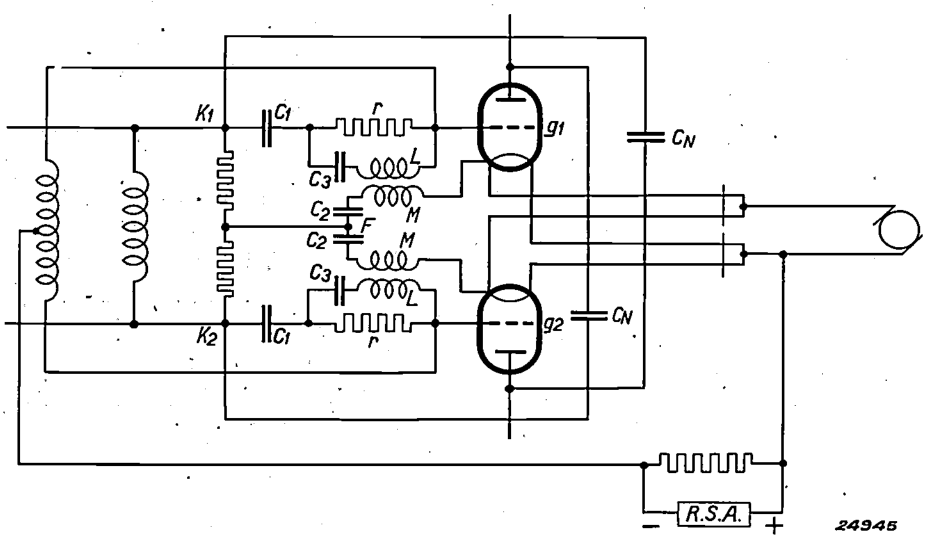

Fig. 12. Circuit diagram of the transmitting valves in which account is taken of the mutual induction M between filament and grid.

filaments may be connected without difficulties.

One of the points H_1 and H_2 is connected to earth and also with the positive terminal of the grid bias apparatus. The negative terminal is connected with the grid terminal through two chokes of copper or resistance wire.

In spite of the fact that a very good neutrodyning is now obtained, it must be kept in mind that stability is assured only for the working frequency. There is no guarantee that parasitic oscillations of very high or fairly low frequencies will not occur, or an oscillation of a frequency in the neighbourhood of the working frequency. Both kinds of parasitic frequencies are suppressed by a resistance r connected directly in series with the grid terminals (*fig. 11*). Such a resistance, however, is also a resistance for the working frequency, and would therefore have to pass the high capacitive currents, as soon as excitation occurred. An air-cooled resistance could, therefore, not be considered. Moreover, such a resistance would absorb excitation voltage. This is prevented by introducing a series circuit consisting of a condenser C_3 with a self-inductance L in parallel with the resistance, which circuit is in resonance with the working frequency. This circuit has a very slight damping, so that even with only a slight difference in frequency it has a higher impedance than the resistance which is in parallel with it. It is actually found that the stage is now quite stable, even with no load on the anode circuit. The above-mentioned series circuit is of course not so sharp that the modulation bands are passed through with decreased intensity; on the other hand, however, the neutrodyning is sufficient guarantee of stability for such small differences in frequency (of the order of 0.07 % at the most) and even for larger differences.

Since the long filament connection is strongly coupled with the grid connection which runs parallel to and at a short distance from it in each of the valves, there is a high mutual induction between anode and grid circuit, so that an undesired back-coupling occurs, and the output cannot be increased. This difficulty is met by compensating these undesired mutual inductions by the introduction of opposite mutual inductions outside the valves (M in *fig. 12*). In the beginning there were still difficulties with the grid and filament series condensers C_1 and C_2 which form the leads of the grids and filaments. These condensers have relatively low capacity and allow high capacitive currents to pass, so that the voltage across the condensers becomes very high with danger of flashover. By giving them proper proportions and especially by rounding off the edges and making the surfaces absolutely smooth, these difficulties were also finally overcome.

The use of two such large triodes for a high frequency means a long step ahead on the road to increasing-output, as was explained in the introduction to this article. There is now nothing to prevent the construction of a broadcasting transmitter for ultra short waves for several hundred kilowatts carrier wave energy. This may be done by combining several 50 to 75 kW transmitters of the above-described type.

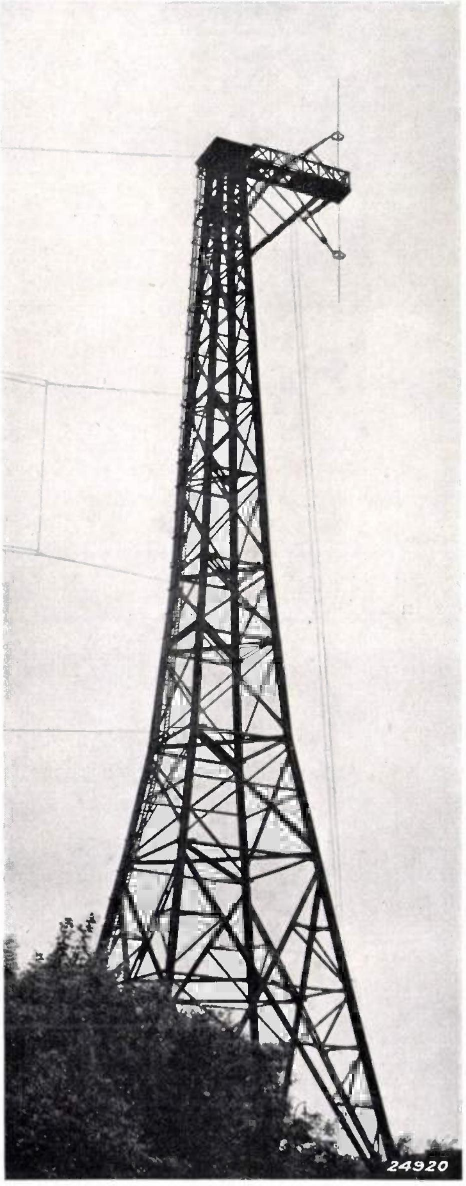

Fig. 13. Wooden mast with platform to which a vertical dipole aerial is fastened.

Aerials

When the transmitter was still set up in the neighbourhood of Eindhoven, experiments were made with a horizontal dipole aerial hung between two masts only about 25 m high. Upon transfer to Huizen an aerial for a wave length of 19.71 m was

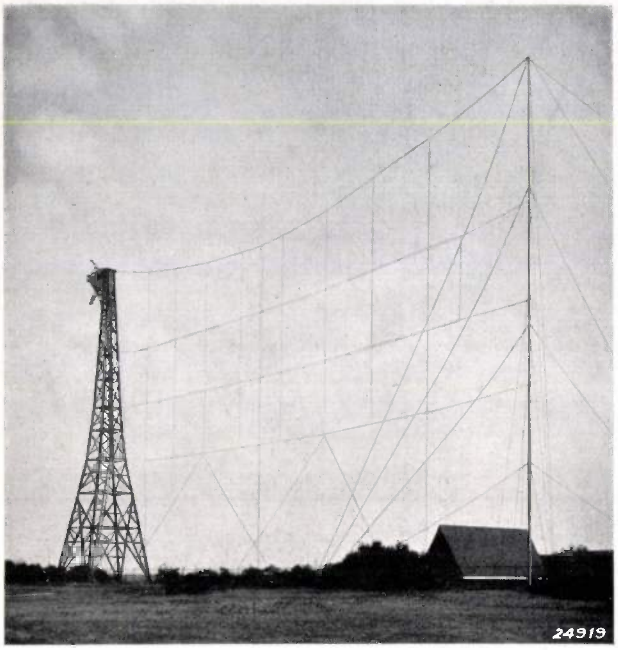

Fig. 14. Directional aerial set up on the grounds of the Phohi transmitter in Huizen.

put into use, which was directed towards the Dutch East Indies, as well as a vertical dipole aerial which was suitable for 19.71 and 31.28 m (15.22 and 9.59 megacycles per sec respectively).

The directional aerial was hung between a wooden tower and a steel Mannesmann mast, both 60 m in height. A tower was chosen in connection with the available space; a wooden tower was chosen

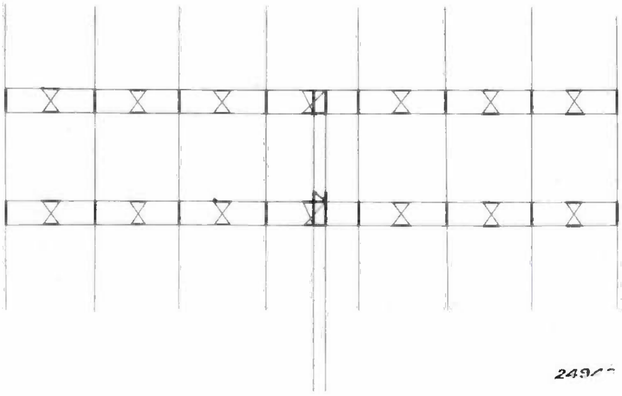

Fig. 15. Circuit diagram for the feeding of the directional aerial.

because the vertical dipole aerial was to be fastened to it, and a steel tower would cause too many losses. In connection with supporting lines and feeders it was necessary to fasten the dipole aerial several metres away from the tower. This made it necessary to build a platform on top of the tower, as well as two arms which slant upwards and downwards respectively, and serve to hold the two ends of the dipole aerial by means of an insulated construction (*fig. 13*). This fastening of the dipole could not be done at a voltage node, because in that case the free end of the rod would be too long in connection with wind pressure and the like. Moreover, the voltage node does not fall in the same place for both the wave lengths to be used. The rods are thus held at a point where some high-frequency voltage is present, so that proper insulation became necessary. The rod is now held between three supporting insulators which lie in a horizontal plane. The insulators are mounted on a heavy brass ring. The two rods are thus firmly fixed and can be fairly easily connected with the feeders which are brought up outside the tower, go through the floor of the platform and then run horizontally to the dipole aerial.

The length of the dipole aerial is chosen so that for a wave length of 31.28 m each of the rods is almost a quarter wave in length and therefore for a wave length of 19.71 m each is about $^3/_4$ of a wave in length.

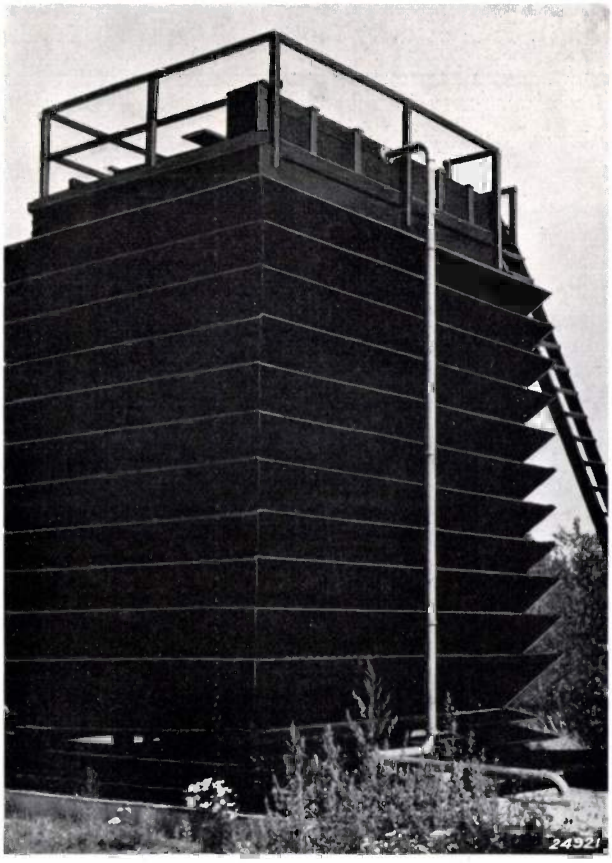

Fig. 16. Scrubbing tower for the water cooling.

The length of the dipole is not critical, if only care is taken that half its length is not more than half the length of the shortest wave to be used, since otherwise opposing currents will be present

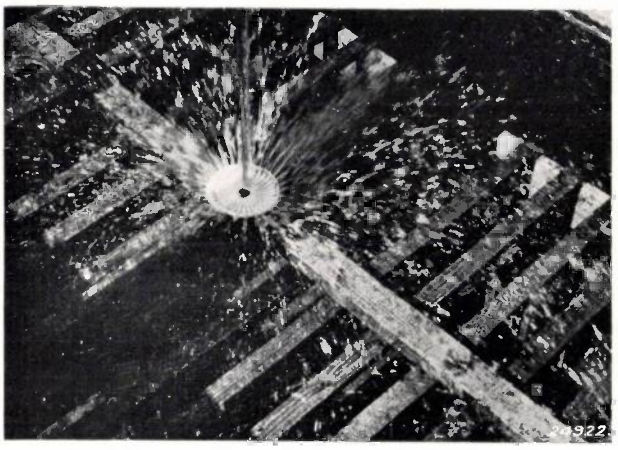

Fig. 17. The cooling water splashes against the wooden latticework of the cooling tower.

in the aerial. Since the half dipoles used in the P.C.J. transmitter are about 7.43 m long, they form for the 19.71 m wave a set of two dipole aerials one above the other, while for 31.28 m they form a single dipole.

The aerial directed toward the Dutch East Indies is for a wave length of 19.71 m and, as has already been mentioned, is hung between the wooden tower and a 60 m high guyed Mannesmann tube (*fig. 14*).

The directional aerial itself consists of 24 vertical half-wave aerials: three rows of eight (*fig. 15*). The

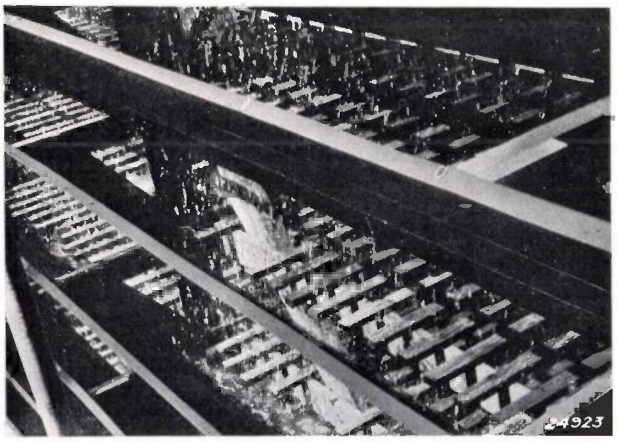

Fig. 18. The cooling water runs through the tower in a finely-divided state.

feeders connected with these aerials run mainly horizontal and feed successively all the aerials at voltage loops of the feeders. Apart from slight losses in the feeder lines themselves, all the aerials are connected with points of equal potential, while the equality of phase is also insured, independent of mutual coupling of the aerials by radiation. The influence of the insulators had previously been studied and also of cross points in the aerial structure on the potential along the horizontal feeders, and it was found that the various lengths must be carefully measured.

Cooling system

With an efficiency of 30 per cent in the triodes of the type TA 20/250, and with an output of about 30 kW per valve, one must count on about 70 kW anode dissipation per triode, and therefore the anode cooling water must remove 140 kW. In addition there is the anode dissipation of the four triodes type TA 12/20 000 K, and of the two triodes of the type TA 10/5 000 K, and finally the heat developed by

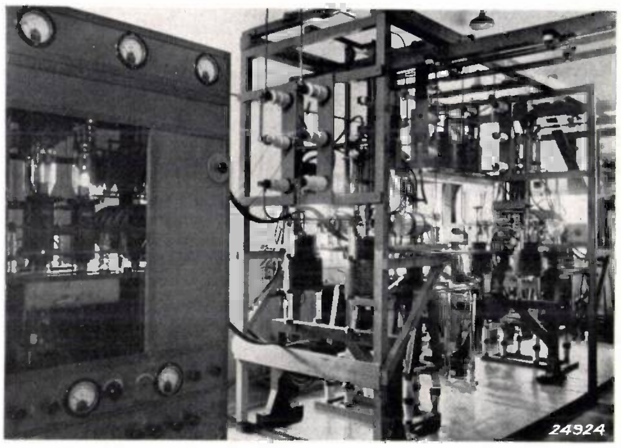

Fig. 19. Third section and end stages for wave lengths of 19.71 and 31.28 m.

the water-cooled resistances, so that the total energy to be removed by the cooling water can be estimated at about 190 kW. A temperature of about 18° C is suitable for the ingoing cooling water. With this temperature there is no trouble with condensation on the lines even in the summer, while with about 240 litres/min for the two triodes type TA 20/250 the temperature of the outgoing water remains far below the highest permissible temperature at which the formation of boiler-scale and similar disturbing phenomena begin to appear. It must be kept in mind, that the highest temperature of the water in the cooling jacket is much higher before the mixing; in the determination on 120 litres/min this is taken into account in the manufacture of the jackets. The remaining smaller water-cooled triodes, as well as the water-cooled resistances, require about 70 litres/min, so that one arrives at a pump capacity of about 300 litres/min.

In order to bring the water which has been heated in the valves back to room temperature, it is allowed to flow through a cooling tower

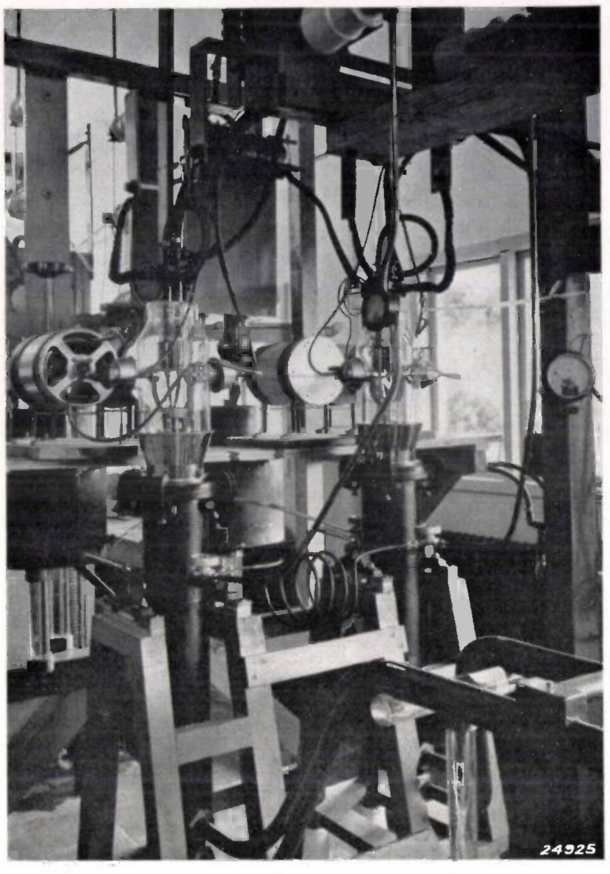

Fig. 20. Closer view of end stage for a wave length of 19.71 m.

(*fig. 16*). The cooling-water is pumped out of a concrete tank containing 44 cubic metres through the cooling jackets of the triodes and returns again to the tank. To prevent a continual rise in the temperature of the water in the tank, another pump raises the water which then falls on a lattice work and is thus divided into drops (*figs. 17* and *18*). In this way an excellent adjustable cooling is obtained. In such a simple system water of course evaporates. This loss is supplied from a well which gives iron- and calcium-free water.

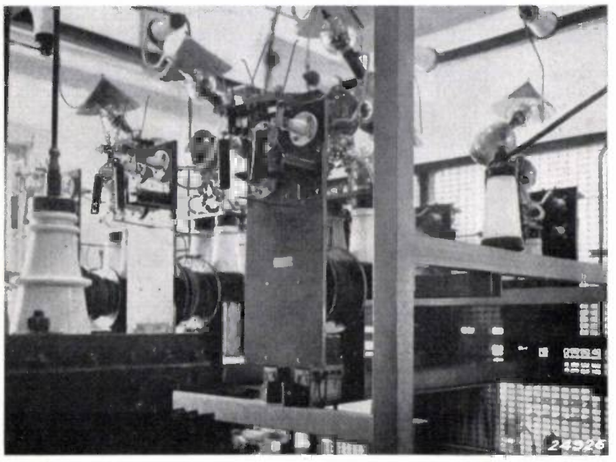

Fig. 21. The rectifiers D.C.G. 10/15 which feed the third section.

Several photographs at the end of this article give an impression of the third section, and of the

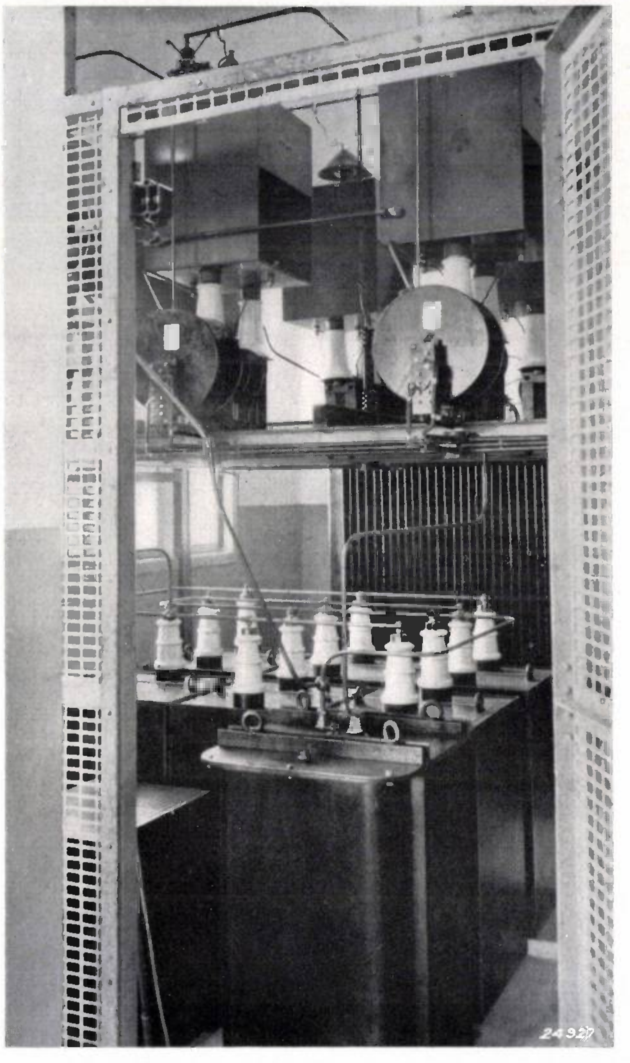

Fig. 22. The rectifier installation for the 60 kW stages.

two final stages for wave lengths of 19.71 and 31.28 m (*fig. 19*). In *fig. 20* may be seen details of the stage for 19.71 m which is still in the experimental state. In this figure the short water-cooled Lecher system with sliding bridge may be seen, as well as the triode TA 20/250, the series condensers C_1, C_2 and C_3 and the coupling M of fig. 12. *Fig. 21* shows the rectifiers D.C.G. 10/15 which feed the third section, while *fig. 22* shows the rectifier installation for the 60 kW stage. The rectifier valves D.C.G. 5/30 are not visible but are situated in the open metal boxes in the upper part of the figure. Below each of these boxes the filament current and starting transformer for the diode above may be seen. On the floor are standing the condensers and choke which form the filter, while in the background part of the high tension transformer is visible.

In the case of the transmitter for 31.28 m which has recently been put into service, use is made of a new kind of directional aerial, whose direction can be varied. A description of this aerial must, however, be given in a later article.

THE UNIVERSAL DECIMAL CLASSIFICATION *)

by L. C. J. TE BOEKHORST. 025.45

A librarian must be able to give a reliable answer when asked what literature is present on a given subject. Besides the subject catalogue of the works present in the library, he refers to the different reference indices and encounters the difficulty that they all make use of different systems of classification. It is therefore not to be wondered at that there has been for some time an effort to unify the various systems of classification.

In 1876 the American librarian Melvil Dewey [1]) published a classification of all sciences according to a decimal system. This system was revised in Europe by the Institut International de Bibliographie in 1895, elaborated and introduced as the "Classification Décimale Universelle". It has been more and more widely used in recent years, and is continually being extended by the International Committee for Universal Classification of the Institut International de Documentation (I.I.D.) which has its seat at the Hague.

Various short treatises [2]) have been published by the I.I.D. itself or with its approval, besides the complete editions which have already appeared or are still in the process of being published [3]).

The decimal classification divides the whole domain of human knowledge into 10 main classes, each of the main classes into 10 other classes, and each of these into not more than 10 sub-classes, etc.

The main classes are the following:

0 generalities, bibliography, documentation
1 philosophy, psychology
2 religion
3 social sciences
4 philology
5 pure sciences
6 applied sciences
7 art, games, sport
8 literature
9 geography, history, biography

Main class 6, applied sciences, is divided as follows:

60 general considerations
61 medical sciences
62 engineering
63 agriculture and dairying
64 domestic economy, housekeeping
65 commerce, business organization, communications, transport organization, accountancy, industrial organization
66 chemical technology
67 various industries and manufactures
68 various trades, industries and professions
69 building construction

Class 62, engineering, is subdivided as follows:

620 general considerations
621 general mechanics and electrical engineering
622 mining
623 military and naval engineering
624 civil engineering (general)
625 road and railroad engineering
626 hydraulic engineering
627 engineering in natural waters: rivers, lakes, sea
628 sanitary engineering
629 traffic engineering

Class 621, general mechanics and electrical engineering, is divided in the following way:

621.0 general mechanical theory
621.1 production and use of steam
621.2 utilization and distribution of water power

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66 chemical technology
67 various industries and manufactures
68 various trades, industries and professions
69 building construction

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621.3 electrical engineering
621.4 internal combustion engines and heat engines other than steam
621.5 pneumatic energy
621.6 blowing and pumping machinery
621.7 factories and workshops, foundries
621.8 transmission machinery
621.9 machine tools, hand tools and workshop practice.

For highly specialized subjects numbers with many figures are obtained in this way. For clarity a decimal point is placed between every 3 figures. As an example of the complete specification of a subject the following may serve:

5 pure sciences
53 physics, mechanics
535 light, radiation, optics
535.3 reflection, refraction, absorption, emission
535.33 spectra, emission spectra
535.338 form of spectra
535.338.3 line spectra
535.338.33 structure of line spectra
535.338.332 fine structure of line spectra

In practice the sub-division does not always go as far as possible; the larger the number of publications present on a given subject, the farther the sub-division.

Here and there in the system numbers are left unused in order to be able to give a logical place to a possible new subject. At the same time the scheme also offers the possibility of further subdivision whenever this proves necessary.

It will often happen that the contents of a publication cannot be represented by a single main number. If a work treats of two independent subjects, for example chemistry, class 53, and physics, class 54, the two corresponding classification numbers are joined by the sign +:

53 + 54 textbook of physics and chemistry.

In the case of a work on "material testing in iron and steel manufacture", in which the two different subjects:

620.1 material testing and
669.1 iron, steel

are joined, the two classification numbers are combined by means of a colon:

620.1 : 669.1 material testing in iron and steel manufacture.

The following are examples of such combinations of classification numbers. To characterize an article on "a double current welding apparatus" we obtain the following index:

621.791.735 : 621.314, in which

621.791.735 indicates a welding apparatus with electric arc, and
621.314 means conversion of electric current.

An article on "The electron microscope as an aid in metallographic investigation is dismembered as follows:

537.553.72 electron microscopy
621.385.833 electron optical tubes
669.016 metallography, and indicated by
537.533.72 : 621.385.833 : 669.016.

In addition to the main classes indicated by one of the numbers from 0 to 9 and their subdivisisons already discussed, the decimal classification, also has auxiliary classes which further define the main concepts and are indicated by certain signs. There are special and general auxiliary classes.

The special auxiliary numbers are used to define more precisely the subjects of a whole group in a particular sense. They may only be used in this sense with that definite class. They are separated from the preceding numbers by a point and always begin with zero.

Special auxiliary numbers in class 621.3, electrical engineering, are for instance:

.027 kind of tension
.027.2 low tension
.029 kind of frequency
.029.5 high frequency
.029.58 short waves

With the following main class numbers from class 621.3:

621.318.4 electric coils
621.396.11 propagation of radio waves in the atmosphere and above the earth's surface, we can therefore make the following combinations:
621.318.4.029.5 high frequency coils
621.369.11.029.58 propagation of short radio waves.

An article on "protection of overhead low tension mains against overvoltage" may be given the following classification number:

621.316.1.027.1 : 621.316.9

In this number 621.316.1 indicates distribution mains and .027.2 low tension, while 621.316.9 represents protection.

An article on "the recording of rapidly occurring electrical phenomena with the aid of cathode-ray tube and camera" receives the following classification number: 61.317.087 : 621.385.832 : 771.3, in which 621.317 indicates electrical measuring technique, and the auxiliary number .087 indicates recording. The numbers 621.385.832 and 771.3 indicate respectively cathode-ray tube and camera.

The general auxiliary numbers may be used with any main concept. The most important are those for "form", "place" and "point of view".

Auxiliary numbers for form, represented by (0):
(021) handbooks
(05) periodicals
Combinations:
621.396 (021) handbook of radio technique
621.396 (05) radio periodical

Auxiliary numbers for place, represented by ( ):
(492) The Netherlands
(73) The United States
Combinations:
621.311 (492) electrical power stations in the Netherlands
654.19 (73) radio broadcasting in the United States

Auxiliary numbers indicating point of view, beginning with .00:
.001.5 scientific and technical research
.003.2 financial point of view
Combinations:
62.001.5 technical research
654.19.003.2 (492) the financial condition of the the Netherlands radio broadcasting.

Auxiliary numbers are only used when the main numbers do not indicate a concept sufficiently clearly.

In the arrangement of the single decimal classification numbers, for instance in a catalogue, one must imagine that from each number a decimal fraction is formed by setting a 0 in front of it. The various numbers are then arranged according to the increasing values of these decimal fractions.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS' GLOEILAMPENFABRIEKEN

1224: W. Elenbaas: Der Gradient der Hochdruckentladung in verschiedenen Metalldämpfen (Physica 4, 747 - 751, Aug. 1937).

The gradient in the high pressure discharge in cadmium and zinc vapour is measured in volts per cm as a function of the tube diameter; the amount of metal vapour per cm of arc and the energy consumption per cm are also measured. These are compared with earlier measurements on high pressure mercury tubes. With the vapours of the three metals named it is found possible to describe the gradient by means of the same formula, in which only the constants vary for the three vapours. The relations between these constants for the different gases correspond to expectations.

1225: J. A. M. van Liempt und W. van Wijk: Die Schnell-analyse des Edelgasgehaltes in Stickstoff (Rec. Trav. chim. Pays Bas **56**, 755 - 759, July 1937).

The method described in **1191** for the analysis of mixtures of argon and nitrogen by absorption of the latter in liquid lithium is further elaborated for mixtures rich in nitrogen, so that at present the content of rare gases in nitrogen can be determined in 3 minutes.

1226: A. Bouwers: Les radiographies instantanées. (Arch. Él. méd. **45**, 89 - 93, Mar.-Apr. 1937).

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1227*: M. J. O. Strutt: Sound distribution over existing heavy current electrical mains. (Gedenkb. Prof. Feldmann, Waltman, Delft, 1937, pp 355 - 362).

A system is developed with which four sound programs can be distributed over existing electric mains from a central transmitter to all those connected with the mains, by means of four carrier

*) An adequate number of reprints for the purpose of distribution is not available of those publications marked with an asterisk. Reprints of other publications may be obtained on application to the Natuurkundig Laboratorium, N.V. Philips' Gloeilampenfabrieken, Eindhoven (Holland), Kastanjelaan.

The general auxiliary numbers may be used with any main concept. The most important are those for "form", "place" and "point of view".

Auxiliary numbers for form, represented by (0):

(021) handbooks
(05) periodicals

Combinations:

621.396 (021) handbook of radio technique
621.396 (05) radio periodical

Auxiliary numbers for place, represented by ( ):

(492) The Netherlands
(73) The United States

Combinations:

621.311 (492) electrical power stations in the Netherlands
654.19 (73) radio broadcasting in the United States

Auxiliary numbers indicating point of view, beginning with .00:

.001.5 scientific and technical research
.003.2 financial point of view

Combinations:

62.001.5 technical research
654.19.003.2 (492) the financial condition of the Netherlands radio broadcasting.

Auxiliary numbers are only used when the main numbers do not indicate a concept sufficiently clearly.

In the arrangement of the single decimal classification numbers, for instance in a catalogue, one must imagine that from each number a decimal fraction is formed by setting a 0 in front of it. The various numbers are then arranged according to the increasing values of these decimal fractions.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS' GLOEILAMPENFABRIEKEN

1224: W. Elenbaas: Der Gradient der Hochdruckentladung in verschiedenen Metalldämpfen (Physica 4, 747 - 751, Aug. 1937).

The gradient in the high pressure discharge in cadmium and zinc vapour is measured in volts per cm as a function of the tube diameter; the amount of metal vapour per cm of arc and the energy consumption per cm are also measured. These are compared with earlier measurements on high pressure mercury tubes. With the vapours of the three metals named it is found possible to describe the gradient by means of the same formula, in which only the constants vary for the three vapours. The relations between these constants for the different gases correspond to expectations.

1225: J. A. M. van Liempt und W. van Wijk: Die Schnell-analyse des Edelgasgehaltes in Stickstoff (Rec. Trav. chim. Pays Bas **56**, 755 - 759, July 1937).

The method described in **1191** for the analysis of mixtures of argon and nitrogen by absorption of the latter in liquid lithium is further elaborated for mixtures rich in nitrogen, so that at present the content of rare gases in nitrogen can be determined in 3 minutes.

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waves of 20 000, 30 000, 40 000 and 50 000 cycles per sec. From measurements of the characteristics of low voltage mains, high voltage mains, telephone cables and special measuring arteries in this frequency range, the technical and economical possibility of this distribution system has been established. On the basis of these measurements the necessary energy for given mains can be calculated. The technical construction of the system is described briefly.

1228: J. van der Mark: Television with the aid of the iconoscope (Hand. 26e Ned. Natuur- & Geneesk. Congr. 63 - 68, 1937).

For the contents of this lecture the reader is referred to Philips techn. Rev. **1**, 16 and 325 (1936).

1229: C. Dorsman: Television with the aid the Nipkow disc and multiplier (Hand. 26e Ned. Natuur- & Geneesk. Congres, 68 - 73, 1937).

For this lecture see Philips techn. Rev. **2**, 72 (1937).

1230: W. G. Burgers: The electron microscope in metallographic investigation (Hand. 26e Ned. Natuur- & Geneesk. Congr. 124 - 127, 1937).

The material discussed in this lecture is given in detail in Philips techn. Rev. **1**, 313 and 321 (1936).

1231: Balth. van der Pol: Concerning an infinite number of coupled oscillation circuits in connection with the problem of the cable (Hand. 26e Ned. Natuur- en Geneesk. Congr. 145 - 148, 1937).

It is explained how the transition can be made from an infinite number of coupled oscillation circuits to a continuous cable; cf. the detailed article in Philips techn. Rev. **1**, 367 (1936).

1232: J. L. Snoek: Magnetic Investigation of structure by means of patterns in iron filings (Hand. 26e Ned. Natuur- & Geneesk. Congr. 157 - 159, 1937).

The patterns produced in iron filings are studied in the case of anisotropic rolled nickel iron band for pupin coils in which the elementary magnets in the plane of the band lie perpendicular to the direction of rolling. On the side surface where the magnetic field is perpendicular to the surface, figures appear in the filings which upon commutation of the field pass over into their complementary figures. On the surface of the band not only regular parallel stripes, but also peculiar branches occur, which are difficult to reconcile with the hypothesis that the iron-filings patterns occur at the edges of the elementary magnetic zones.

1233: C. J. Dippel: A very sensitive quartz spiral spring balance (Hand. 26e Ned. Natuur- & Geneesk. Congr. 167 - 169, 1937).

The apparatus described in Philips techn. Rev. **1**, 126 (1936) is discussed in this lecture. It is possible to weigh down to a microgram in a vacuum with this apparatus.

1234: N. W. H. Addink: The relative positions of emission and absorption bands in fluorescence (Hand. 26e Natuur- & Geneesk. Congr. 171 - 172, 1937).

In the case of the fluorescing dye rhodamine, the lowest frequency absorbed is lower than the highest frequency emitted, so that the absorption and emission bands overlap. The absorption is found to be independent of the state of excitation of the rhodamine. The emission band is much extended toward lower frequencies with an increase in the temperature from —75°C to +20°C, while it changes little at the higher frequency side. Such an influence of temperature cannot be explained by means of the theory proposed in 1928 by Perrin for fluorescence, but follows logically from a consideration of potential curves.

1235: A. Bouwers: New possibilities for radiotherapy (Hand. 26e Ned. Natuur- & Geneesk. Congr. 251 - 257, 1937).

In this lecture given before the medical section of the congress a survey was given of modern research on the transmutation of atomic nuclei (cf Philips techn. Rev. **2**, 97, (1937)). The possibility is discussed of the application of induced radio activity in therapeutics.

1236: W. J. Oosterkamp: The quality of X-ray films (Ned. T. Geneesk. **81**, 2805, June 1937).

This article is a briefer edition of the article reviewed in **1214**.

1237: A. Bouwers: Die Schärfe des bewegten Röntgenbildes (Fortschr. Röntgenstr. **56**, 148 - 149, July 1937).

Ordinary X-ray photographs are sharpest when the geometrical lack of sharpness, kinematic lack of sharpness and screen lack of sharpness are equal, because their product determines the amount of light available. With the rapid successive exposures used in cinematography, however this is not so, since the energy supplied by the tube increases more slowly than proportional to the width of the focal spot. In taking X-ray films therefore it is well to take the geometrical lack of sharpness smaller than the kinematic and screen lack of sharpness. The periods of rest between the exposures may be taken about three times as long as the time of exposure. When projection is used, 90 per cent of the light is lost even when very fast objectives are used so that ten times as much light is needed. The total lack of sharpness therefore becomes about twice ($\sqrt[3]{10}$) as great in the most favourable case.

1238: J. van Niekerk: Anaphylaxis and Vitamin C (J. Allergy, **8**, 446 - 449, July 1937).

The influence of vitamin C on anaphylactic sensibilisation and shock was studied in experiments with Guinea-pigs. Horse serum was used as an anaphylactogen, the sodium salt of l-ascorbic acid as vitamin C solution. It was found that an avitaminosis C, a scorbutic condition, of the experimental animals had no influence on creating an anaphylactic sensitivity or on the shock phenomena caused by a second injection in sensitised animals.

In the amounts used of 25 to 100 mg and administered in the manners described, vitamin C has no protective action against the usually fatal shock symptoms in consequence of the second injection of the anaphylactogen. Addition of vitamin C to horse serum does not hamper the anaphylactogenic action of the latter.

1239: B. D. H. Tellegen and V. Cohen Henriquez: Inverse Feedback (Wirel. Eng. **14**, 409 - 413, Aug. 1937).

After an introduction giving the general characteristics of inverse feedback (cf. Philips techn. Rev. **2**, 289, 1937), its application to amplifiers for sound reproduction is discussed. A two-valve amplifier is described which feeds a loudspeaker and is provided with inverse feedback of the voltage on the loudspeaker to the cathode of the first valve with a dependence on frequency such that the desired frequency characteristic is obtained.
Measurements of frequency characteristics and distortion show the advantages of inverse feedback.

1240: A. Bouwers und A. Kuntke: Ein Generator für drei Millionen Volt Gleichspannung (Zs. techn. Phys. **18**, 209 - 219, Aug. 1937).

A critical survey is given of the different methods used for generating the very high voltages such as are necessary in nuclear physics (cf Philips techn. Rev. **2**, 97, 1937). The cascade generator is described (cf Philips techn. Rev. **2**, 161, 1937). The pulsation of the voltage and the fall in voltage upon loading are calculated. The high voltage is measured with the help of an oil-cooled measuring resistance, through which a maximum current of 1 mA flows. The high voltage is then read from a static voltmeter which is connected with one one-thousandth of this resistance. The high voltage is regulated by means of the energizing field of the alternating current generator. Experience with 1.6 MV direct voltage with respect to earth is discussed, as well as measurements with a spark gap between spheres 1 m in diameter, up to 1.3 MV with respect to earth. It must be especially noted that after long functioning the occurrence of a flash-over across unexpectedly long distances seems to be unavoidable.

1241: D. M. Duinker: Siebkreise für Gleichrichter (Elektrotechn. u. Masch.-Bau **55**, 381 - 388, Aug. 1937).

The filter used with a rectifier is for the purpose of smoothing the pulsating direct voltage. In designing the filters account must be taken of the highest permissible currents which occur in the rectifier valves upon switching in as well as in a condition of steady output. It is moreover desirable to obtain a direct voltage which depends as little as possible on the load. From these facts the minimum values may be deduced for the product and the quotient of the self-inductance and the capacity, from which values the value of the quantities themselves may be found.

In addition the influence is discussed of asymmetries on the pulsation of the direct voltage and the considerations which hold in the determination of the number of filter units.

1242*): W. Geiss: Der Einfluss der Spannungsschwankungen auf Glühlampen und Beleuchtung. (Bull. Schweiz. Elektrotechn. Ver. **28**, 225 - 226, May 1937).

In this lecture, given before the Schweizerische Elektrotechnische Verein, the consequences are discussed of the fact that the light flux and the length of life of a filament depend to a large degree upon the voltage to which it is subjected. An incandescent electric lamp which, at 220 volts, has a light flux of 125 decalumens and a guaranteed life of 1000 hours, will, at 231 volts give 150 decalumens and burn only 500 hours, while at 209 volts it gives only 100 decalumens and lasts for 2000 hours. If it is desired, on so-called 220 volt mains, to guarantee an average of 220 volts even for consumers at some distance, the average voltage for a user near a point of distribution may have to be as high as 230 volts. For this consumer lamps intended for 220 volts will of course prove unsatisfactory. It seems obvious that the lamps should be constructed for the average voltage over the whole mains; in this case, therefore, for 225 volts. Since the practice of voltage regulation is not the same for all mains, it is desirable that the conception of what is meant by 220 volt mains should be standardized.

1243: M. J. O. Strutt: Characteristic contants of H.F. pentodes. Measurements at frequencies between 1.5 and 300 Mc/s (Wirel. Eng. **14**, 478 - 488, Sept. 1937).

The four fundamental constants of short wave valves are the input and output impedance, slope and feed-back impedance; they are measured by means of a resonance circuit and a diode voltmeter in parallel with it. The manner in which the diode voltmeter is calibrated to 300 Mc/s, and the different measuring arrangements are described. The input impedances up to 300 Mc/s and the output and feed-back impedances up to 60 Mc/s are given for the Philips valves AF 3 and AF 7, the Mullard valves SP4B and VP4B and an acorn pentode. The theoretical derivation is given of the fact that the reciprocal of the resistance and the capacitance in the case of impedances measured during action, must be functions of the square of the frequency. This is also confirmed by experiment. The causes of the decrease in these impedances at higher frequencies are briefly indicated. Finally it is found to be possible to attain a more than fivefold high-frequency amplification for waves of 5 m with the valves AF3 and VP4B.

1244: Balth. van der Pol: Discontinuous phenomena in radio communication (T. Inst. el. Eng., London, **81**, 381 - Sept. 1937).

In this lecture before the radio section of the Institution of Electrical Engincers the advantages are pointed out, in the theoretical treatment of various phenomena in radio, of making use of discontinuous functions, such for example as the unity function of Heaviside, which is zero for all negative values of the independent variable, and at zero value of the independent variable jumps to a value of unity, which it retains further for all positive values of the independent variable. Non-periodic phenomena can, as a matter of fact, only be described in this way. A detailed treatment is given of one-dimensional wave functions and two-dimensional potential functions with, instead of the usual differential equations, difference equations, which give a deeper insight into all kinds of concepts such as divergence and rotation and line integrals.

1245*): W. Uyterhoeven:Gas discharge lamps with "low pressure" of gas. (Ned. T. Natuurk. **4**, 199 - 216, Sept. 1937).

This lecture was given in the vacation course of the Genootschap voor Verlichtingskunde in 1936, and begins with an explanation of the difference between the mechanism which produces visible radiation in a discharge in sodium and in a discharge in mercury (cf. Philips techn. Rev. **1**, 2 and 70, 1936). The various forms of discharge and the stabilization of the discharge are then discussed. Finally neon lamps, low pressure mercury lamps and different types of sodium lamps are discussed.

1246: W. G. Burgers: Unmittelbare Beobachtung von Gefügeumbildungen bei hohen Temperaturen mit Hilfe des Elektronenmikroskopes (Z. Metallk. **29**, 250 - 251 Aug. 1937).

For the contents of this lecture before the second international congress of the International Association for Testing Materials (London April 1937) compare Philips techn. Rev. **1**, 313 and 321 1936).

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LAMP MANUFACTURE AND VITAMINE RESEARCH

by A. VAN WIJK. 535.61-31 : 621.327.31 : 612.392.01

In this article it is shown how in the Philips Laboratory the problem of formulating the requirements to be met with sources of ultraviolet radiation in view of their application has led to research on vitamines, and particularly to new ideas about vitamine D.

Introduction

If the human or animal diet is deficient in certain organic substances the health of the human being or animal will suffer. These generally quite complex substances are necessary only in extremely small amounts and are called *vitamines*. Vitamines have become familiar to the general public within a very short time. Although the idea of malnutrition due to their lack was introduced into the science of health only a few decades ago, almost everyone is now acquainted with the terms vitaminosis and deficiency diseases and realizes that the former are essential for preventing the latter. It will, however, probably be a surprise to many that a lamp factory may be concerned with vitamines, as the title of this article suggests. It is our intention to show how it came about naturally that the Philips Laboratory began to work on vitamines, and how this work resulted in investigations which led to deeper insight into the subject.

Ultraviolet radiation

The invisible ultraviolet radiation forms the direct connection between the lamp factory and the subject of vitamines. Since the beginning of the present century ultraviolet radiation has been playing an increasingly important part in medicine and biology. It is not only true that a great many diseases can be cured by irradiation of the patient with ultraviolet light, the significance of this radiation for health is even greater, since a certain minimum dosage of ultraviolet light may be regarded as essential for the maintenance of good health. The sun is the natural source of this light. Ultraviolet radiation is thus a necessity of life, and when this necessity is not supplied deficiency phenomena appear. Various diseases, including infectious diseases, are particularly apt to occur after a continued lack of ultraviolet light, for example toward the end of winter, when the sun gives little or no active ultraviolet radiation. Certain diseases are called "diseases of the dark", because they occur or are promoted by habits of living or circumstances in which too little light — and in this connection especially ultraviolet light — is received. Among these are rickets, scrofula and tuberculosis.

Since the natural source of the necessary ultraviolet radiation, the sun, has been found to fall short, particularly in the winter and in our climate, use has been made of artificial sources of radiation to supply the deficiency. In the first place this was done in purely therapeutic cases, but as the realization of the necessity of ultraviolet radiation for the healthy individual gained ground, such sources began to be employed for protection against disease, as a prophylactic or, one might say, as a measure of hygiene. It will be clear that this presents a problem to the lamp factory laboratory. What are the requirements which a lamp must satisfy in order to replace the sun when used as a hygienic measure? What must be the nature of the radiation emitted and how intense must it be? It must be kept in mind that ultraviolet radiation is a collective term for all the radiation with a wave length between 400 and about 185 mμ (shorter wave lengths also, although they may be excluded immediately, since they do not penetrate through the atmosphere). The ultraviolet radiation from the sun is confined to the longer wave lengths, and the shortest wave length occurring in the sun's spectrum depends upon the part of the earth's surface and the season of the year. It is, however, never below about 290 mμ. It is therefore certain that a source of radiation which must substitute the sun need give no radiation with a wave length less than 290 mμ. It is, however, another question whether it is permissible for it to do so; in other words whether there would be disadvantages if it did. Further there is the question whether it is actually necessary that radiation with a wave length in the neighbourhood of 290 mμ be present, or whether it is sufficient to have only longer waves?

Direct observation can supply answers to the above questions only imperfectly and incompletely. Since we are here concerned with the prophylactic action on persons in good health, the wide deviations existing among such persons necessitate a statistical treatment so extensive that is rarely possible to be carried out.

One general conclusion can, however, be accepted with confidence, and that is that practically all the beneficial action of every source of radiation is lost if the radiation is filtered through ordinary glass. Whereas ordinary glass transmits a large part of the radiation with a wave length greater than about 320 mμ, but absorbs shorter wave lengths to a large extent, the effectiveness of ultraviolet light thus appears to be chiefly confined to the radiation having a wave length shorter than about 320 mμ. A further reaching conclusion may be arrived at by closer study of a single definite effect of the ultraviolet light, an effect which is distinguished from most of its other effects by the fact that the mechanism has to a large extent been clarified, and therefore lends itself to the study of specific questions.

Rickets (rachitis) and ultraviolet radiation

The above-mentioned effect is the prevention and cure of rickets. This is a disease of the growing skeleton, and there fore a children's disease, the main feature of which is that the newly formed bone tissue does not calcify but remains soft.

As a result deformities may appear as bowlegs and knock knees, too narrow a pelvis or chest, curvature of the spine, etc.

On the basis of the research of Mc Collum and others, rickets was known to be a deficiency disease caused by the lack of a certain vitamine which should be present in cod liver oil and was called vitamine D. Only a few natural foods contain vitamine D in amounts of any importance. This fact explains the widespread occurrence of the disease in children, when no preventive measures are taken. Huldschinsky now observed that rachitic children could be cured by irradiation with a quartz mercury lamp ("sunlight lamp treatment") just as well as they were cured by administration of a substance containing vitamine D. The further study of this very interesting dualism in the cure of rickets was greatly simplified by the fact that experiments could be carried out on animals, in this case young rats, which can be rendered rachitic by means of an experimental diet of a definite composition. It soon appeared that animals which received this diet, but which were at the same time irradiated with ultraviolet light, developed no rickets. This was also the case, when a sufficient amount of a substance containing vitamine D was added to their diet. Vitamine D and ultraviolet light appeared therefore to be equivalent, although at first the connection between them was not evident. The explanation was found after Steenbock and Hess discovered that ultraviolet light had the same effect when, instead of the animals themselves, their food was irradiated. This irradiation gave the food the properties of a substance containing vitamine D, that is, vitamine D was formed in the food under the influence of ultraviolet light from some component which was originally inactive. In a remarkably short time further facts were established about this component which was indicated by the name *provitamine D.*

Provitamine D

In the beginning the provitamine D was thought to be identical with *cholesterol* (Hess). Cholesterol is a very common component of animal fats. When separated from the other components, and after having been irradiated with ultraviolet light, it was actually found to exhibit the action of vitamine D. At the same time Hess and Weinstock were able to show that irradiation caused a physically observable change in cholesterol: an absorption band in the ultraviolet at about 300 mμ gradually disappeared.

Various other sterols and cholesterol derivatives were investigated by Rosenheim and Webster with a view to their possible activation. They found that irradiated ergosterol is very active compared with irradiated cholesterol.

In addition they ascertained that rigorously purified cholesterol had lost its activatability, so that it was not the cholesterol itself, but an impurity which must be considered to be the provitamine. This conclusion was also reached by Pohl and Windaus, who studied the change in the absorption of cholesterol upon irradiation more quantitatively. Cholesterol purified in a certain way, and no longer capable of being activated, also failed to show the characteristic absorption bands between about 270 and 300 mμ.

On the basis of the sensitivity of provitamine D to certain chemical treatments (bromination and de-bromination, oxidation) it was considered possible that ergosterol, occurring as an impurity in natural cholesterol, was the provitamine. Ergosterol was actually found to give the characteristic absorption bands which are also shown by the cholesterol, when not specially purified and still capable of being activated. The ergosterol, however, exhibited much more intense absorption: about 0.5 mg of ergosterol gave an absorption of the same order of magnitude as 1 gram of cholesterol dissolved in an equal amount of solvent, and with the same thickness of layer. When it was then found that after irradiation the 0.5 mg ergosterol was also equivalent to 1 gram of cholesterol for the cure of rachitic rats, the conclusion seemed to be obvious. There was, however, a fault in the logic which remained unnoticed for many years. It had indeed been proved that ergosterol receives antirachitic properties by irradiation with ultraviolet light, and further, that in this case no impurity of the ergosterol must be considered responsible, a fact which we shall not discuss in this article. Ergosterol therefore may with truth be called provitamine D. From this, however, it by no means follows that the impurity of cholesterol with provitamine D properties is actually ergosterol. It was only shown that the provitamine D in cholesterol has a similar absorption spectrum to that of ergosterol, and in the same way as ergosterol is sensitive to certain chemical treatments.

Although the conclusion seemed very seductive and plausible, and although it was accepted for nearly 10 years explicitly or implicitly by science, it has been found to be incorrect. It has been found that ergosterol is not the only provitamine D, and, especially, that it is not the provitamine which occurs in cholesterol. This fact in no way detracts from the value of ergosterol as provitamine D or from that of the vitamine D obtained from it as a preventive against rickets in children or rats or pigs, nor from most of the scientific research on the photochemistry of vitamine D formation. It would lead us too far afield if we were to explain precisely how one was gradually brought to this conclusion. It will suffice here to state that the credit of being the first to formulate this fact is due to Waddell, on the basis of a series of very fine experiments on chicks, in which it was found that the effect of a definite amount of irradiated cholesterol was much greater on these animals than that of an amount of irradiated ergosterol which was equivalent to the cholesterol in the case of rachitic rats [1]).

In order to demonstrate the great similarity between the absorption spectra of the provitamine in cholesterol and of ergosterol, *table I* and *fig. 1*

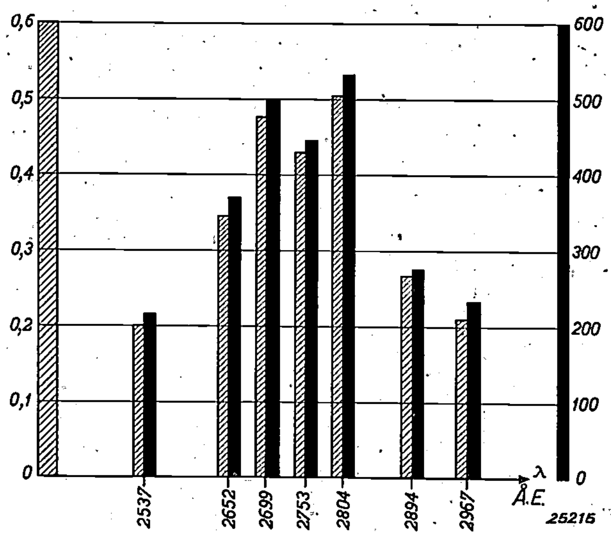

Fig. 1. Comparison of the absorption indices of provitamine in cholesterol (cross-hatched) and of ergosterol (black) for various mercury lines.

show one of our measurements on these spectra. In the table below is given the absorption index for various wave lengths of a solution of cholesterol containing the provitamine which had been repeatedly purified by recrystallization ($\alpha_{\text{chol + pr}}$).

This absorption index α is defined by $I/I_0 = e^{-\alpha c d}$, where I_0 and I are the intensities of the incident and transmitted radiation respectively with a thickness of layer d (in mm) and a concentration c (as unit of concentration 80 g/liter was used). By repeated irradiation, treatment with absorption charcoal and repeated recrystallization, cholesterol containing no provitamine is obtained (α_{chol}). The difference between the absorption indices before and after the treatment gives the absorption index of the provitamine (α_{pr}) multiplied by its unknown concentration in the cholesterol (c_{pr}). A comparison of the value obtained with the absorption index of ergosterol (α_{erg}) gives a ratio which is very nearly independent of the wave length. If

[1]) The vitamine D obtained from an impurity of cholesterol and that obtained from ergosterol are therefore different substances, both of which show antirachitic activity. If one compares quantities of the two substances which have the same antirachitic action on rats, the first substance has a much stronger action on chicks than the other. At the end of this article the experiments on chicks are discussed as well as the difference between the two kinds of vitamine D (cf. page 38).

we assume not only that the form of the provitamine spectrum is similar to that of the ergosterol spectrum, but also that the absolute magnitude of the absorption is the same, the constant ratio indicates the concentration of the provitamine in the cholesterol.

Table I. Comparison of the absorption spectra of provitamine in cholesterol and of ergosterol.

λ mμ	$a_{chol+pr}$	a_{chol}	$a_{chol+pr} - a_{chol} = a_{pr} \cdot c_{pr}$	a_{erg}	$\frac{a_{pr} \cdot c_{pr}}{a_{erg}}$
296.7	0.233	0.019	0.214	232	0.00092
289.4	0.286	0.022	0.264	278.5	0.00094
280.4	0.536	0.028	0.508	530.5	0.00095
275.3	0.559	0.028	0.431	441.5	0.00097
269.9	0.525	0.050	0.475	499	0.00095
265.2	0.438	0.089	0.349	370.5	0.00094
253.7	0.370	0.065	0.205	212	0.00096

Photochemistry of vitamine D formation; connection with the requirements for sources of ultraviolet light

After this digression on provitamine D, let us return to the point in the argument where we stated that rickets can be prevented (or cured) on the one hand by irradiation of the test animals, and on the other hand by irradiation of their food. The link between these two statements lies in the fact that the skin (of human beings as well as of rats, pigs or chicks) contains cholesterol which, because of its content of provitamine D, can be rendered antirachitic in action. The vitamine D so formed can be taken up by the body from the skin, just as it can be taken from the food. In this way is explained the apparent dualism of rachitis — a deficiency disease due to lack of vitamine D, avitaminosis or hypovitaminosis on the one hand, and on the other hand a "disease of the dark". At the same time this is a case (the only one up to now) of the action of ultraviolet light on the human (and animal) organism in which the primary process taking place in the skin is known, and permits of closer study in the laboratory. As far as measures against rickets are concerned, therefore, the question as to the most favourable composition of the light of a "hygienic" ultraviolet light source comes down to the question of the best composition for the radiation in the photochemical formation of vitamine D from provitamine D.

In our investigations of this question [2]) we found that the composition of the radiation is by no means unimportant. These investigations were carried out in the first place on the provitamine ergosterol, and were later repeated as far as necessary on other provitamines D. The results are the same with the different provitamines D, a fact which was to be expected on the grounds of the similarity of their absorption spectra in the ultraviolet region. On the one hand the absorption spectrum determines the radiation which can be absorbed and therefore which can be active, and on the other hand the nature of the atomic bonds in the molecule, and thus the possible transitions due to the absorption of radiation, are expressed in the absorption spectrum.

For the study of the transformations taking place extensive use was made of the observation of the changes occurring in the absorption spectrum. For the purpose of irradiation the provitamine was dissolved in a suitable inactive solvent with a high transparency in the ultraviolet (especially purified hexane, ether or alcohol). It was found necessary to remove all oxygen before irradiation, also that dissolved in the solvent, as otherwise oxidized and inactive reaction products were obtained (for the technique and that of the determination of the absorption spectra the reader is referred to the publications mentioned in the footnote). After having been irradiated for some time a part of the provitamine will be changed into one or more reaction products which in turn may be liable to further change.

If the absorption spectrum of such a solution is measured, it represents in general the spectrum of a mixture of substances. One of these is known, namely the provitamine. The amount of provitamine in the mixture which has not been changed can be determined with the aid of digitonine. (Digitonine gives with the provitamine an addition product insoluble in alcohol, while it does not with the reaction products). By subtracting from the absorption found for the mixture the part due to the provitamine still unattacked, the absorption of the mixture of the reaction products is obtained. For each case this can be recalculated to give that which would be found for any arbitrarily chosen concentration, that is the specific absorption of the reaction mixture. The result of measurement and calculation now depends closely upon the composition of the radiation used for bringing about the reaction. It was found possible to distinguish between two typical kinds of active radiation mixtures, namely:

a) that in which only radiation with a wave length

[2]) E. H. Reerink and A. van Wijk, Proc. Kon. Acad. Wet. 32, 845, 1929; Bioch. J. 23, 1294, 1929; Bioch. J. 25, 1001, 1931; Chemistry at the Centenary (1931) Meeting Brit. Ass.; E. H. Reerink, A. van Wijk and J. v. Niekerk, Chem. Wbl. 29, 645, 1932.

greater than about 280 mμ acts; this will henceforth be called "long wave" radiation,

b) that in which only radiation with a wave length of about 250 mμ acts, which we will call "short wave" radiation.

The mixture indicated under a) was obtained by filtering the radiation of a quartz mercury lamp (mercury vapour pressure of the order of 1 atmosphere [3])) through a solution of benzene or xylene in alcohol, which absorbs the radiation of wave lengths shorter than about 272 and 284 mμ respectively.

For the "short wave" radiation use was made of a low pressure column discharge in mercury vapour, which emits almost exclusively the two resonance lines of mercury. By filtering the radiation through an absorption cell filled with chlorine gas the active radiation was practically limited to the resonance line with a wave length of 253.7 mμ.

It was now found that with "long wave" irradiation the reaction proceeds quite simply, at least as long as the transformation is not carried too far: the specific absorption a_R of the reaction mixture in the latter case then no longer depends upon the degree of decomposition (cf. *table II*). In other words the reaction may be described as if primarily only one reaction product is formed, which is then changed in its turn only after long continued irradiation. Or a mixture of reaction products may be formed whose amounts are in a constant ratio to each other independent of the degree of decomposition (when the reaction has not proceeded too far), that is to say that it is more probable that the products are formed simultaneously than that they are formed successively one from the other. The simple process described holds approximately up to a degree of decomposition of the order of 60 per cent.

Table II. Absorption spectra of "long wave" irradiated ergosterol solutions at different degrees of decomposition.

			Degree of transformation					
			10.4 %		23.5 %		31.4 %	
λ mμ	a_{erg}	a_R, long wave	$a_{calc.}$	$a_{exp.}$	$a_{calc.}$	$a_{exp.}$	$a_{calc.}$	$a_{exp.}$
313.0	0.7	25	3	2	6	6	8	10
302.2	27	87	33	31	41	41	45	47
296.7	232	172	226	222	218	217	213	215
289.4	278	264	277	274	275	274	274	275
280.4	530	395	517	515	499	494	488	489
275.3	441	418	439	438	436	433	435	436
269.9	499	452	494	495	488	484	484	485
265.2	370	442[5]	387	377	384	387	393	389
253.7	212	371[5]	228	227	249	252	261	272
248.3	144	336	165	165	189	193	205	215

The columns a_{erg} and $a_{R\ \mathrm{longwave}}$ give respectively the absorption indices of ergosterol and of the reactionproduct obtained by longwave irradiation (both a mean of a great number of measurements). The following columns give the agreement between the absorption values obtained in a definite set of measurements and the calculated values according to $a_{calc} = (1-a)\ a_{erg} + a\ a_{R\ \mathrm{longwave}}$, where a is the degree of the decomposition.

The primary reaction product of the "long wave" irradiation could be separated from the unattacked ergosterol by carrying out the necessary operations in a vacuum (precipitation of the ergosterol with digitonine, recrystallization). The crystalline substance obtained possessed an antirachitic activity of an intensity which had up to then never been found namely 26.9×10^3 I.U.D. (International Units Vitamine D) per milligram [4]). Since it was later found possible by chemical methods (esterisation with 3.5 dinitrobenzoic acid, recrystallization and saponification) to obtain a crystalline substance from the irradiation product of ergosterol, which was called calciferol, and which had a still higher activity (about 40×10^3 I.U.D. per mg), and must be considered to be pure vitamine D from ergosterol, it follows that the crystalline reaction product obtained directly by "long wave" irradiation was not yet pure vitamine D, although its content of the latter was very high.

The reaction with "short wave" irradiation proceeds in quite a different way. The specific absorption of the reaction mixture depends closely, even when the change is still small, on the degree of transformation, the spectra found differ widely from those of the "long wave" product and the antirachitic activity is much lower. For the purpose of comparison in *table III* and *fig. 2* the spectrum is given of the reaction product obtained after 30 per cent transformation with "short wave" radiation. The activity per mg was in this experiment about 11.5×10^3 I.U.D./mg. The content of vitamine D is therefore much lower than after "long wave" irradiation. This indicates that a large part of the provitamine has heen changed into other products than vitamine D. In other words the "short wave" radiation is unfavourable for the formation of vitamine D.

As to the question of the most favourable composition of the radiation for a lamp of hygienic

[3]) For the spectrum compare Philips techn. Rev. **2**, 18, 1937, spectrum of the Philips "Biosol".

[4]) By international agreement the effect tested on rats of a definite amount of an arbitrary but well defined preparation, prepared in the laboratory of the "Medical Research Council" is accepted as a unit. The following may serve to give an idea of this unit. A dosage of about 800 I.U.D. per day is considered suitable for prevention of rickets in nursing children.

value, it follows from the foregoing that absence of "short wave" radiation which does not occur in the sun's spectrum is to be desired, since

Table III. Absorption spectrum of a "short wave" irradiated ergosterol solution at a degree of decomposition of 30%.

λ mμ	$\alpha_{exp.}$	α_R short wave 30 %
313.0	22	72
302.2	73	180
296.7	263⁵	337
289.4	333	460
280.4	554⁵	610
275.3	486	589
269.9	530	602
265.2	412	538
253.7	264	385
248.3	199	323

the prevention of rickets is one of the chief demands made on such a lamp. On the basis of the absorption spectrum of ergosterol (or of another provitamine D) it again follows that only radiation with a wave length less than about 310 mμ can be effective, since only

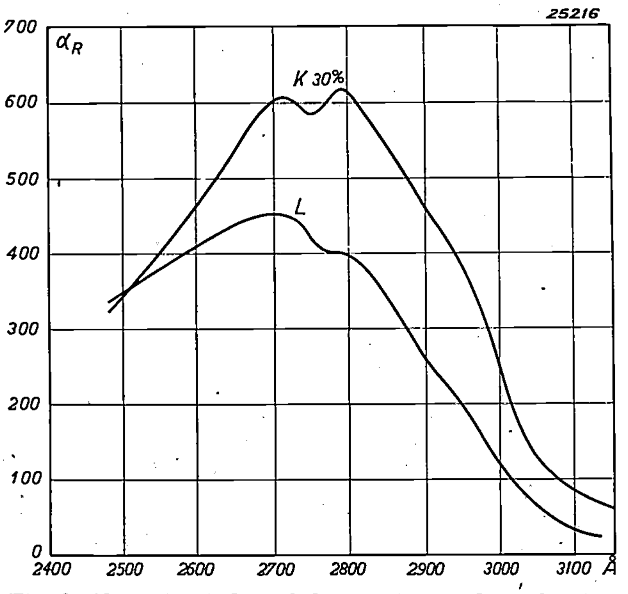

Fig. 2. Absorption index of the reaction product after irradiation of ergosterol as a function of the wave length. L irradiation with "long wave" light. The absorption index is independent of the degree of decomposition. K irradiation with "short wave" light. The curve reproduced refers to a degree of decomposition of 30%.

such radiation is absorbed. These considerations have led to the introduction of a special kind of glass, which transmits in the ultraviolet down to about 280 mμ, but which absorbs the "short wave" region, as a container for the lamps which must serve to replace the sun (cf. Philips techn. Rev. 2, 18, 1937, filter for "Biosol").

Vitamine D and chicks

Rickets is very prevalent among chicks which do not receive sufficient sunlight. This is the case with a large part of the chicks as those are often raised entirely indoors in incubators. Provision must then be made for sufficient addition of vitamine D to their food or for ultraviolet irradiation of the birds themselves. In general an unexpectedly high dosage of vitamine D obtained by irradiation of ergosterol is necessary for the prophylaxis (or therapy) of rickets in chicks. If for example one of two similar groups of chicks is given a certain amount of "natural" vitamine D which is just sufficient, and the other is given an amount of vitamine D from ergosterol which is equivalent to the first when tested on rats, it is found that the action of the latter amount is not sufficient for the chicks. About 25 times as much must be given to obtain the same result. If one now compares in the same way irradiated cholesterol and vitamine D from a natural source, amounts which have equivalent effects on rats are found to be equally effective in the case of chicks (Waddell). We have already given the explanation of this: the provitamine D in cholesterol is not ergosterol. In connection with the great importance of the prevention of rickets in chicks, and in order to remove all doubt about the correctness of the above conclusion, it seemed essential to us to attempt to actually separate the provitamine from cholesterol and to study the properties of the pure substance. All the cholesterols described in the literature contain only extremely slight traces of the provitamine (deduced from the absorption spectrum and the activatability) of the order of 0.1 per cent. Since it did not seem promising to begin with such poor sources of provitamine, we first examined cholesterols from the most varied sources spectro-photometrically for their provitamine content. We finally found a raw material, Chinese duck egg yolk, which contained a cholesterol which was sufficiently rich in provitamine to promise success in its separation. This cholesterol contained about 4 per cent provitamine [5]). As a matter of course it had still to be ascertained by comparative chick and rat experiments whether this very unusual cholesterol after irradiation still had the same favourable properties as the more ordinary kind. This was found to be the case.

A method was worked out for the separation in which use was made of fractional adsorption and desorption on aluminium oxide. The solution followed

[5]) Boer, Reerink, van Wijk, van Niekerk, Proc. Kon. Acad. Wet. 39, 622, 1936.

by large quantities of the solvent was filtered through a thick layer of aluminium oxide. The outstreaming liquid was continuously examined by means of absorption measurements, and on the basis of these measurements fractions rich in provitamine and fractions poor in provitamine were separated. Before the cholesterol was subjected to this treatment it was first esterised, with acetic acid for instance. After repeated application of this process to the fraction rich in provitamine a very high concentration was finally obtained. By recristallisation until the properties were constant (absorption spectrum, optical rotation, melting point) the pure acetate of the provitamine was obtained. After saponification and recrystallisation the provitamine itself was obtained and was actually found to differ from ergosterol. It could be identified as 7-dehydrocholesterol. *Table IV* shows the great similarity in the absorption spectra of ergosterol and the provitamine from cholesterol. The provitamine separated from cholesterol as well as the synthetically prepared 7-dehydrocholesterol, exhibit after irradiation the expected high activity with respect to rickets in chicks, with which fact the chain of evidence seems to us to be complete.

Table IV. Comparison of the absorption spectra of ergosterol and provitamine from cholesterol.

λ mμ	a_{erg}	$a_{provit.}$	$a_{7\text{-dehydrochol.}}$ (synth.)
313.0	0.7	2	0.7
302.2	27	27	26
302.2	27	224	218
289.4	278⁰	278	268
280.4	530⁰	521	504⁰
275.3	441⁰	438	424⁰
269.9	499	495	473
265.2	370⁰	367	354⁰
253.7	212	204	202
248.3	144⁰	140⁰	139

For practical uses [6]) finally other animal sterols, that from the mussel for example, were found to be much more convenient because of their high content of provitamine, which can be activated to "chicken active" vitamine D.

[6]) The above-described vitamine D preparations for use by human beings as well as for animals are marketed by the N.V. Philips-van Houten company under the general name of "Dohyfral".
The preparation for animals consists of a solution of the "chicken active" vitamine D in oil. The preparations for human consumption are in the form of chocolate tablets which are supplied with "ergosterol vitamine D" as well as with the "chicken active" vitamine D ("Neo-Dohyfral").

PROPERTIES AND APPLICATIONS OF ENAMELLED WIRE.

by J. HOEKSTRA.

621.315.617.4

Enamel lacquer is an excellent insulation material for copper wires because of its high breakdown potential, its high electrical resistance, its slight absorption of water, its unusual thinness and its good mechanical and chemical properties. In this article details are given about these properties of enamel lacquered wire.

Introduction

In electrical engineering the term "enamelled wire" (more accurately "enamel lacquered wire") means metal wire which is insulated by a thin layer of hard baked lacquer. The name often leads to misunderstanding by laymen, who usually think of the glasslike enamel of pots and pans.

The first insulated conductors were made by Faraday when he carried out his pioneer experiments on electromagnetism. According to his diary of 1831 he used "twine and calico". Textiles are thus the oldest form of insulation for wires, and they are still employed on a large scale. Later Werner Siemens used gutta-percha und rubber, which have undoubtedly much higher insulation resistance, but which must be applied in much thicker layers. It is for this reason that insulation materials to be applied in the plastic state, such as rubber, gutta percha and various similar synthetic products of recent years, have no great practical significance for coils and windings. Cotton insulation is also relatively thick, so that the appreciably more expensive silk was often used. The idea, which originated in America around 1900, of applying a thin layer of a lacquer or varnish directly to the copper also was attractive from the point of view of expense.

The idea also was worked out in Germany about 1905. The first task was that of finding a lacquer which was resistant to the chemical influences acting during impregnation and use of the wire, and which in addition was elastic and strong enough for the purpose. Difficulties were encountered at the beginning in satisfying these chemical requirements, since black enamelled wire in particular is made with lacquers having an asphalt base which is not resistant to oil.

In 1906 it was stated [1]) that in Germany copper wire was being made with a coating of cellulose acetate and also of an enamel lacquer. This wire was at first intended for the coils of measuring instruments, etc. but was later available for small machines and transformers.

The early cellulose acetate coatings have almost entirely gone out of use. The lacquers with an asphalt base (for black enamelled wire) have become much less important: all black enamelled wire is not however made with an asphalt base lacquer. Modern enamelled wire is made with the help of oil lacquers. These lacquers consist of mixtures of resins and drying oils (for example wood oil and linseed oil, usually a large proportion of the former in order to produce quick drying, chemically resistant lacquers). Kienle and Adams [2]) give as an example the following composition: 60.5% wood oil, 18.3% raw linseed oil and 21.2% resin neutralized with lime; these components are heated together $1^1/_4$ hours at 270°C, then cooled to 160 ° C and mixed with 10 % benzine and 90% solvent naphtha. It is therefore a very "oily" lacquer, and the oils are already partially polymerized by heating. Of course every manufacturer of enamelled wire uses his own lacquer, the composition of which is usually not published. Artificial resins are also often used in these lacquers.

In the manufacture of enamelled wire a small amount of such a lacquer is applied to the surface of a bare copper wire. This may be done by drawing the wire vertically out of a lacquer container. The wire is then passed through an oven in which the lacquer is "baked". In this process, the polymerisation of the drying oils, which has already begun during the heating of the lacquer, is continued by the oxygen of the air. A strong, cohesive, insoluble layer is the result.

The resins provide for greater hardness of the coating, the (highly polymerized) oil content is responsible for the great elasticity. There is as yet no absolute certainty about structure of the polymerized oils. The properties of the polymer indicate a rubberlike structure, that is a network of quite large molecules are joined to each other at relatively few points. In oil lacquers this structure gives a practically insoluble whole, a fact which indicates very strong, probably chemical bonds at the few points of contact of the large molecules. The pres-

[1]) Electrotechn. Z. 27, 16, 1906.

[2]) R. H. Kienle and L. V. Adams, Ind. Eng. Chem. 21, 1279, 1929.

ence of the "rubber structure", however, includes the posibility of the swelling of the polymer in liquids whose molecules have the tendency to penetrate between those of the polymer. It may in general be said that great elasticity is very often accompanied by a tendency to swell in suitable liquids. This tendency is very slight with enamel lacquer, but is not entirely absent. Particularly in the case of an insufficiently baked lacquer coating is it noticeable (for instance in the presence of melted beeswax and also benzine at room temperature), it is at a minimum with a normally baked lacquer coating.

The application of fluid lacquer and the subsequent baking are repeated a number of times (sometimes 6 or even as many as 30 times), until the desired thickness of insulation has been reached. This is done by leading the wire as it comes out of the oven back to the lacquer container and so on.

The application of enamelled wire has become progressively more important in the last few decades. Not only thin and very thin wire (up to 25 microns core diameter; 40 micron wire is used to a large extent at present for high tension transformers) but also the thick kinds (for instance 3 mm core diameter) are at present employed everywhere for transformers, motors, coils, low current cables and connections (telephone installations). The product has at present almost a monopoly in the radio industry (transformers, coils, so-called stranded or "litz" wire); in other branches (motors, power transformers) it is steadily gaining ground. The reasons for this increasing application are the following:

a) *The enamel coating has a very high breakdown potential.*
b) *It has a very high electrical resistance and the dielectric losses are not high.*
c) *The enamel is an unusually watertight material.*
d) *The insulation layer is unusually thin.*
e) *The mechanical properties of the coating satisfy rather high requirements.*
f) *The enamel coating is very resistant chemically.*

We shall examine these characteristics one at a time in order to build up as complete a picture as possible of the final product.

a) Breakdown potential

The breakdown potential of the enamel layer itself may be said to be very high: namely of the order of 10^6 volts/cm for layers from 5 to 20 microns thick.

The breakdown potential of the insulated wire exhibits quite wide variation; the reason for this is that the condition of the metal surface exerts a tremendous influence on breakdown. This phenomenon, which is theoretically perhaps not entirely unexpected with such thin insulating layers, results in the fact that the condition of the bare wire is actually at least as important as that of the insulating layer.

The high breakdown potential of the material is not completely manifested in the finished product because of this influence of the metal surface. Nevertheless the breakdown potentials attained, compared with the thickness of the insulation, are still very high, as may be seen from the table below, in which several of the standard specifications of various countries are noted.

Breakdown potentials (in volts eff), minimum requirements according to the standard specifications of different countries for enamelled wire of various diameters and with an alternating voltage of 50 c/s.

Diameter of wire	Standard thickness of of enamel coating (approx)	Standard Specifications			
		DIN Germany	C 31 France	C 8.7 U.S.A.	BSS 156-Engl. (1936)
mm	microns				
0.1	6	350	150	200	200
0.5	16	500	400	700	1000
0.1	22	750	550	800	1200

The breakdown potential of the various standard specifications are not entirely comparable because the methods of determination differ rather widely. The American and English specifications determine the breakdown of two enamelled wires against each other, the others of one wire against polished metal. Division by two of this "double breakdown" in order to obtain a figure comparable with the other requirement is not permissible, because of the fact that the breakdown found as an average with enamelled wires with different thicknesses of insulation is far from proportional to the thickness of the insulation. This latter fact is again explained by the above-mentioned influence of the metal surface.

Wire with twice the standard thickness of coating exhibits about three times the normal breakdown potential [3]).

Fig. 1 which reproduces the result of a large number of determinations of breakdown (DIN method) on two different lots of enamelled wire of the same kind, gives an idea of the relatively wide

[3]) By the term breakdown value is meant the voltage which is withstood by 80% of the samples tested, in agreement with the German standard specifications.

variation which occurs as a rule in the determination of this property of enamelled wire. The small difference in the average thickness of the enamel layer found in samples which broke down

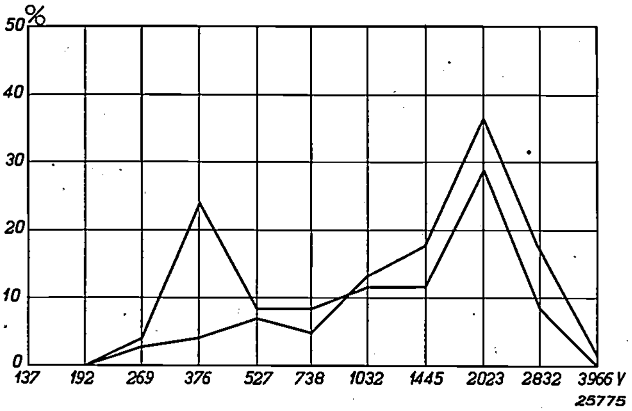

Fig. 1. Statistics of the breakdown potentials measured on two spools of different manufacture (about 60 measurements per spool). The wire was enamelled wire of 0.4 mm core diameter. The thickness of a single layer of lacquer was about 13 μ.

The method was that of the DIN: the wire was wound a bare polished cylinder of 30 mm diameter, and one end was weighted with 50 grams. A sinusoidal alternating voltage of 50 c/s was then applied between the copper wire and the cylinder, and gradually increased from 0 to the breakdown potential. E_{eff} was measured. The breakdown values found were then divided into groups in such a way, that each group begins with a voltage 1.4 times as great as the initial voltage of the previous group. The percentage of the total number of measurements in each group is plotted.

at widely different potentials, illustrates the fact stated above, that the thickness of the layer of lacquer is by no means the only factor which determines breakdown.

A comparison of the breakdown value of enamelled wire with that of ordinary kinds of wire insulated with textiles is of little value, considering the fact that the textile insulation of these wires, when tested in the unimpregnated condition, has a breakdown potential which is lower than that of the layer of air situated between the textile fibres. The spark travels along the fibres; a breakdown of the insulation material does not take place. This is also shown by the fact that the same breakdown potential may be measured repeatedly at a single point. For example, a wire of $^1/_2$ mm core diameter double covered with cotton (thickness of insulation 120 μ) has a breakdown potential of about 350 volts against a polished cylinder. If a textile wound coil is impregnated, the breakdown potential is somewhat higher, but is entirely dependent on the quality of the impregnation.

b) Insulation resistance and dielectric losses

The enamel layer itself has a very high resistance. With 1000 volts direct voltage on a meter of enamelled wire of $^1/_2$ mm diameter, immersed in mercury, a resistance of 10^{12} ohms is found if the enamel layer has no faults. With standard enamel wire, however, several faults per 10 m are permissible. Such faults are not usually "bare places" while the English expression "pinholes" is also incorrect, but they are spots with an unusually low breakdown potential. If there is such a spot on the wire immersed in mercury, breakdown occurs far below 1000 volts. Another kind of fault which may influence the insulation resistance under certain circumstances is demonstrated by immersion of the wire in water: in this case several otherwise entirely invisible faults are found to exist where the insulation resistance is lower. This holds however only for standard enamelled wire; with extra thick (so-called double-enamelled wire) these faults are practically nonexistent. This latter fact, even more than the higher breakdown potential, explains the reason for the increasing demand for extra insulated wire. The "faults" in the insulation described here are of absolutely no importance in the employment of the wire for ordinary coils and transformers because of the fact that with the small number of permissible faults it practically never occurs that two of them lie next to each other. There is a possibility that the faults might be of importance in special applications such as for unimpregnated coils to be used in a moist environment or very low power connections (telephony). In such a case it is better to make sure by using wire with an extra thick enamel layer.

The dielectric loss angle δ, measured at radio frequency, is small ($\tan \delta = 0.015$), and keeps

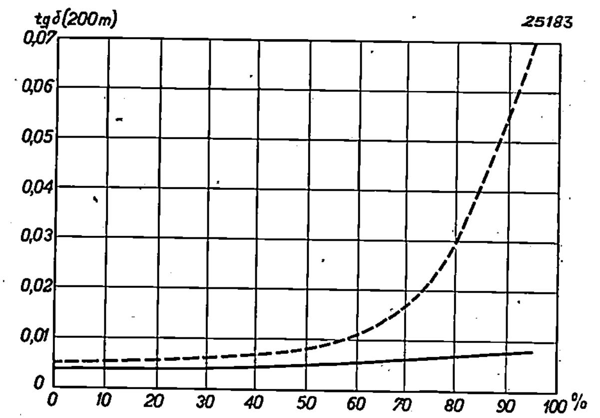

Fig. 2. Dielectric losses ($\tan \delta$) measured at 200 m wave length between two insulated wires with 0.5 mm core diameter, which were twisted together with a pitch of 2.6 cm, as a function of the relative humidity of the surrounding atmosphere. The wires were kept in this atmosphere until the losses had become constant and they were not removed from it for taking the measurements.

———————— enamelled wires

---------------- wires wound twice with silk.

the same value after long continued storage of the wire in a humidity chamber or moist room. If it is measured while the wire is in an atmosphere of definite relative humidity, the loss angle of enamelled wire is found in practice to increase slightly with the moisture content of the air. *Fig. 2* gives a clear picture of the very favourable behaviour of this material compared with silk. Double silk covered wires and enamelled wires, both with $^1/_2$ mm cores, were twisted together for this purpose with a pitch of 2.6 cm.

The insulation resistance and the dielectric losses are of particular importance when the wire is to be used for cables and for the wiring of apparatus. The application is of great and increasing importance, particularly in telephony. For wiring, enamelled tinned wire is usually used, since tinned wire is more easily soldered.

The dielectric constant is about 3.2, a relatively low value. This is an advantage for the purpose of telephony.

c) Resistance to water

Enamel laquer is an electrical insulation material with an extremely small attraction for water. Because of this characteristic its electrical characteristics remain practically unchanged in very moist surroundings. The insulation value given above of more than 10^{12} ohms may also be measured in water after 48 hours immersion. For wiring in moist surroundings or in a tropical climate enamelled wire is therefore the most suitable. Coils and motors wound with enamelled wire are found to suffer much less from moisture than when wire insulated with textiles and the like is used.

It might be thought that the same resistance to water and moisture could be obtained by soaking paper or textile insulation in impregnating lacquer, that is, also in an oil lacquer, and then baking the whole. This impregnation gives reasonably satisfactory results, but the watertightness of enamel lacquer is not even approached. As an example of this we give the following results of measurements on bifilarly wound coils of 300 double turns with 0.35 mm diameter core, wound on a central core of 3 × 3 cm.

Insulation	Resistance at	
	45% rel. humidity	60% rel. humidity
Enamel lacquer	2×10^{10} ohms	2×10^{10} ohms
Cotton, well impregnated	10^7 ohms	5×10^5 ohms

In the first place the smaller resistance to moisture of the impregnated insulation may be ascribed to the textile or paper fibres which pass straight through the layer of impregnating materials and take up considerable water. In the second place to the fact that baking in this case is carried out at about 100 ° C compared with 300-400 ° C in the case of enamelled wire. In the third place the absorption of water depends also upon the kind of oil in the impregnating lacquer. It is much less for example with wood oil than with linseed oil.

As an example of the second and third points, the table below gives the absorption of water by films of linseed oil and wood oil after 21 days immersion in water [4]) compared with that of a film of enamel lacquer kept for 21 days in an atmosphere saturated with water vapour.

Absorption of water by			
Linseed oil		Wood oil	Enamel lacquer
Dried at room temp.	Dried at 70°C	Dried at room temp.	Baked at 300-400°C
150%	30%	$\leqq$ 45%	$1^1/_2$%

In judging these figures it must be kept in mind that the thin films here investigated generally have a higher absorption of water than thick or massive pieces of the same material. Hartshorn, Megson and Rushton [5]) thus found with a film of artificial resin 0.165 mm thick an absorption of water of 3 per cent, while a film 0.675 mm under the same conditions only absorbed 1.94 per cent.

In spite of the slight permeability to water of the enamel layer it is often of importance to impregnate coils which are intended for use in moist surroundings. The purpose of this operation is then, in addition to obtaining a mechanically strong unit, to prevent the penetration of macroscopic amounts of moisture into the coil.

d) Thickness of insulation, "space" factor.

The unusually thin insulation of enamelled wire offers great advantages to the electrical engineer. The "space factor", that is the percentage of the cross section of a wound coil occupied by copper, is at present an important quantity for the constructor, now that every effort is being made to

[4]) W. Rinse and W. H. G. Wiebols, Verfkroniek, 10, 231, 250, 1937. Ind. Eng. Chem. 29, 1149, 1937.

[5]) L. Hartshorn, N. J. L. Megson and E. Rushton, Chem. Ind. 56, T. 266, 1937.

construct electrical apparatus as small as possible.

The space factor is of course dependent on the method of winding, and with completely regular mutual contact of the wires reaches a maximum, which depends then only on the thickness of the insulation. The usual (and in most countries standardized) thicknesses of the commonest wire insulations are given below for wires with core diameters of 1 and 0.3 mm respectively.

Fig. 4 shows a few pieces of standard enamelled wire which have been laid across each other on an anvil and then struck with a hammer. The result gives the impression that the lacquer coating is considerably stronger than the copper beneath it. If, however, the same lacquer films are made on thin copper foil, and the copper is then dissolved in such a way that the enamel lacquer film is not damaged, the stretch of these films by themselves is found to be considerably less than that of enamel layers on copper wire. Several stress-strain diagrams determined with a special dynamometer on such membranes are reproduced in *fig. 5*. The membranes which have a stretch of 19 per cent are decidedly soft ("un-cured"), those with 3.5 per cent stretch correspond to the layers which, on copper have a stretch of 50 per cent.

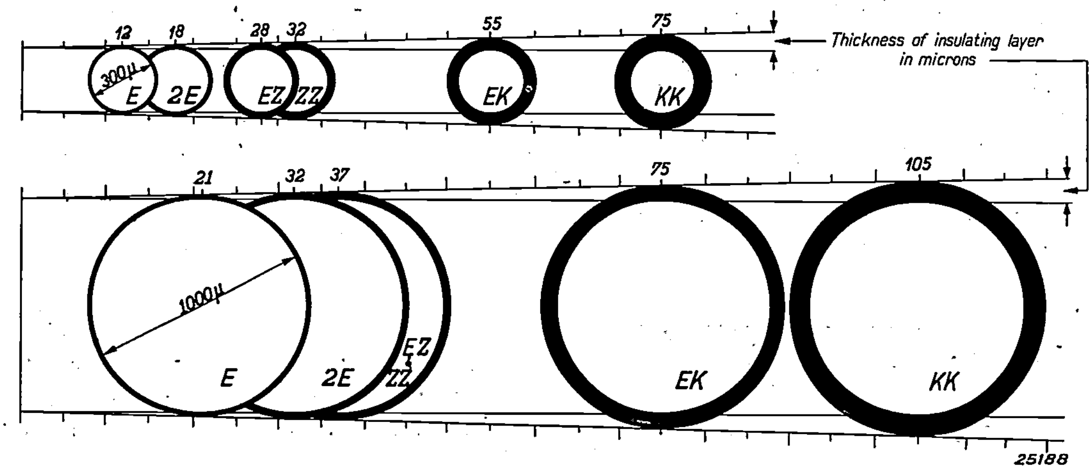

Fig. 3. Thickness of the insulation of ordinary wires for winding.
E = enamelled.
2E = thick enamel layer (twice the normal breakdown potential).
EZ = enamelled and then wound once with silk.
ZZ = bare copper wound twice with silk.
EK = enamelled, wound once with cotton.
KK = bare copper wire, wound twice with cotton.
((In upper right figure) thickness of a single layer of insulation in microns).

e) Mechanical properties of the enamel layer

The mechanical properties of lacquer baked at a high temperature form an interesting subject by themselves. If a piece of enamelled wire is stretched until it breaks, that is, often more than 30 per cent, it is found, except sometimes with very thick layers, that the layer of lacquer is quite undamaged, and that even after this test of strength it can scarcely be removed with the finger-nail.

Most of the wires also easily stand being wound about a mandrel equal in diameter to the wire. In this test the enamel layer is stretched 50 per cent on the outer side. It must, however, be noted that an enamelled wire, which has been wound around a mandrel of such a diameter that the lacquer coating only just remains intact, runs the chance of developing cracks in the coating when the wire is heated in the shape of a spiral (as is done in impregnation). The "critical" mandrel has a diameter 1 to 3 times that of the wire, depending on the material of the wire and the thickness of core and insulation.

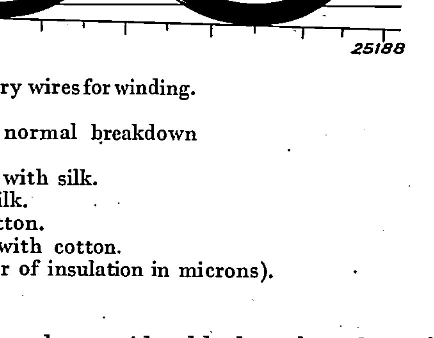

Fig. 4. Several pieces of standard enamelled wire laid across each other on an anvil and then struck with a hammer. The enamel layer shows no injury.

There is therefore no possible doubt that the bond between the lacquer and the copper beneath it is very strong, and that this bond makes possible the great stretch of the layer of lacquer. This also

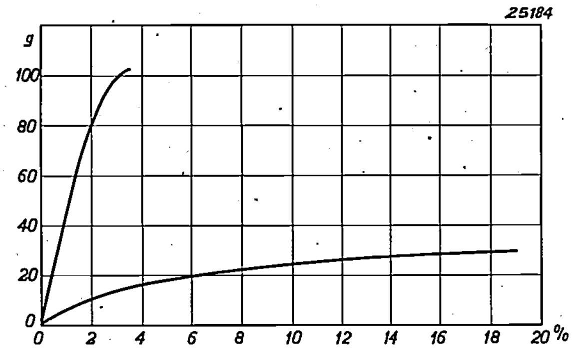

Fig. 5. Stress-strain diagrams of membranes of baked wire enamel. The membranes were baked on copper foil in four steps. each time a given thickness of layer of fluid lacquer was applied and baked a given number of seconds at the same temperature. The copper was dissolved electrolytically. The strips of lacquer film were 5 mm wide and 20 to 30 μ thick. Pieces 5 cm long were stretched about 2.5 mm/min. Upper curve: hard baked, lower curve: soft baked.

explains why thicker kinds of wire, which have thicker lacquer coatings will withstand less mechanical deformation than thin wires; the great elasticity in the thicker layers can no longer be transferred to the metal *via* the bond to the same degree as is the case with thinner layers of lacquer on thinner wires. Wire 2.5 mm thick, for example, with a lacquer coating of 30 microns will not withstand being wound around a mandrel of its own thickness, although the lacquer is in general softer (less well baked) than with wire of 0.2 mm.

The above-mentioned enamelled tinned wire has less favourable mechanical properties with respect to its insulation (mandrel test) than ordinary enamelled wire. The adhesion of the enamel lacquer to tin is much less strong than to bare copper.

f) Chemical properties of the enamel layer

The chemical qualities of the insulating layer, which are important to the electrical engineer, are chiefly stability against aging, sensitivity to temperature, resistance even at high temperature to impregnating substances (paraffin, waxes, asphalts, oil lacquers and other insulating lacquers), to moisture and oils (chiefly transformer and lubricating oil). The requirements of resistance to concentrated alkali and acid solutions, which is still often made, is actually of little importance. The resistance of the enamel to alkalies and especially to acids is high, and can be increased quite easily, although at the expense of the mechanical properties.

The stability of enamelled wire is found to be very high providing the enamel layer has been correctly baked through. In the case of clearly under-baked wire it has sometimes been found after years that the enamel coating had lost some of its elasticity. In the case of standard enamelled wire we have not been able to discover any changes in the course of years.

The sensitivity to high temperatures is slight when the heating is not of long duration. This is obvious considering the high temperature to which the product is exposed during manufacture. No anxiety need be felt for injury to the enamel layer as a consequence of the drying and impregnating of coils at temperatures up to 150 ° C. On the contrary, the heating of enamelled wire from 6 to 24 hours in air at temperatures up to 120 ° C causes a slight further hardening of the enamel layer, which is to be desired for many purposes, and is in fact deliberately brought about by many manufacturers of electrotechnical products.

In the long run the enamel layer has a resistance to high temperature which is as good as, or better than, that of textiles.

The resistance to impregnating materials and lacquers is extremely satisfactory in the case of good modern enamelled wire. Less favourable experiences, which will occasionally be met with in the literature of the last few years, may be ascribed to the fact that the most suitable raw materials were not used for the enamel lacquer.

The so-called finishing lacquer, insulating lacquers which are used cold and only applied to the outside of coils, usually soften the enamel layer temporarily. This is no disadvantage, but serves to bring about a strong adhesion between the dried layer of insulating lacquer. The constructor must of course take this temporary softening into account.

Of the impregnating materials which are applied without solvent (and therefore in a molten state), especially those which contain higher esters have some action on the enamel layer at a high temperature.

Good enamel, however, may be allowed to remain in molten beeswax for 15 minutes or longer at 150 ° C, after which, upon cooling, no depreciation of the enamel can be discovered.

The conducting core

Now that we have studied several characteristics of the layer of enamel lacquer, we shall consider briefly the conducting core of the enamelled wire.

Very much the largest proportion of enamelled wire is electrolytic copper wire. This copper, which has to be very pure to give the highest conductivity required for electrical purposes, has an annealing temperature below 200 ° C. The annealing temperature of a cold worked metal is closely dependent on the degree of deformation; the cold deformation is very great in wire-drawing, considering that this process is at present always carried out as a continuous operation on drawing machines with 10 or more dies, each of which gives the wire an elongation of from 20 to 30 per cent.

The low annealing temperature, which results in enamelled copper being always soft, is due to this cause.

The tensile strength and elongation attained are approximately as follows:

Diameter of wire	Tensile strength	%
mm	kg/mm²	Stretch
1	23—27	30—40
0.05	27—30	20—26

This annealing of the copper during enamelling is an advantage, in the first place since hard (springy) wire cannot be used for winding coils, and moreover the conductivity of annealed copper is 3 to 4 per cent higher than of copper in the hard state.

With very thin wire, however, the tensile strength of annealed copper is very low for its use in coils. In this case other metals and alloys may be used, although the choice is very much limited by the requirement that the conductivity of the metal chosen may not be very much less than that of copper. Three metals which satisfy this requirement are copper-cadmium (which is also used for overhead high tension lines), silver and copper-silver. The high price of the raw materials silver and copper silver is not all important in this case because of the fairly high cost of working this extremely fine wire. Copper-silver is very strong, but has among other disadvantages a tendency to be corroded by moisture. Copper-silver and silver have little or no stretch in the hard state. They have not reached the annealing point at enamelling temperature. If the material is annealed at a high temperature before enamelling, the advantage of great strength disappears. The absence of stretch is a disadvantage in the use of copper-silver wire on winding machines, since the wire breaks at the slightest irregularity in the winding.

The table below gives a comparison of the properties of thin wire with various core materials.

Core	Diameter of the core	Force at break	Stretch	Specific resist.
	microns	gram	%	ohms mm²/m
Copper	40	36	15—23	0.0175
Copper-cadmium 0.6%	40	56	5—16	0.0183
Silver	40	70	2—3	0.019
Copper-silver	40	138	1—2	0.022

A special use of thin enamelled copper wire is in the manufacture of insulated stranded wires for high frequency purposes.

These wires, which are divided into a large number of thin units in parallel and insulated from each other, are used especially where solid wires would give too much eddy current loss. Because of the fact that very thin wires, sometimes less than 100 μ in diameter, are used, it is clear that a good space factor can only be attained with the use of enamel as insulator.

Various requirements are made in the braiding of these wire, because for example care must be taken that the same wires do not keep to the outside of the strand over the whole length or a large part of the length. The method of braiding is otherwise noticely dependent on the purpose for which the wire is intended. Stranded wires are manufactured in great variety, from very thin for winding small coils, to very heavy interlaced braids with many hundreds of wires for carrying large currents, for instance in transmitter installations (*fig. 6*).

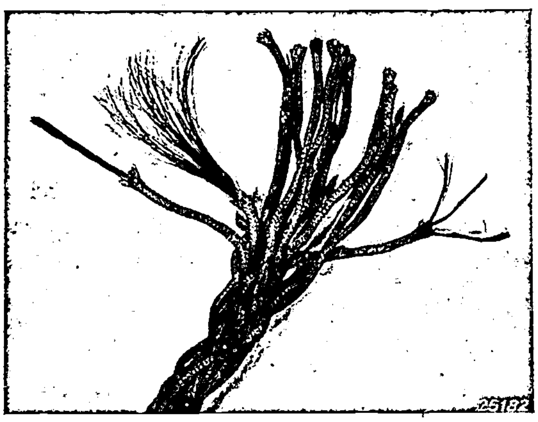

Fig. 6. Heavy stranded wire consisting of 12 bundles braided together. Each bundle is covered with silk and consists of 90 enamelled wires of 0.07 mm core diameter.

In this brief review of enamelled wire in its various forms, the practical applications of the material have not been exhausted. We have confined ourselves to the cases where the enamel forms the only insulation of the wire. Enamelled wire is also employed in combination with the spinning or braiding of paper, cotton, silk or artificial silk about the wire, impregnated or otherwise, or subsequently covered with a layer of lacquer (usually of cellulose derivatives). The discussion of these kinds of wire, however, is outside the scope of this article.

THE COLOUR REPRODUCTION OF INCANDESCENT LAMPS AND "PHILIPHANE GLASS"

by P. J. BOUMA.

535.62 : 621.326.78 : 666.24

"Philiphane" or "Neophane" glass is a glass which has an absorption band in the yellow region of the spectrum. By using this glass as an envelope for the bulb of an electric lamp, most of the colours, and particularly the blue, are reproduced with greater saturation. The loss of light flux is only slight.

Introduction

As has already been explained in a previous article [1]) the estimation of the quality of colour reproduction of a source of light depends very much upon the purpose for which it is desired to use that source of light.

In some cases it is a question of obtaining light which shall resemble daylight as closely as possible. This ideal may be approached by providing the electric lamp with a blue bulb which makes the spectral distribution of the light more nearly like daylight. Good results may also be obtained with gas discharge tubes in which the line spectrum of a gas is complemented by the continuous spectrum of fluorescent substances.

In many cases however there is no desire at all to imitate daylight, and the requirement is made that the surroundings shall have a pleasant, cosy appearance, that persons shall have a healthy appearance, that certain articles appear fresh and tasteful, etc. We shall concern ourselves here with the question of whether electric light can also be improved in this respect by the introduction of a coloured envelope.

What absorption curves must be considered?

It is clear that large portions of the spectrum must not be weakened to any extent by the bulb. This would cause too great a decrease in the efficiency. Only an improvement of the colour which can be attained with a relatively slight loss of light is to be considered.

The only possibility thus consists of the absorption of one or more relatively small regions of the spectrum. What colours may be considered in this connection?

In general, absorption of a given colour is accompanied by the following two objections:

1) An object which reflects almost exclusively this colour appears too dark.
2) Objects which exhibit the colour under consideration in a less saturated form appear still less saturated.

The first objection holds particulary for the colours at the extremities of the spectrum, thus for red and blue. Very saturated red, for example, can only occur when a material reflects practically exclusively red and orange. The same is true of blue.

For yellow, however, the situation is different. Highly saturated yellow occurs in nature as a rule, not because only a narrow region of the spectrum is reflected, but because red and green as well as yellow are fairly well reflected, and only blue and violet are absorbed to a large extent. Such a phenomenon appears in the case of the yellow "Selectiva" light used for automobile and bicycle lamps. In this case a very saturated yellow colour is obtained by the absorption of the blue and violet only. Strong absorption of the yellow therefore will cause only a relatively slight decrease in brightness of saturated yellows.

The second objection also holds particularly at the extremities of the spectrum: the blue, which is reproduced in electric light in a much less saturated form than in daylight, may certainly not be made still duller. The saturation of the red

1) Philips techn. Rev. 2, 1, 1937.

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[1]) Philips techn. Rev. 2, 1, 1937.

may also not be decreased too much, since otherwise the reproduction of skin colour would be made worse.

For the reasons mentioned above the second objection is also of much less importance in the case of yellow. Moreover the sensitivity of the eye for the observation of differences in saturation is lowest in the yellow [2]). It is therefore obvious that the influence exerted by absorption of part of the yellow on the reproduction of colour deserves further study.

The influence of "Philiphane" ("Neophane") glass

Absorption of part of the yellow can be attained by surrounding the lamp with a bulb of the so-called "Philiphane" ("Neophane") glass. The influence of this glass was studied by comparing an incandescent lamp with an ordinary opal glass bulb to one suspended in a bulb of "Philiphane" glass, according to the method described elsewhere in this periodical[1]). The main feature of this method is that a large number of differently coloured cards are lighted partially with one and partially with the other source of light. *Fig. 1* gives the result of the comparison; the arrows indicate the direction in which the "Philiphane" glass causes the colours to be displaced.

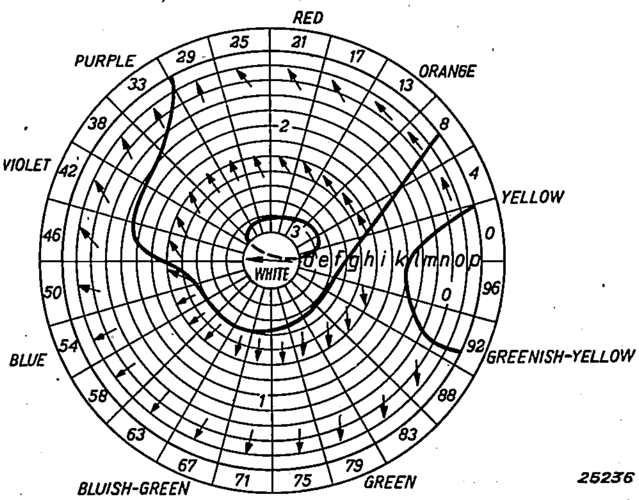


Fig. 1. Comparison of the colour reproduction by an incandescent lamp with and without a bulb of "Philiphane" glass. The arrows indicate the direction of colour change when the lamp is enclosed in a "Philiphane" bulb. The rectangles represent the colour cards of Ostwald's colour atlas. The saturated colours lie toward the outside, the least saturated in the middle.

3: region of poor correspondence
2: region of moderate correspondence
1: region of good correspondence
0: region of perfect correspondence.

For the sake of comparison *fig. 2* shows how the incandescent lamps shift the colours as compared with daylight. In the first place it will be noted that the differences in fig. 1 are in general relatively small.

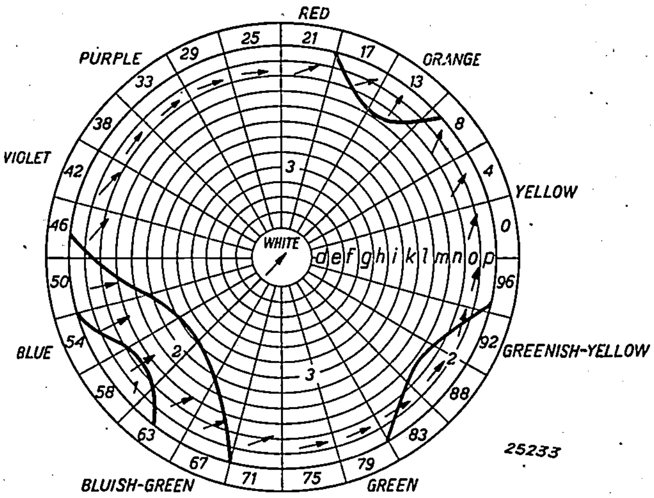


Fig. 2. Comparison of the colour reproduction by an electric lamp and by daylight. The arrows indicate how the colours are shifted upon transition to electric light.

Only with some very unsaturated colours, especially with orange, is the shift so large that we must speak of a totally different colour (region 3).

We must therefore pay particular attention to the direction of the displacements occurring at such points. As was to be expected the arrows point in general away from the saturated yellow.

The majority of the colours have become more saturated, a change which is to be desired especially at relatively low levels of illumination. In particular the blue, which upon changing from daylight to incandescent light has become considerably less saturated (fig. 2) is again reproduced in a more saturated form.

The orange is shifted toward the red: the shift in the direction yellow to red is in general experienced as an increased "warmth" of that colour. This shift also gives a more healthy appearance to the complexion.

The green, which upon transition from daylight to incandescent light had become a somewhat dubious yellow-green, goes back to green again under the influence of "Philiphane" glass.

Finally we note that white and the very unsaturated colours are shifted in the direction of blue-violet. This may certainly not be considered an advantage; since however the change is not very great (less than $^1/_3$ of the difference between incandescent light and daylight), and moreover since it lies almost in the same direction as the shift on transition from daylight to incandescent light, it is not very disturbing.

The results found may be summarized as follows.

The use of "Philiphane" glass has the advantage

[2]) D. Judd, J. Opt. Soc. Amer. 25, 24, 1935.

of reproducing most colours in a more saturated form and of making the orange-yellow warmer. Various disadvantages in the use of incandescent light, such as the faded appearance of blue and the shift of green toward yellow-green, are partially overcome. The most important advantages of electric light such as the high saturation of orange and of the colours in its neighbourhood, the greater intensity of red, etc. are retained.

In those cases where the first condition of the illumination is the creation of a pleasant, cosy sphere in which many warm colours occur and in which we ourselves appear especially well, "Philiphane" glass is actually able to give a better reproduction of colour than ordinary incandescent light. This is especially true in the case of its use in hotels and restaurants, for illumination of certain show windows such as those of provision shops and florists etc.

A special warning must be given against considering the use of "Philiphane" glass as an attempt to imitate daylight. For this purpose quite other methods must be used.

The improvement in colour is of course accompanied by an inevitable loss in light intensity. In the case of "Philiphane" glass this loss is about 20 percent. If it is desired to increase the colour effect by the use of thicker or more deeply coloured glass, the loss of light will also increase.

We must add a few words about the ratio of the brightness of the colours with and without "Philiphane" glass.

For several of the colour cards with fairly saturated colours used in the above described comparison the point is determined in the colour triangle upon illumination with each of the two kinds of light, under the simplified assumption that Neophane glass absorbs completely the radiation from 570—600 mμ. The shifts were entirely in the directions already found in fig. 1. From the same calculation we found the coefficients of reflection for both cases. These are tabulated in *table I*.

Table I.

No.	Colour	Coefficient of reflection	
		ordinary	"Philiphane"
	Ideal white	100	100
nc 25	red	36	32
nc 17	orange	40	34
nc 4	yellow	57	55
nc 83	green	23	25^{5}
nc 58	blue	17	19
nc 42	violet	17	17
nc 33	purple	31	29

We see that no very great differences appear. There can therefore be no question of the "disappearance" of any colours. It is true that the differences can become somewhat greater with very saturated colours, but on the other hand "Philiphane" glass always continues to transmit a part of the yellow. In this respect therefore we need not fear unpleasant surprises.

Finally we note that "Philiphane" glass may have other practical applications besides its use in envelopes for electric lamps. The advantages offered by the use of this glass can be of two sorts:

1) The glass offers the possibility of removing an unpleasant excess of yellow from a mixture of several kinds of light. Especially in combination with mercury lamps such cases may occur.
2) As may be seen immediately from fig. 1, "Philiphane" glass displaces the red and green in opposite directions, so that the difference between these colours becomes more pronounced. The use of spectacles of "Philiphane" glass by persons with a poor sense of colour might be able to make easier the distinction of coloured traffic signals.

APPLICATIONS OF CATHODE RAY TUBES I

by H. VAN SUCHTELEN.

621.317.755 : 621.385.882

An early use of cathode ray tubes was in the study of alternating electric potentials. In comparison with mechanical oscillographs, such as the loop oscillograph, the cathode ray tube has the advantage that the beam of electrons has practically no inertia, and that therefore very high frequencies and phenomena of very short duration can be made visible. The rising technique of radio communication immediately seized upon the cathode ray tube as a welcome instrument of investigation. Due to the presence of two sets of deflection plates the beam of electrons can be deflected in two mutually perpendicular directions; this offers the possibility of examining the behaviour of a voltage as a function of the time in a very simple way, or of examining the relation between two voltages. In the first case the set of plates for horizontal deflection obtains its voltage from a "linear time-base", that is, a potential which increases linearly with the time for the whole period, and then falls back to its initial value again. In the second case one of the two voltages is applied to each set of deflection plates. If for instance the two alternating voltages are sinusoidal with an integral ratio between their frequencies, so-called Lissajous figures result.

The voltage necessary to obtain good deflection of the beam is about 20 to 60 volts, the load on the source of potential is then nearly zero, since a capacity of a few $\mu\mu$F only is present as loading impedance between the deflection plates. It is obvious that this will sometimes be an important advantage, especially in working with high frequencies.

One very good feature of the newer cathode ray tubes is the fact that the fluorescence image on the screen is sufficiently brilliant to be observed in fairly light surroundings, thanks to its great brilliancy, so that photography is unnecessary unless it is desired to record a particular result. The direct observation of the fluorescence image is obviously of great advantage, since, as in the investigation of particular circuits, the result of changes in the circuit may be seen immediately on the screen of the cathode ray tube.

The technical development of cathode ray oscillographs has taken these possibilities into consideration. In addition to large permanent set-ups of oscillographs with complete photographic equipment, more and more small portable equipments have recently come into use. In these smaller sets the cathode ray tube with its power supply, amplifier and time-base generator and synchroniser is built as a compact unit [1]), so that it becomes very simple to transport the oscillograph to the place where the phenomenon to be investigated is taking place. This is one of the factors which have led to the more general use of the cathode ray tube as a measuring instrument in the investigation of innumerable phenomena. In a series of short articles we shall discuss some of the applications. The articles will be illustrated by photographs of the oscillograms [2]) which may be observed on the screen of the cathode ray tube.

Oscillograms of power mains

The most obvious oscillograms of alternating currents or voltages are the curves recorded for power mains. The oscillograms of *figs. 1* and *2* offer little that is new to the electrical engineer. Upon careful consideration it will be seen that the curve of fig. 1 is slightly more angular than that of fig. 2. Such differences may be of influence on the calibration of voltmeters or wattmeters. Although the two curves in this case were obtained from two different mains, the same mains can also give different curves according to the nature of the load. It is sometimes important to know whether or not a given load on the mains, added to the circuit at a point far from the power station, will cause a distortion of the voltage curve, and to know the extent of such deformation. In such cases a convenient portable oscillograph offers great practical advantages. A rapid demonstration on the spot in the form of an oscillogram can obviate long discussions.

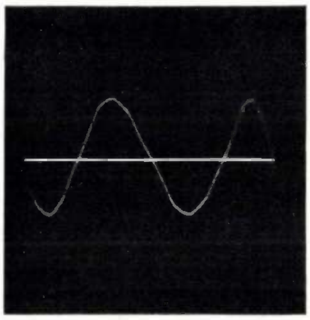
Although upon comparison of fig. 1 and 2 the latter appears the least deformed, it is not permissible to conclude immediately that the latter source of voltage will in every case give less departure from normal behaviour. If the voltage to be investigated is not connected directly but through a simple RC circuit (*fig. 3*) with the cathode ray oscillograph, the main frequency may be very much suppressed with respect to harmonics of

1) Such an apparatus was described fully in Philips techn. Rev. 1, 147, 1936. The cathode ray tube of the type DG 9-3 with a screen diameter of 9 cm is used. The following oscillograms were recorded with this apparatus.

2) The photographic recording of oscillograms is usually very simple. The photographs in this article were made with a very simple box camera and an extra lens; time of exposure about $^1/_2$ sec.

much higher order. Afterwards these harmonics which have been filtered out may be amplified until they give a clearly visible amplitude on the cathode ray tube. In this way a new oscillogram

1



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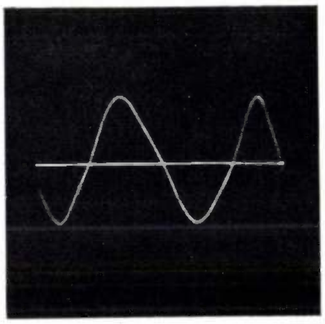

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Fig. 1 and 2. Oscillograms of the voltage of two different 50 cycle power mains. The first mains voltage appears to have slightly more deformation.

(*fig. 4*) is obtained with the voltage of fig. 2. In fig. 4 fine indentations may clearly be seen which indicate a frequency very much higher than the 50 cycles of the mains. Such a ripple, which may be caused by the slots in the rotor of the alternating current generator, was found to present difficulties in the initial development of radio valves heated with alternating current. An oscillogram like that in fig. 4 gives a decisive answer to the question as to the cause of the interference tone heard. At the same time it may be concluded from the evidence which the oscillogram presents that the interferences, due to their high frequency, could be suppressed by connecting a fairly simple filter into the mains circuit. This was actually found to be the case.

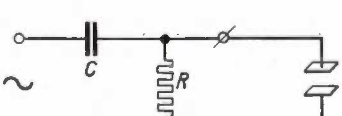

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Fig. 3. A simple filter is connected between the mains and the deflection plates of the cathode ray tube in order to decrease the intensity of the fundamental frequency of the alternating voltage between the plates.

It will have been noticed that the general character of the curve in fig. 4 differs very much from that of figs. 1 and 2. The horizontal deflection is in this case not obtained from a linear time-base,

24858

Fig. 4. The oscillogram of the mains of fig. 2 recorded with the use of the filter of fig. 3. The horizontal deflection is obtained from the mains voltage and not from a linear time-base. With a vertical deflection obtained by means of a sinusoidal alternating voltage of 50 cycles, an ellipse will generally be formed The fine indentations of the approximately elliptical line indicate disturbances which may, for example, have their source in the slots of the rotor of the alternating current generator.

but from the 50 cycle alternating voltage of the mains. Since the filter does not completely suppress the fundamental frequency, an alternating voltage of 50 c/s is also present on the other set of deflection plates. The frequency of the latter is shifted in phase with respect to the mains voltage, and thus gives rise to the elliptical figure. This device is used in this case in order to make sure immediately without careful adjustment that the time-base is accurately synchronized with the mains voltage. This is important for the purpose of deciding whether the fine indentations have their source in the mains themselves or are due to an external source of interference. In the latter case the ripple will never be completely stationary, but will run more or less rapidly along the ellipse.

Finally a practical example will show how quickly it is possible to work with the cathode ray oscillograph. As was mentioned, an *RC* filter was necessary to obtain the curve of fig. 4. The resist-

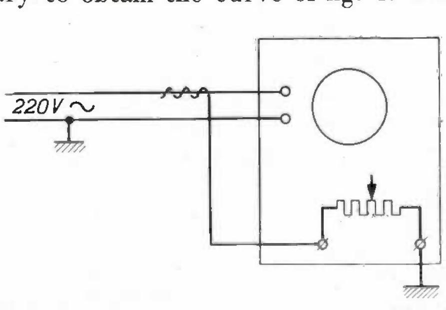

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Fig. 5. Practical construction of the filter circuit represented in fig. 3.

ance was here formed by the input potentiometer of the built-in amplifier, while the capacity was formed by winding an insulated wire connected to the input terminal of the amplifier several times loosely around the unearthed power mains cable (*fig. 5*).

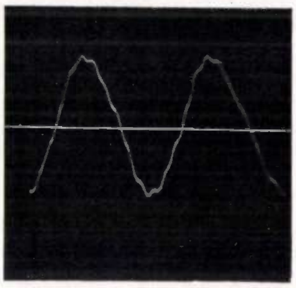

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Fig. 6. Oscillogram of "standard mains" which are used in laboratory to test radio receivers.

Phenomena such as that illustrated by fig. 4 led to the construction of "standard mains" with an artificial interference by overtones. In the Philips Laboratory such mains of 50 c/s with 3 per cent overtones of 500 cycles are used for testing radio receivers. The oscillogram of these mains is given in *fig. 6*.

The oscillograms of direct current mains also may sometimes be important. When the mains are not fed from an accumulator battery a ripple will usually be observed superposed on the straight line which represents the direct voltage. *Figs. 7, 8* and *9* give examples of what may be encountered in practice. Fig. 7 is the oscillogram of 130 volt mains from a source of 40 kW, used in connection with various small motors. The ripple, which is due in general to the dynamo, can scarcely be distinguished. At night, instead of the 40 kW dynamo, a 6 phase rectifier of lower power is used. This makes no difference to the running of the machines connected with the mains. On the oscillogram, however, the successive crests of the sine curves of the six alternating current phases can be very clearly distinguished. The short oscillation of much higher frequency at the end of each crest is typical. It is from such an oscillation that interference may be expected. We shall return in a following article to give an example of this.

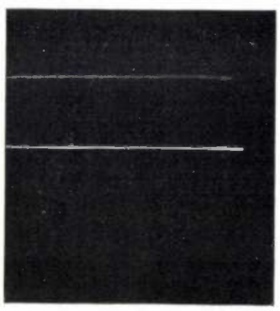

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Fig. 7. Voltage of direct current mains fed by a 40 kW dynamo.

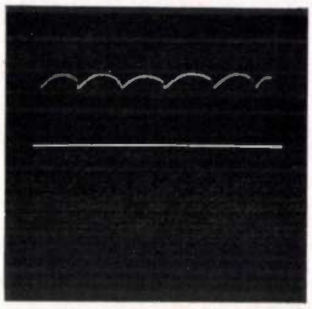

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Fig. 8. Voltage of direct current mains fed by a 6 phase rectifier. The voltage exhibits a pronounced ripple, formed by the crests of six sine curves. In addition, at the end of each section of sine curve an oscillation of much higher frequency may be seen.

A third example of direct current mains is the oscillogram in *fig. 9* of a 5 kW motor generator. With this smaller machine the ripple is much larger than in the cases of the more powerful machine.

If it is desired to examine the ripple of direct current mains in more detail, the same device may be employed as that with which fig. 4 was obtained, but in this case of course using a normal linear time-base.

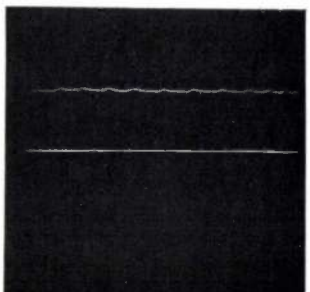

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Fig. 9. Voltage of direct current mains fed by a 5 kW dynamo. Smaller machines often have a much larger ripple than larger machines (fig. 7).

WORK-BENCH ILLUMINATION

by N. A. HALBERTSMA.

628.976

Fine assembly work demands a very intense illumination of the work shop. Because of the fact that the sensitivity to glare increases very much with the level of brightness, the illumination must in addition be very uniform. In this article the points of view are discussed which determine the design of fixtures for an illumination of about 100 ft. candles on the working surface. General as well as local illumination is discussed. The examples give the development of the illumination of an assembly bench for the making of radio receiver valves.

The illumination of the work-bench and the machine-tool where fine work is done has passed through a remarkable development in the last 25 years. Local illumination, upon which one had to rely out of considerations of economy in the time of the carbon lamp and relatively high prices for current, was gradually displaced by general illumination when, with the large metal filament lamps (especially the gasfilled lamp), light sources with a much higher efficiency and units with a considerable luminous flux became available.

With increasing insight into the value of good and strong illumination for the rapid and accurate carrying out of fine work, there is now a tendency to revert to local illumination without neglecting general illumination.

For the correct evaluation of the two very different methods of illumination it is important to find out what objections there are to general illumination when it is desired to raise the illumination to high levels (around 100 ft-candles and higher), and to what extent these objections can be met by local illumination.

We have in mind operations which make greater demands on the eye than writing or reading (for which purposes general illumination of sufficient intensity can be easily obtained).

These are many such fine operations. It is well known that drawing makes greater demands than writing. Fine lines and the slightness of contrast in the case of lines drawn with a hard pencil or in tracing etc. require many times the light intensity with which one can write comfortably.

At the work-bench and on the machine-tool the small dimensions of the details to be observed or the slight contrasts create a demand for a strong illumination.

In a report of the English "Industrial Health Research Board" and the "Illumination Research Committee" [1]) Weston arrived at the following table of illumination required with decreasing size of details to be distinguished:

Table I.

Visual angle of the details to be distinguished	Dimensions at a distance of 14 inches	Required illumination
10′	1 mm	1.6 ft. candles
8′	0.8 ,,	1.8 ,,
6′	0.6 ,,	2.4 ,,
5′	0.5 ,,	3.2 ,,
4′	0.4 ,,	5.4 ,,
3′	0.3 ,,	24.0 ,,
2′	0.2 ,,	80.0 ,,
1′	0.1 ,,	160.0 ,,

As test object Weston used cards with Landolt rings of the type shown in *fig. 1.* The test person had to mark the rings whose slit was in one of the eight positions given. As a measure of the performance of the eye Weston took the number of rings marked in a given time. The number generally increases in such tests with the size of the rings on the card and the intensity of the illumination. There is, however, a limit to the speed of observation, even although the light intensity is increased to that which occurs in the summer in full sunlight (about 10 000 ft. candles). The "required illumination" according to the above table is the illumination at which this limit has been reached within 2 per cent. If it is desired to pass this limit, spectacles or a magnifying glass must be used [2]) in addition to strong illumination.

In the table is the width of the slit, it may be noted, that the slits as printed in fig. 1 subtend a visual angle of about 3′ at normal reading distance. To give some idea of this angle, it may

[1]) H. C. Weston. The relation between illumination and industrial efficiency I. The Effect of Size of Work, London, His Majesty's Stationery Office, 1935.

[2]) The magnifying glass of the watchmaker and engraver is the best known example of this, but H. C. Weston and A. Adams state in Report No. 40 of the Industrial Fatigue Research Board, London 1927, that operators of knitting machines with normal eyes needed spectacles of + 1,5 D when they had to transfer the stitches of the stocking to needles on the knitting machine of which there were 20 to 22 per inch and whose separation therefore was slightly more than 1 mm.

be noted that the angle is also about 3′ for the details of the type of this periodical at a normal reading distance.

In Weston's tests the contrast was as great

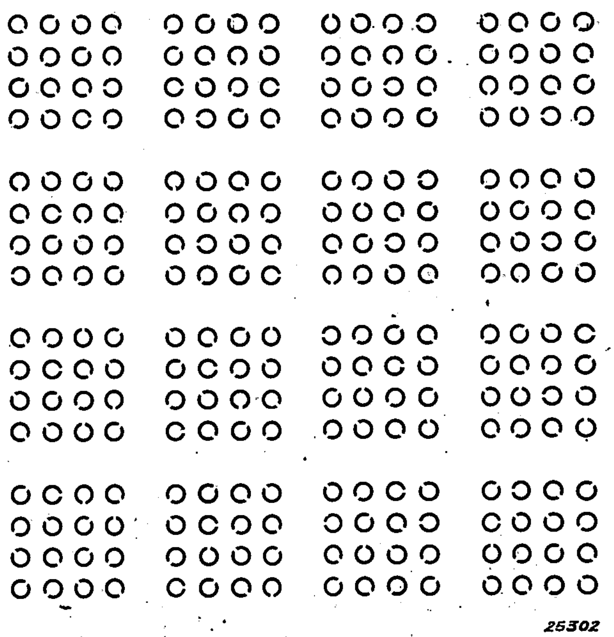


Fig. 1. Test card for the determination of speed of observation. The visual angle of the details to be distinguished is 3′.

as possible (black and white). Lower contrasts require greater intensity of illumination.

The curve of the sensitivity of the eye to contrast as recorded by König-Nutting (*fig. 2*) shows that contrasts with ratios of brightness of 1 : 1.02 ($H/\Delta H = 50$) at a brightness of 1 cld/sq. ft., that is at intensities of illumination of 300 lux on a surface with 10 per cent reflective power, are no longer observed.

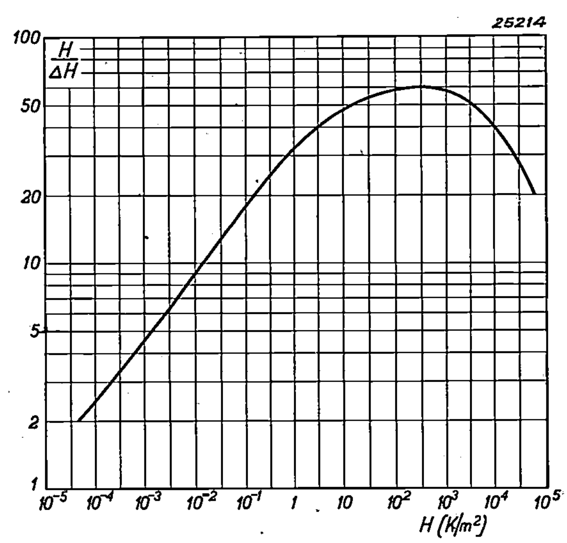


Fig. 2. The smallest observable contrast in brightness $H/\Delta H$ as a function of the brightness. H is the lowest of the two brightnesses, ΔH the difference in brightness. The ratio of brightness is therefore $1 + \Delta H/H$.

The high standard of accuracy which is now often required to ensure interchangeability of parts, has in many factories necessitated high intensities of illumination (100 ft. candles for example) at the work-bench.

Objections to very intense general illumination

The following difficulties are encountered with very intense general illumination:

a) The operating costs and the heat development increase proportionally with the power installed and become important at 10 W/sq. ft. and higher. The development of heat can become disturbing not only because of the heating of the air, but also due to the radiant heat (the use of the mercury vapour lamps e.g. decreases the heat developed at equal intensities of illumination by about one half).

b) In rooms of moderate height (12 to 14 ft.) the brightness of the fixtures quickly becomes disturbing when the intensity of illumination is increased to a large degree. Ward Harrison [3]) pointed out that two phenomena must be kept in mind in designing installations with high intensities of illumination. The first is that the contrast in brightness, between the lightest and the darkest parts of the field of vision, which is permissible in connection with the risk of glare, decreases with increasing of brightness of the field of vision. This fact was discovered by Nutting [4]). With tenfold increase of brightness of the field, the maximum brightness may only be doubled if the size of the source of light remains the same. In the second place account must be taken of the fact that increase in size of the brightest parts of the field of vision (the lighting units, which in this case were diffusing globes) decreases the maximum permissible brightness. This practically means that no use can be made of the above-mentioned doubling, and that one is limited to a maximum brightness. Therefore with increasing light flux from the light sources the apparent surface of the diffusers must in fact increase in the same ratio. With reflectors, and with indirect illumination this must be kept in mind. Entirely apart from economic considerations, increase in the number of light points, for instance to 1 per 20 sq. ft., provides no solution of the difficulty met in a radical increase of the level of illumination (up to 100 ft. candles for example).

c) In long factory rooms the workmen are distracted by the glare of the large number of

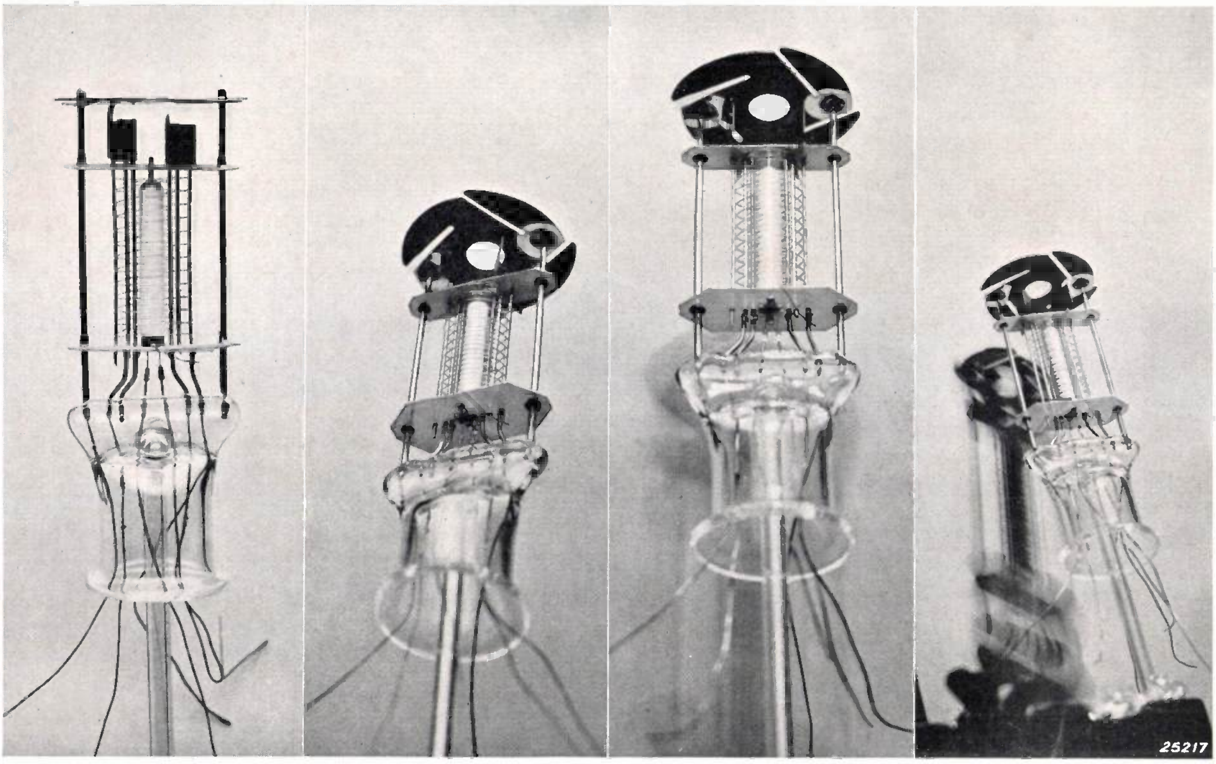
[3]) W. Harrison. What is wrong with our 50 ft candle installations? Trans. Ill. Eng. Soc. **32**, 209, 1937.

[4]) P. G. Nutting, Trans. Ill. Eng. Soc. **11**, 943, 1916.

fixtures or the strongly lighted ceiling where the brightness is a multiple of the brightness of the working surface toward which his eye is directed.

d) In the presence of a large number of closely spaced lighting points the intensity of the shadows will as a rule be small, but on the other hand, with pieces of work made of shiny material, there is an increase in the number and intensity of the bright spots caused by specular reflection.

eyes of the worker not supporting a diffusing bulb in the immediate neighbourhood. The proximity of the lighting unit to the head of the worker, however, makes it necessary to make special provisions against radiant heat. Neither the hands nor the head of the worker should feel the presence of heat radiation. If the reflector becomes very hot, it also becomes a source of radiant heat. Generous dimensions and good ventilation for removing the warm air which collects in the reflector will keep the temperature of the reflector low.



a) b) c) d

Fig. 3. The inside of a pentode, photographed under various systems of local illumination.
a) and b): Large enamelled reflector with lamp shielded from below. Different positions of the piece of work. The supporting wires may, according to their position, appear bright or dark against the background, but exhibit no disturbing light reflexions.
c): Lamp, with inside frosted bulb, not shielded from below. Disturbing light reflexions appear.
d): Lamp with clear glass bulb and a very small reflector. The reflexions become stronger. In addition a disturbing shadow appears.

Characteristics of strong local illumination

We shall now try to see to what extent the above-mentioned objections to strong general illumination can be met by strong local illumination.

a) The heating of the air is less, because even if one assumes 10 sq. ft of work-bench space per man, and installs 100 W for that space, the number of work-bench places in a room still remains so much smaller than the total floor space in sq. ft that the heating of the air will be much less owing to the smaller current consumption.

As lighting unit a reflector only can be used, the

b) A reflector for local illumination which is so formed and so placed that the worker sees neither the lamp nor any part of the strongly lighted inner surface of the reflector [5]) will not be blinding by itself. On the other hand very disturbing light reflexions will appear on a piece of work made of polished material. These reflexions are caused by

[5]) When a small portion of the inner surface of the reflector remains visible, this cause of dazzle may be removed by painting that part dull black.

the lamp as well as by polished parts of the inner surface of the reflector. The brightness of an inside frosted lamp of 75 W is of the order of 150 cdl/sq. inch. Under this lamp metal objects will show reflexes of about 60—120 cdl/sq. inch, enamelled or lacquered objects (with a direct reflection power of 4 per cent only 6 cdl/sq. inch). If, however, the lamp is shielded from below, there will still be reflexions of the lamp in an enamelled reflector of 6 cdl/sq. inch which cause reflexions of 3 cdl/sq. inch in metal objects.

c). Since the fixtures for local illumination are at about eye-level, it is also important for the other workers in the hall not to be able to look into the reflector. Otherwise the objection to general illumination mentioned under c) above would hold equally and perhaps still more for local illumination.

d). Shadow formation may be of great importance according to the mixture of the work to be done. Sometimes it is desirable, in other cases every attempt is made to avoid it. While in the case of general illumination we have long been aware of the influence of the system of illumination, direct, half indirect or indirect, on the formation of shadows on the piece of work, it has only recently been realised that local illumination can also be adapted to the requirements of shadow formation. Reflectors with small diameters which used to be used exclusively for local illumination will always produce strong shadows. With increasing diameter of the reflector the shadows become less pronounced because the light coming from a reflector with a diameter of 12 or 15 inches subtends a fairly wide solid angle because of the short distance between reflector and work. At a distance of 13 inches this solid angle is approximately equal to the solid angle of the incident light in indirect lighting where a surface of about 7 ft diameter on a ceiling 12 ft high radiates the light.

Fig. 4. Work-bench in an assembly department with the illumination whose results are shown in figs. 3a) and b).

The lighting of an assembly bench.

The ideas on local illumination of high intensity (about 100 ft. candles) summarized in the foregoing have led to a special lighting installation on the work benches in the radio valve department of the Philips Incandescent Lamp Works, where in 1934 repeated complaints of eye troubles occured with persons who were employed in the assembly of radio valves.

This assembly work must be considered as the limit of performance of the eye, since in addition to the handling and selecting of often very fine parts, the eyes are not focussed on a single plane surface only, as in the work of a watchmaker, but they must be focussed at different distances in space.

The modern radio valve contains a number of concentric grids which are born by supporting wires insulated from each other. Helices of fine wire are used as grids.

These bare metal wires can be observed in two ways. They can appear dark against a light background and thus be seen as silhouettes, when the source of light is not reflected in the direction of the eye by the bright metal (see *fig. 3a*), or they may appear bright in contrast to the background when they reflect the light.

Whether or not the reflection is observed by the eye depends upon the position of the work. The same supporting wires, in the position shown in *fig. 3b* are lighter than the background due to reflection of the light source.

In order to prevent the reflexion in the metal parts and also in the glass from becoming too strong and causing dazzle, and in that way making observation in the immediate neighbourhood more difficult, the surface brightness of the light source must not be too great. The lamp is therefore shielded from below in the reflector. The surface brightness is then of the order of magnitude of 3 cdl/sq. inch at the most, if the reflector is dull white on the inside. If the enamel of the reflector has a shiny surface, reflexions of greater brightness about 30 cdl/sq. inch, occur, and the brightness of the metal wires in certain positions also becomes greater.

The importance of the shielding of the lamp may be seen from a comparison of the illuminations obtained with shielding (figs. 3a and b) and the illuminations obtained without shielding (figs 3c and d).

The omission of shielding has two results when a lamp of normal type, matted on the inside, is used:

1) The light reflexions on the metal wires and in the glass become much stronger (see fig. 3c).
2) Heavy shadows occur.

Both phenomena become more pronounced if a clear glass lamp and a very small reflector are used. In fig. 3d) the strong reflexions and heavy shadow may be seen.

The impressions given by these photographs have been confirmed by tests on a large scale.

With an illumination as that used in figs 3a and b the production per hour was highest and the number of mistakes smallest. The number of complaints of eye trouble also decreased considerably.

Considering the nature of the work and the great demands which it makes on the eye, a careful selection of personnel by means of a thorough eye examination is to be desired.

The enamelled reflector, matted on the inside (*fig. 5*) and 15 inch in diameter is fastened 15 inch

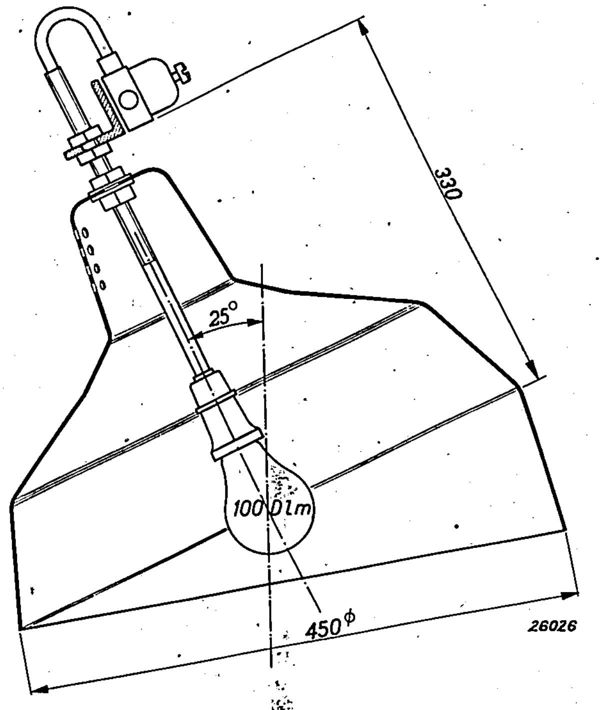
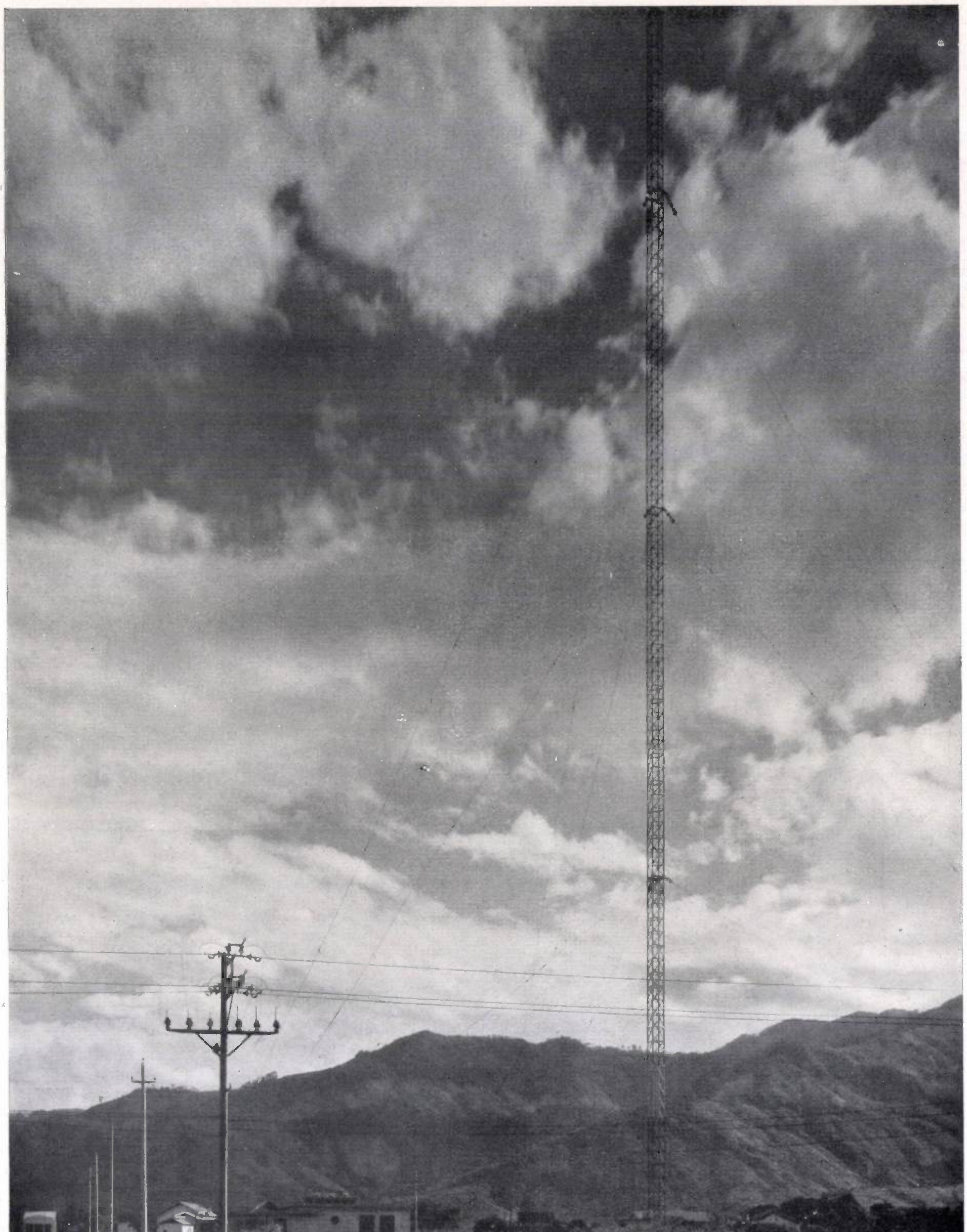


Fig. 5. Sketch of the enamelled reflector for the work-bench illumination of figs. 3a and b.

above the work-bench in a slanting position. The source of light is a lamp of 75 Watt with a clear glass bulb.

Directly under the lamp on the work-bench there is an illumination of about 55 ft. candles. As a rule the work is held at some distance above the bench. At this point a light orange-coloured, matt-finished, metal plane is placed at an angle of 75 ° with the horizontal. This serves as background for the work to be inspected.

When the work is held before this plane, the horizontal intensity of the illumination is about 80 ft. candles, and the light falls from the wide opening of the reflector on the work from all sides and through the various wires in such a way that the parts in the interior of the piece of work can also be seen without hindrance from shadows. The large surface of the reflector and the ventilation holes in its neck prevent it from becoming hot, and thus prevent disturbing heat radiation on the head of the worker.



Transmitter building (middle) with cooling tower (left) of a 20 kW Philips medium wave broadcasttransmitter at Rio de Janeiro.

In the background can be seen part of a self-radiating aerial tower belonging to this station.

A ROTATING DIRECTIONAL AERIAL FOR SHORT WAVE BROADCASTING

by P. J. H. A. NORDLOHNE. 621.396.677.029.58

A description is given of the rotating directional aerial erected at Huizen for broadcasting from the transmitter PCJ on 31.28 m.

On November 16th 1937 a rotating directional aerial was put into use at Huizen for experimental transmission by the short wave station PCJ described in the previous number, with a frequency of 9.59 megacycles (wave length 31.28 m). This directional aerial was designed for a power of about 75 kW carrier wave energy.

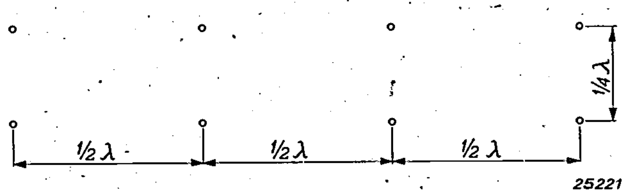


Fig. 1. View from above of a uni-directional aerial.

The purpose of building a rotating aerial was to make possible the restriction of the energy to any chosen direction with a single aerial installation, and in this way to concentrate all the available energy toward a given part of the world, for example toward a given part of North, Central or South America.

Instead of a rotating aerial, up to now it has been usual to build a number of permanent fixed directional aerials for this purpose, each of which served for a single definite direction. The angle of emergence of the beam sent out by a directional aerial can be chosen more or less wide. If aerials with small angles of emergence are built, a narrower section of the earth's surface is reached; but with greater intensity; if, on the other hand, the aerials have large angles of emergence a wider section of the earth's surface is covered, but at the cost of intensity.

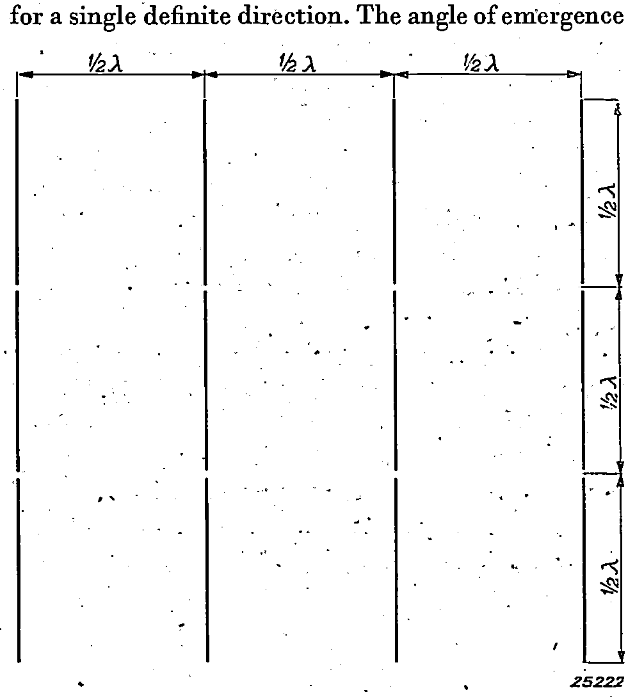


Fig. 2. Position of the half wave aerial wires in the vertical plane.

Considering the fact that the wave length with which one serves a certain distant part of the world is almost always less suitable for another distant part of the world in the opposite direction, since, for example, the state of ionization along the path to the latter is different, it is advantageous to make the directional effect of the aerial one-sided. Thus the aerial must not radiate any energy in

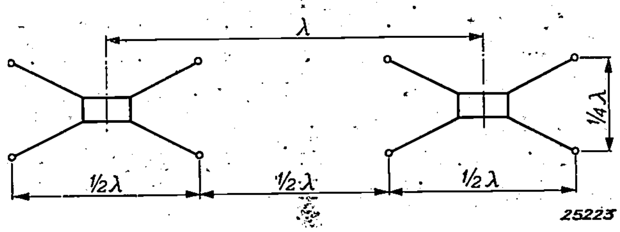


Fig. 3. View from above of the towers with the aerial wires hung on the projecting supports.

a backwards direction. This uni-directional action can be attained by grouping the various aerial wires in two parallel planes which lie at a distance of a quarter wave length from each other, and by giving the currents in front and rear planes a phase difference of 90°.

Fig. 1 gives a view of the uni-directional aerial from above. The supply is such that the currents in the aerial wires of the front plane are 90° in phase behind those of the wires of the rear plane, so that the waves which the two planes radiate, when they have the same amplitude, will reinforce each other toward the front and exactly neutralize each other toward the back.

The sharpness of the beam depends upon the number of aerial wires next to each other in each plane and on the mutual separation of the wires. If directional aerials for a given frequency are available, the intensity of which falls to one half at 13° on either side of the main direction, and if it is desired to reach all parts of the world from East Africa to and including the United States with an intensity which is never less than half of that in the main direction of the aerial, eight such aerials would be necessary. These directional aerials should be placed along the sides of a polygon to prevent one aerial from exerting too much influence on the radiation diagram of another. The transmitter building may be situated at the centre of the circumscribed circle of the polygon. With this method very large areas of ground are necessary, as well as one or more complicated and expensive switching arrangements.

These objections are avoided with the rotating aerial.

For the rotating aerial of PCJ for a wave length of 31.28 m two 60 m wooden towers were built which can rotate together around a central pivot. In connection with the desired concentration of radiant energy in a horizontal as well as in vertical plane, four aerial wires a half wave in length were suspended next to each other, and three were suspended one above the other in the same vertical plane (*fig. 2*), while two of such planes were placed one behind the other as shown in fig. 1. These aerial wires may be hung in the usual way between the two towers, but in that case the diameter of the circle described by the rotating towers is relatively large. When steel masts are used, moreover, in order to limit the currents induced in them, which would have an undesirable influence on the radiation pattern, the distance from the aerials to the poles would have to be made equal to at least about one wave length, so that the structure would be unreasonably large. The diameter of the circle described can be considerably decreased by using wooden towers and by not suspending the aerials between them, but on the towers themselves. In this case the towers must be provided with projecting supports in the form of platforms (*fig. 3*). Each tower then carries four vertical wires, each of which is divided by two insulators into three half wave aerials whose currents are in phase with each other. Since the distance between the aerials in the front as well as the rear plane is fixed at a half wave length, the distance between the centres of the towers becomes exactly one wave length.

In order to allow the whole structure to rotate easily, the two towers must be joined together at their bases. They must therefore be mounted in the same rotating undercarriage. The latter consists of a steel construction about 54 by 13 m which rests at each of the eight points where it bears the towers on a set of wheels (*fig. 4*) consisting of two wheels in a line (16 wheels in all). Four of these sets of wheels are on the outer track, the

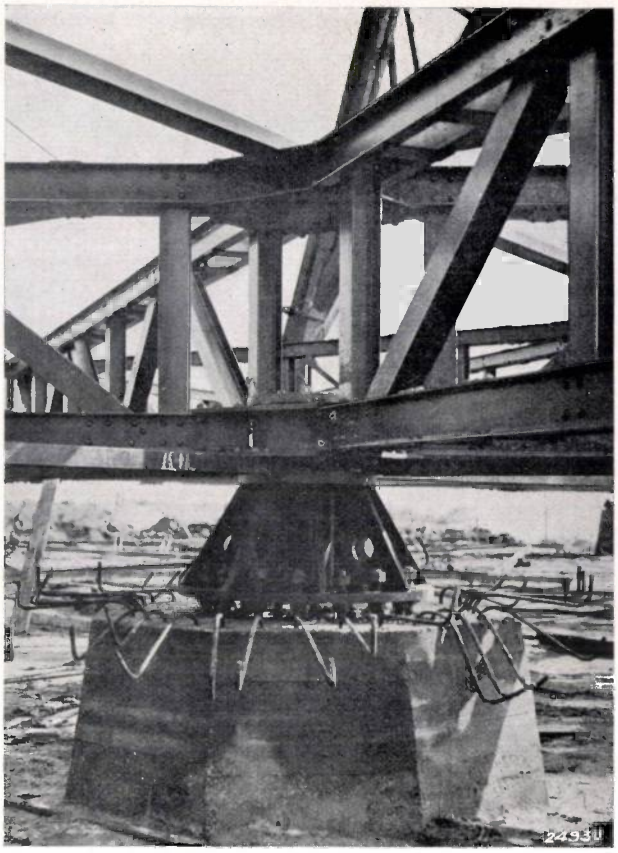


Fig. 5. Anchorage of the steel pivot about which the whole rotating aerial is centred.

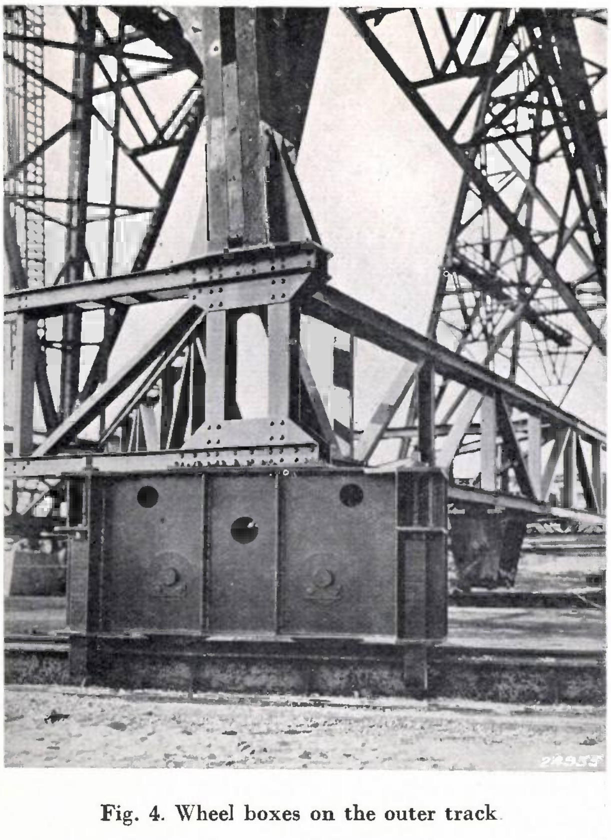


Fig. 4. Wheel boxes on the outer track

other four on the inner track. The outer track has a diameter of about 46.4 m and consists of a normal rail AH no. 1, welded to an I-girder, DIN 24, which in turn is anchored in a (circular) wall of reinforced concrete. The inner track is constructed in exactly the same way but has a diameter of about 22.4 m.

Since during rotation in a strong wind the wheels might be pressed against the sides of the rails, the whole structure is centred about a steel pivot which is anchored in a block of reinforced concrete with a heavy base (*fig. 5*). This pivot therefore serves exclusively to take up the horizontal forces.

The wooden towers are made of Norwegian fir impregnated with a substance to prevent rot. The general construction can be seen clearly in *fig. 6*. All joints are made with nuts and bolts, while several important points are provided with iron butt straps. Each wooden tower weighs 18 tons, the steel framework weighs about 40 tons, while each set of wheels weighs 1.5 tons.

The wheels rotate on double ball bearings. There is one set of wheels under each of the four posts of the two towers.

The heaviest load occurs as the result of a wind diagonally against the towers. There is then a 38 ton compressive stress on one side and a 38 ton tensile stress on the other side. In this way a maximum load of 51 tons may occur on one set of wheels: 4.5 tons from the wooden tower, 7 tons from the steel undercarriage, 1.5 tons from the wheels themselves and 38 tons due to the force of the wind. In the construction of the towers the calculations were made, according to the Netherlands Normalization Committee, with a wind force of 70 kg/m^2 for the first 20 m of height, while for every additional meter of height an increase of pressure of 1.5 kg/m^2 is calculated. This increase, however, goes only to a height of 40 m, so that above

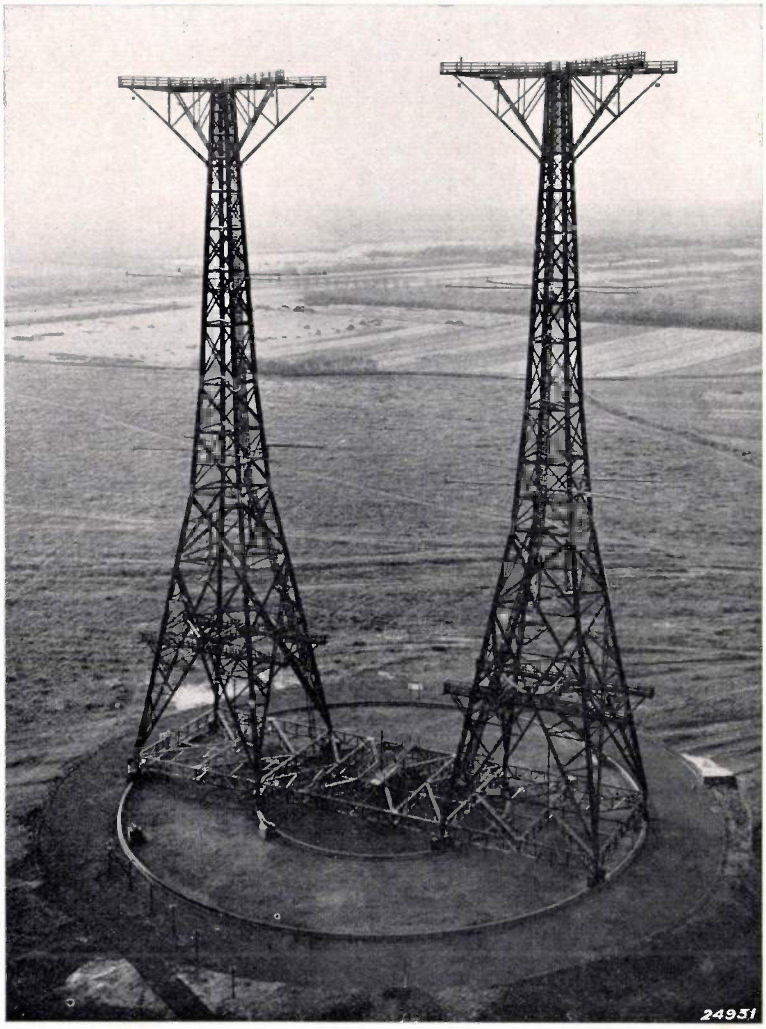


Fig. 6. The wooden towers of the rotating aerial.

this height the pressure remains 100 kg/m². The wind pressure makes heavy demands on the wood construction, considering the fact that the surface on which the wind exerts its force is much greater than that of an iron tower, which can always be constructed in a more slender form. Moreover, the platforms in the shape of a cross, on whose extremities the aerial wires are hung, represent loads which are effective at the top, and thus make large contributions to the moment of the forces acting on the tower.

The construction of a tower was carried out in

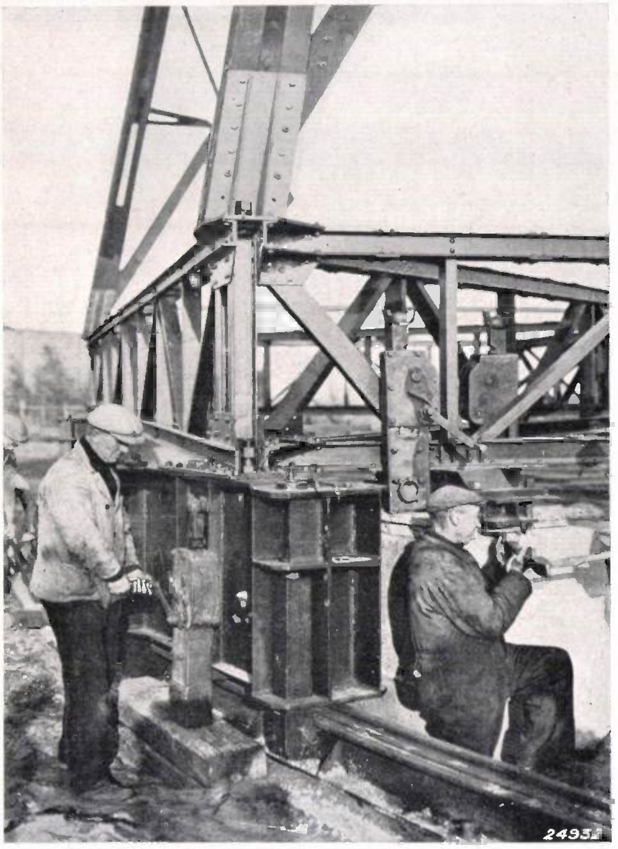


Fig. 7. The bridge with the towers is freed from the concrete blocks and rests upon the wheels.

the following way. Two sides about 20 m high were first built on the ground. After these had been erected in their places the necessary struts etc. were added to complete the other two sides. The tower was finished with the help of a hoist and tackle and temporary scaffolding.

The iron framework was also made ready in large sections which were put together on the spot. The specially constructed wheels are mounted in practically closed long narrow boxes (cf. *fig. 4*). Since the wheels could not be delivered quickly enough, and it was impossible to wait for them because of the lateness of the season, the framework and towers were first mounted on eight concrete blocks. It was then possible, during the construction of the towers and the putting together of the

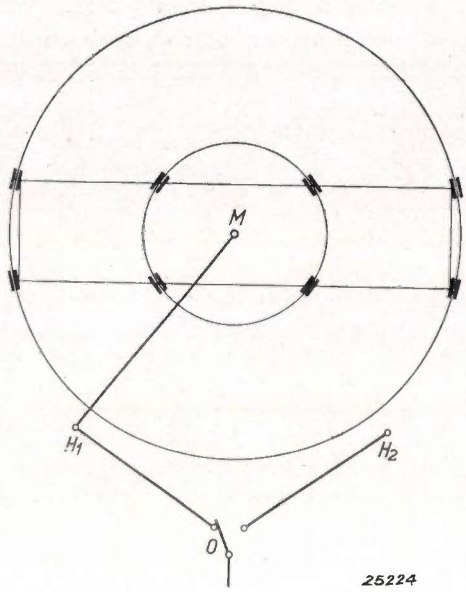


Fig. 8. Diagram of the supply of the aerial. The aerial can be fed from O *via* H_1 or H_2 to M.

framework, to level the ground, cast the concrete walls for the tracks and to fix the I-girders and finally the rails.

The last operation was that of fastening the wheel boxes underneath the framework, and then freeing the connection between the framework and the eight concrete blocks (*fig. 7*). During this operation the framework was supported by jacks. The whole was easily lowered a distance of 0.5 cm on to the rails.

Rotation, is, for the time being, carried out by hand,

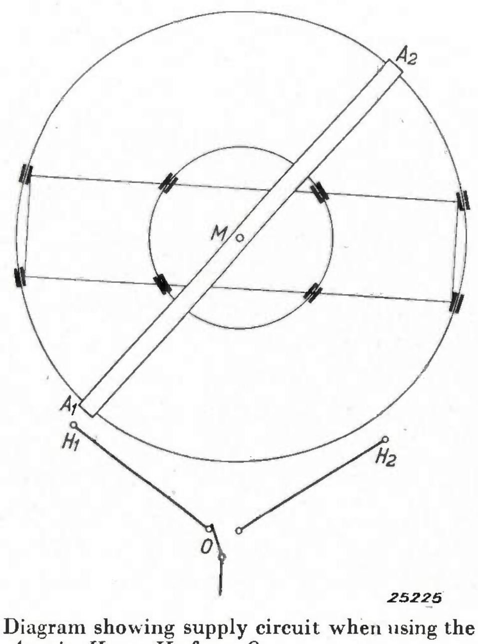


Fig. 9. Diagram showing supply circuit when using the supply arm A_1A_2, *via* H_2 or H_2 from O.

for which purpose two hand capstans are fastened underneath the framework. From each capstan a steel cable passes to a single pulley block which is fastened to a shoe on the outer rail, and from there to one of the wheel boxes. Two men at each capstan can easily set the towers in motion, while the motion can be maintained with one or two men. An electrical drive is now being designed, about which we cannot yet give any details.

The energy is supplied to the aerial vertically above the pivot by means of overhead connections. With a permanent, fixed supply connection there would be certain positions of the tower at which they would touch the supply lines or the latter would have to be conducted through the towers. By arranging the supply from two points this difficulty is avoided (*fig. 8*). In order to change from the supply point H_1 to the point H_2 the switch at *0* must be reversed and the supply line MH_1 is replaced by MH_2.

Changing the supply line is rather time-consuming, and in the long run a troublesome operation. For this reason a rotating supply arm has been designed, which can rotate independently of the towers on a pivot of its own above the framework pivot and which rests on the track. From the position drawn in *fig. 9* the towers can then be turned to the right over about 100° without the necessity of any commutation. With a greater rotation than this, the aerials must be supplied *via* H_2 by means of the other half of the rotating arm. For this purpose the switch at *0* must be reversed, the switch at H_1 opened and the rotating arm turned to the right until A_2 reaches H_2. The switch at H_2 can then be closed. *Fig. 10* gives a view of the Phohi grounds at Huizen where the PHI transmitters of the Phohi and the PCJ transmitter are set up. The new rotating directional aerial which serves for broadcasting on 31.28 m may clearly be seen at the left. In the centre background is the permanent

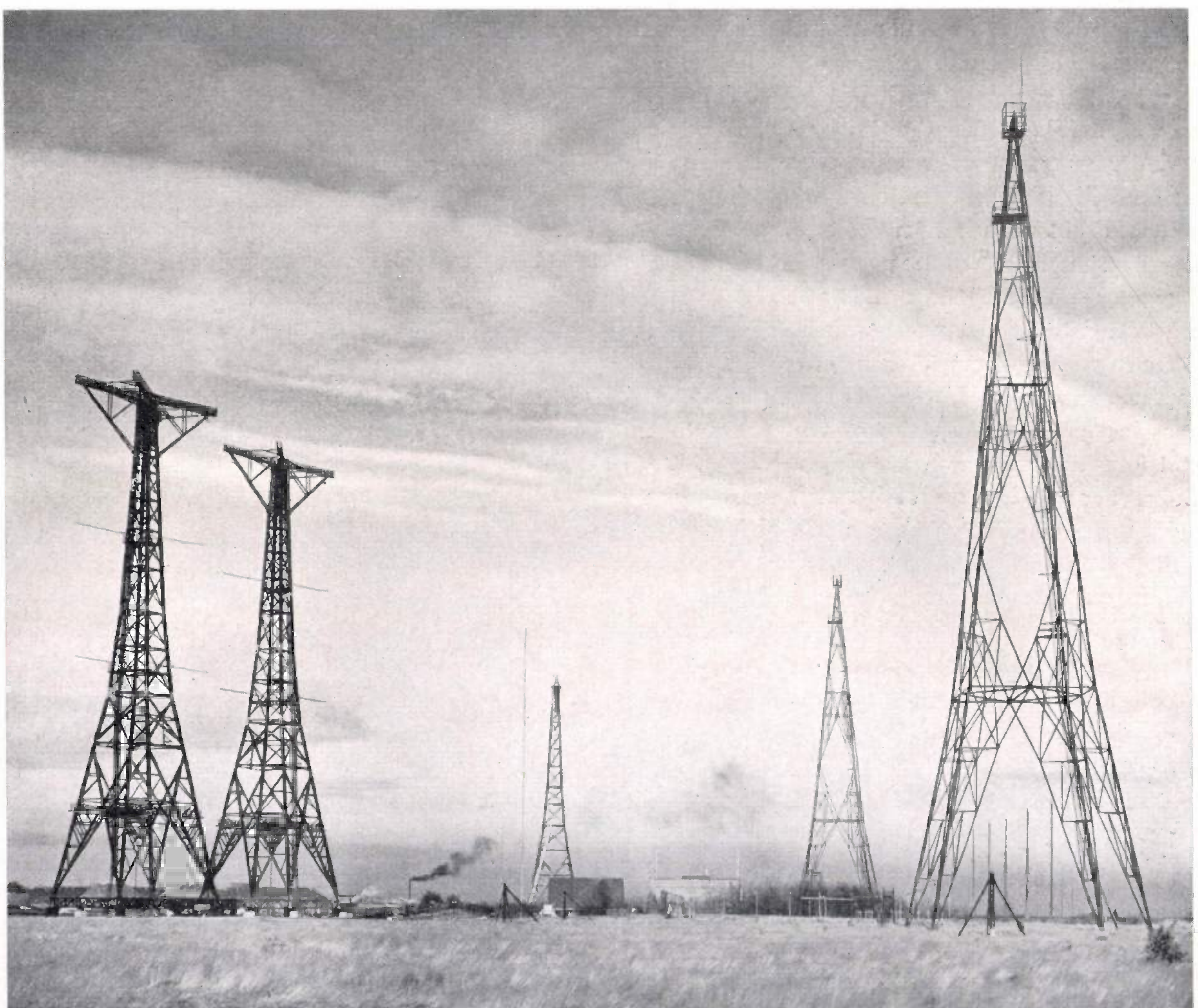


Fig. 10. General view of the grounds of the PHI transmitters of the Phohi and PCJ.

directional aerial used up to now for experimental broadcasting by PCJ on 19.71 m to the Dutch East Indies and British India. At the top of the wooden pole of this fixed directional aerial is a dipole aerial for non-directional broadcasting on either 19.71 or 31.28 m. To the right between the two high towers hangs the fixed directional aerial of the Phohi transmitter for broadcasting to the Dutch East Indies on a wave length of 16.88 m. Under the tower to the right, in the background, may be seen the fixed permanent aerial for transmission on a wave length of 25.37 m to the Dutch East Indies.

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Compare **1238** for the investigations described in this article.

1248: J. H. de Boer and E. J. W. Verwey: Semi-conductors with partially and completely filled $3d$ lattice bands (Proc. phys. Soc. **49**, extra part, 59 - 71, Aug. 1937.

Semi-conductors and insulators are discussed with incompletely filled bands of the energy levels of $3d$ electrons. If they possess a whole number of electrons per atom their lack of conductivity may be explained by the very great probability that the moving electron will be drawn back by its original atom, if only the potential barriers to be passed through are great enough to keep the probability of transition sufficiently small. If the number of electrons per atom of a definite level is not a whole number the hindrance to conduction disappears. This is the case for example in lattices of Fe_3O_4 and nickel or copper oxide when some of the places of the metal ions are unoccupied.

Substances with completely filled energy bands may become conductive as a result of irradiation with light which is capable of raising electrons into a band of higher energy. If several oxygen places in a zinc oxide lattice are empty, the excess electron in the lattice will be able to change the Zn^{++} ions into Zn^{++} ions by passing over into the energy band of the 4 s-electrons.

1249*: M. J. O. Strutt: Moderne Mehrgitter-Elektronenröhren, Erster Band: Bau, Arbeitsweise, Eigenschaften (138 pages; Julius Springer, Berlin 1937).

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*) An adequate number of reprints for the purpose of distribution is not available of those publications marked with an asterisk. Reprints of other publications may be obtained on application to the Natuurkundig Laboratorium, N.V. Philips' Gloeilampenfabrieken, Eindhoven (Holland), Kastanjelaan.

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AUDITORIUM ACOUSTICS AND REVERBERATION

by A. TH. VAN URK.

534.844

Within a closed space the sound which reaches the ear after one or more reflections at the walls has an important influence on the quality of music and the intelligibility of the spoken word. In this article this phenomenon of reflection, called reverberation, is studied by means of measurements and calculations. This study leads finally to several simple conditions which must be fulfilled if an auditorium is to have satisfactory acoustics.

Introduction

Speech and music both consist of sound vibrations which must reach the ear in a definite succession in order to be intelligible. If the source of the sound and the hearer are both in the open, the succession of vibrations is the same as that in which the sound waves were produced. In a closed space, however, it sometimes happens that vibrations which were produced earlier, after several reflections at the walls, reach the ear simultaneously with those which were sent out later. The order of the sound impressions is therefore disturbed to some extent. It is easy to understand that speech can in this way become less intelligible and music can be run together in such a way that the notes no longer sound distinct. If the source of sound is suddenly silenced, the sound of the reflected waves is still heard for a short time. This phenomenon, reverberation, is of the greatest importance in the acoustics of an auditorium, because of its influence on the intelligibility of speech and quality of music.

Why is it not possible to render the reverberation so slight that no reflection of sound takes place? This would seem to be an excellent plan in the case of speech.

In the first place it is not so easy to achieve this; moreover, if it were done, the sound from a source of a given power would be much weaker than otherwise, as can be ascertained by comparing the music of a violin in a hall with the same music out of doors. Finally there are psychological and aesthetic reasons why we desire a certain amount of reverberation in music. The rolling tones of the organ in the vaulted roof of a Gothic cathedral have a peculiar beauty.

We are thus compelled to look for a compromise between two effects:

a) when the reverberation time is long, the speech is loud but non-intelligible;

b) when the reverberation time is short, the sound is not sufficiently reinforced by reflections, nor is it uniformly distributed over the space.

It has been found that the optimum reverberation time is about two seconds for large halls, while for small halls, where the sound does not need so much reinforcement, one second is the best time.

Sabine's experiments

In the middle of the last century the importance of reverberation to the acoustics of an auditorium was already recognized. It was also known that reverberation was dependent on the volume and shape of the hall, but the nature of this dependence was not known. An exact copy of a hall with good acoustics, but on a larger or smaller scale, was sometimes found to give unexpectedly bad results. Moreover, such knowledge as there was had not penetrated sufficiently well into the sphere of architecture. Architects believed in a certain aesthetic relation between the dimensions of a hall, which was supposed to produce good acoustics. It was not until 1900 that a quantitative insight was obtained into the phenomenon of reverberation through the experiments of the American physicist Sabine, and it was still about twenty-five years before the principles so discovered were applied in the actual building.

Sabine was commissioned to make improvements on an auditorium with bad acoustics by shortening the reverberation time. In order to be able to do this, he wanted to find a relation between the various quantities which affect the reverberation time, namely, the volume of the space, the area and the absorption of the walls. To do this, he measured the reverberation time of various halls at Harvard University with an organ pipe and a chronograph. As reverberation time he took the time elapsing between the moment of shutting off of the source of sound after a steady sound intensity has been reached and the moment when the tone of the organ pipe became inaudible. He then found that reverberation is directly proportional to the volume of the hall, and that in general the curve of the walls and their relative positions have no noticeable effect on reverberation.

By the introduction of seat cushions from a theatre auditorium, he was able gradually to reduce the reverberation time of a given space from 5.35 sec to 1.14 sec. In this way he obtained a relation between the reverberation time and the number of metres of cushions introduced. This criterion was later changed into square metres of cushions, and still later into square metres of open window, after he had found out on a quiet night that the absorption of an open window was proportional to the size of the opening, and that his cushions absorbed 80 per cent of what the same area of open window absorbed. It is obvious that all the sound incident upon an opening passes through it. The number 0.8 was called the absorption coefficient by Sabine. This number, which is less than one, indicated the fraction of 1 sq metre of open window which has the same absorption as 1 sq metre of the material under consideration. In other words, this number is the ratio between absorbed and incident sound. Sabine was now able, by means of measurements in halls with different wall coverings, to determine the corresponding absorption coefficients. Account had also to be taken of the absorption of sound by the audience, which corresponds to about 0.5 sq m of open window per person.

By combining all these measurements experimentally, he obtained the very well known relation:

$$T = 0.16 \frac{V}{\Sigma \alpha_i S_i}, \quad \cdots \cdots \quad (1)$$

in which T is the time elapsing between the shutting off of the source of sound and the fall of the sound intensity to inaudibility; V is the volume of the space, S_i the area of a part of the wall and α_i the corresponding absorption coefficient. $\Sigma \alpha_i S_i$ is then taken over the total surface of the walls and other absorbers (such as the audience), each with its own absorption coefficient.

Considering the complexity of the phenomenon of reverberation, Sabine's formula is extraordinarily simple. It is striking that reverberation is independent of the shape of the space and of the nature of the distribution of the absorbing surfaces over the walls, and of the position of the source of sound and of the observer. It is this very simplicity which makes Sabine's formula so useful for calculating the reverberation time of a hall from its plans and from the known absorption coefficients of the material of the walls, or for the determination of absorption coefficients from reverberation measurements such as Sabine himself carried out in the above-described experiments with chair cushions.

From equation (1) it is immediately clear that when there is no absorption ($\alpha = 0$) the reverberation time is infinite, a fact which is clear enough even without a formula.

If the whole wall surface consisted of material which was completely absorbent, or equivalent to "open window" no reflections would take place and the reverberation time would have to be zero. This result is not confirmed by equation (1); $\Sigma \alpha_i S_i$ becomes equal to the area of the surface of the wall and not infinitely great, so that T cannot become arbitrarily small. From this it may be seen that Sabine's formula is subject to limitations.

Later in this article we shall attempt to derive Sabine's formula from the elements of the theory of sound. It will then become clear that equation (1) holds accurately only when the absorption of sound is so small that each wave is subjected to a large number of reflections before it dies out. When this is not the case, reverberation is dependent on the shape of the hall and on the distribution of the absorbing materials. It is, however, possible to improve upon Sabine's formula in such a way that it remains valid for an approximately cubical shape of hall and a uniform distribution of the absorbing material even with strongly absorbing walls.

Measurements of reverberation

Before we consider the reverberation phenomenon theoretically, we shall first give some details about the measurement of reverberation time.

Sabine carried out his measurements with an organ pipe and a chronograph. Such subjective measurements, which are sometimes still done, require a great deal of practice. Many measurements must be carried out, and their average taken, in order to obtain anywhere near a reliable result.

The accuracy of the measurements can be much improved if there is available a gramophone, an amplifier with loudspeaker, a potentiometer calibrated in decibels and an electric clock. A record with a continuous so-called warble tone is put on the gramophone. This tone is one whose frequency varies rapidly over a short range. It is used to prevent the occurrence of standing vibrations in the measuring space, which would be certain to occur if a single frequency were used, and which would adversely effect the measurement.

The circuit can be arranged in such a way, that the clock only runs when the loud-speaker is shut off. The potentiometer is set 60 db above the limit of audibility which was determined previously and the loudspeaker is switched on. One now waits until the sound intensity has become steady, the switch is then reversed so that the loudspeaker is silenced and the clock begins to run. When the limit of audibility is reached, the switch is again reversed, the loudspeaker begins again, and the clock stops. This can be repeated as often as desired, and the clock adds up the reverberation times measured. It is only necessary to divide by the number of measurements to obtain the average reverberation time. With some practice an accuracy of 5 to 10 per cent can be attained by this method.

Instead of setting the potentiometer at 60 db it may also be set at 50, 40, etc. A linear relation must then be found between these numbers and the time which elapses until the sound is inaudible. The slope of the straight line which most nearly connects these points gives the reverberation time.

For short reverberation times these subjective methods give poor results. In order to exclude the human element, an entirely automatic apparatus has been constructed for measuring reverberation time. Wide use is being made nowadays of recording meters with a logarithmic scale. They have the advantage of recording not only the reverberation time but the whole course of the reverberation, which makes it possible to determine the reverberation time with more judgment. Moreover, certain conclusions may be drawn from the course of the reverberation about the acoustic properties of the space measured. This may be illustrated by the three curves recorded with a logarithmic voltmeter. The first curve (*fig. 1*) is a normal reverberation curve, which, aside from numerous small fluctuations, corresponds well with a straight line. The reverberation time is 1.75 sec. The second

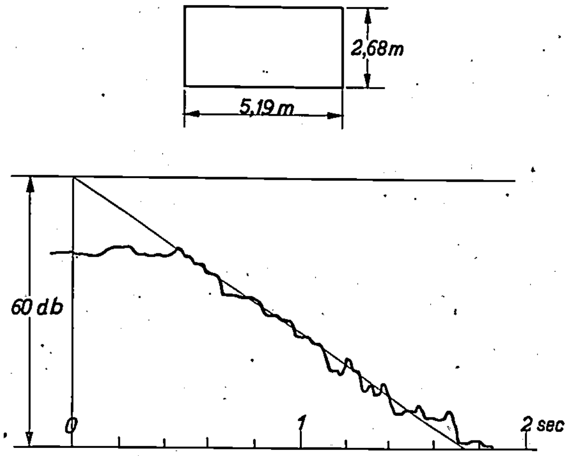


Fig. 1. Recording of the sound intensity in a room as a function of the time. The reverberation time is 1.75 sec. Above, a vertical cross section of the room.

curve (*fig. 2*) was recorded after damping material was introduced into the room on the floor and along the walls, but only extending half way to the ceiling. We can no longer speak of a single reverberation time. The curve consists of two overlapping parts. The first has a reverberation time of 0.6 sec, the second of 1.75 sec; the latter corresponds to the upper half of the room. The reverberation time, which according to equation (1) is determined by the ratio V/A for equal absorption coefficient, is the same as in fig. 1 because V/A has the same value for this half as in the first recording for the whole room. The short reverberation time is that of the damped half. It may

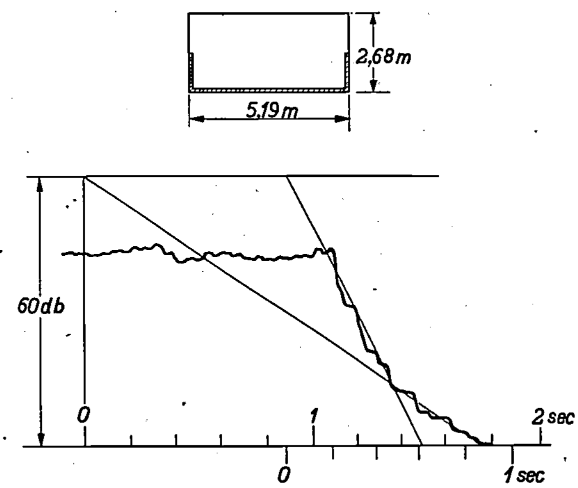


Fig. 2. Recording of the sound intensity in the same room while the lower half is covered with absorbing material, as indicated in the cross section. In the recorded curve two reverberation times may now be seen belonging to the two halves of the room.

therefore be seen from this record that when the damping material is so unequally distributed as in this case, reverberation does not follow a simple exponential course, but consists of two connected exponential branches. In the figures these are straight lines, since the intensity scale is exponential.

In the third recording the door of the room was

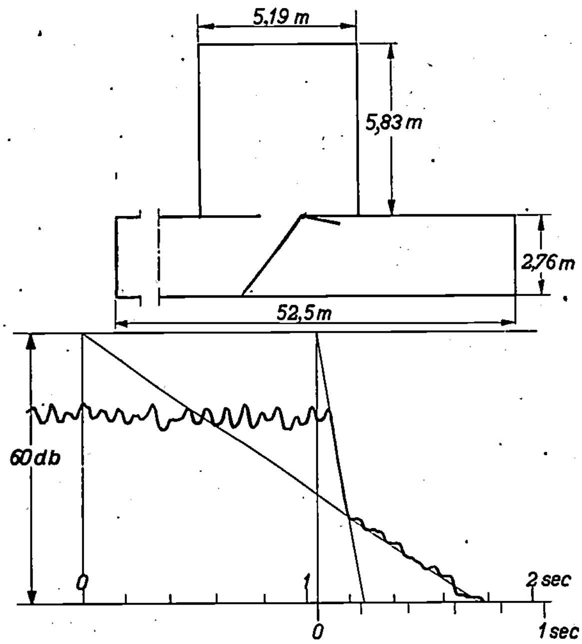


Fig. 3. Recording of the sound intensity in the same room with the door to the corridor open, according to the plan drawn above. In the recorded curve may be seen the short reverberation time of the room itself as well as the long reverberation time of the corridor. The impression is actually given that the last reverberation enters the room from the corridor.

opened wide and care was taken that the sound passing through the door did not immediately re-enter the room. This was done by means of a screen placed diagonally outside the door. The reverberation time of the lower portion of the room has now become shorter, namely 0.2 sec, because the absorption of the open door is now added to that of the material introduced; the other reverberation time has become somewhat longer because the sound which enters the corridor, after it has passed from one end to the other of the corridor, re-enters the room and again passes out through the door. This occurs several times until the sound has died out. The transition between the two parts of the reverberation curve is much sharper here. The reverberation time of the corridor is also expressed in this curve, and is somewhat longer than that of the room. The recorded reverberation times are obtained by picking up the sound with a microphone, whose alternating voltage, after amplification, is traced on the above-mentioned recording meter on a film 35 mm wide, which is made especially for mechanical tracing. In these recordings the speed of the film was 25 mm per sec, and the tracing speed 600 db per sec, that is, the maximum speed at which the needle moves is such that it would be able to cover 30 cm in the direction of the intensity scale in one second. Since on this scale 60 db correspond to 3 cm, we call this 600 db. The principle of this recording arrangement is based in general on an amplifier with potentiometer, which is operated mechanically in such a way that the output voltage of the amplifier remains constant. Thus, if the input voltage falls or rises, the potentiometer is mechanically adjusted so that the output voltage always remains the same. The mechanical control arrangement at the same time operates the tracing needle and provides for sufficient force and speed for tracing.

Photographic copies of these records can be made in unlimited mumbers.

A few years ago an important new method was developed for measuring the reverberation in concert halls filled with people, without its being necessary to take other measures than those which are necessary for broadcasting such music over the radio. In the score of a piece of music to be broadcasted, portions can be picked out where the music suddenly stops at a sufficiently high level of sound and is followed by a sufficiently long rest. A radio set may now be connected in front of the recording apparatus and the recording begun just before beginning of the rest. During the rest the dying out of the sound is recorded and thus the average reverberation time under the conditions under which the hall is normally used.

If it is desired to measure the reverberation time as a function of the frequency in this way, the output voltage of the radio set must be filtered, so that only a narrow frequency band enters the recording apparatus. By using successive octave filters (filters which pass only one octave) the whole course of the reverberation time as a function of the frequency can be recorded. It is then found that the reverberation time is practically never entirely independent of the frequency, and usually becomes greater towards the low frequencies.

Theoretical treatment of reverberation

Considering the importance of Sabine's formula, it is not surprising that there have been repeated attempts to derive this formula theoretically. The value of these attempts lies in the fact that each time a better insight is obtained into the limits of the validity of Sabine's law. We shall give here a simple deriviation of Sabine's law for an idealized case.

Let us assume that we are concerned with a space enclosed by hard walls which reflect sound totally. Let us now allow a source of sound to act for a short time and then stop. After some time *it is probable that the space is uniformly filled with sound due to repeated reflections,* at least if the shape of the space is not of such a regular form that the sound continues to travel in certain paths.

In order to understand this we shall find out with the help of *fig. 4,* how the filling of the space with

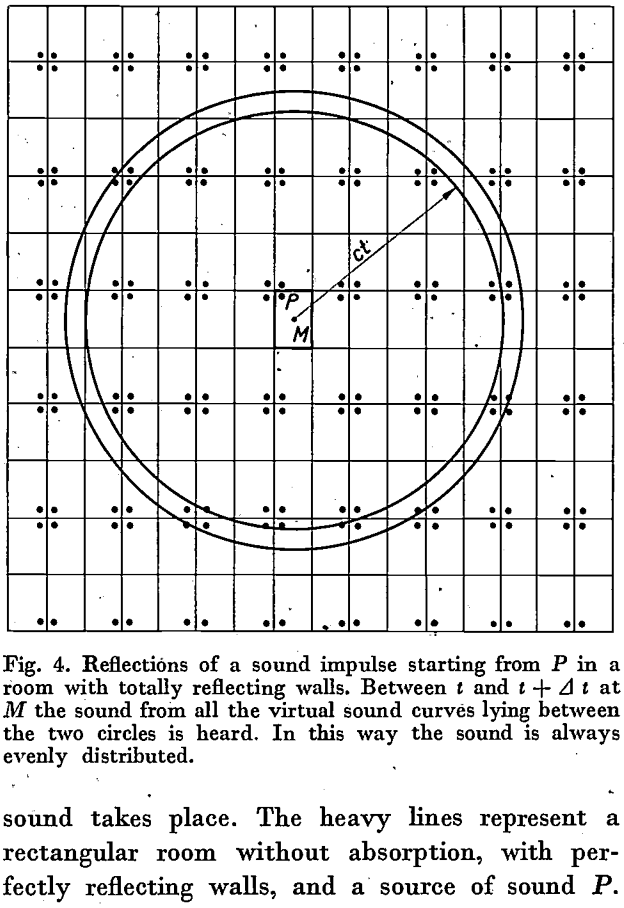


Fig. 4. Reflections of a sound impulse starting from P in a room with totally reflecting walls. Between t and $t + \Delta t$ at M the sound from all the virtual sound curves lying between the two circles is heard. In this way the sound is always evenly distributed.

sound takes place. The heavy lines represent a rectangular room without absorption, with perfectly reflecting walls, and a source of sound P. We wish to know the intensity of sound at an arbitrary point M of the room between the moments t and $t + \Delta t$. The direct sound comes from P; the sound once or oftener reflected comes from the mirror images of P, which are indicated in the figure by small dots. If, for example, at the moment $t = 0$ a sound impulse (a pistol shot) left P, then between t and $t + \Delta t$, the impulses of the virtual sources of sound, which lie in the spherical shell between ct and $c\ (t + \Delta t)$ are heard at M. The volume of the shell between the spheres ct and $c\ (t + \Delta t)$ increases proportionally with the square of the time, and this holds also for the average number of virtual sources of sound in the shell. As may be seen from the figure, the virtual sources of sound form a regular pattern in space, so that they will on the average be distributed uniformly over the shell. There are naturally always accidental fluctuations about the most probable value, not only in the number of virtual sources of sound, but also in their distribution over the sphere. These fluctuations, however, become continually smaller as the number becomes greater. The average number of virtual sources of sound, and therefore the average intensity, is independent of the position of M in the space, and the sound waves travel on the average with equal intensity in all directions.

The space is now filled with an energy VE_0, when V is the volume and E_0 the uniform energy per cc. Let us now make a hole in the wall with an opening A. All the energy will finally pass out through this opening. The energy passing through 1 sq.cm of the opening per second is $cE/4$ (see small print for proof). The energy incident on the whole opening is $cEA/4$. This is therefore equal to the decrease in total energy per second, since all the energy falling upon the opening is lost. Therefore

$$\frac{\mathrm{d}\,(VE)}{\mathrm{d}t} = -\frac{cEA}{4} \quad \ldots \quad (2)$$

The solution of this is

$$E = E_0\, e^{-\frac{cA}{4V}t} = E_0\, 10^{-0.434\,\frac{cA}{4V}t}.$$

Now in order to find a definite reverberation time T we assume that the reverberation becomes inaudible when the energy density of the sound has become a factor f smaller. Therefore

$$\left.\begin{aligned} &10^{0.434\,\frac{cA}{4V}\cdot T} = f\,, \\ &T = \frac{4V}{cA}\,\frac{\lg f}{0.434} = 0.0268\,\frac{V}{A}\,\lg f, \\ &\text{while Sabine had found} \\ &T = 0.16\,\frac{V}{A} \end{aligned}\right\} \quad . \quad . \quad (3)$$

It may thus be seen that Sabine's experiments agree with the assumption that the reverberation becomes inaudible when the density of the energy has decreased by a certain factor. From the two equations (3) it follows approximately that

$$\lg f = 6; \qquad f = 1.000\,000.$$

The density of the energy therefore decreases to one millionth of its original value before it is inaudible, thus it decreases by 60 decibels. The sound pressure varies proportionately to the square root of the density of the energy, and therefore decreases during the reverberation to one thousandth of its original value.

The reverberation is of course heard somewhat longer by one observer than by another. For physical measurements, however, the time corresponding to a decrease in intensity of 60 db is generally taken as the reverberation time.

Derivation of the formula $cE/4$ for the energy flux through the unit of area.

In a volume dV (*fig. 5*) there is an amount of energy $E dV$. The portion of this which falls upon the surface dS lies within

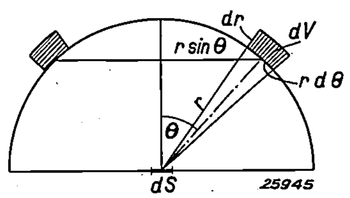


Fig. 5. Derivation of the energy flux through the unit area.

the solid angle $d\Omega$ within which the surface dS is seen from dV. This solid angle is given by

$$d\Omega = \frac{\cos\Theta\, dS}{r^2},$$

when r is the distance between dV and dS, and Θ the angle between the normal to dS and the line joining dV and dS. We begin with the assumption that the energy flows from dV uniformly in all directions. Of the energy $E dV$, which is in dV at a given moment, the amount $E dV \frac{\cos\Theta\, dS}{4\pi r^2}$ will fall on the surface dS.

All the volume elements from which a line to dS makes an angle Θ with the normal to dS, now lie in a torus whose diameter is $r d\Theta \cdot dr$, and the large circle has a circumference of $2\pi r \sin\Theta$. Therefore, $dV = 2\pi r \sin\Theta \cdot r d\Theta \cdot dr$. The energy from dV which falls on dS is thus

$$E \cdot \frac{\cos\Theta\, dS}{4\pi r^2} \cdot 2\pi r^2 \sin\Theta\, d\Theta\, dr = \frac{E}{2} \cdot \sin\Theta\, \cos\Theta\, d\Theta\, dr\, dS.$$

The total energy received by dS in the time dt comes from a hemisphere with radius cdt, when c is the velocity of sound

$$E_{tot} = \frac{E}{2}\, dS \int_0^{cdt} dr \int_0^{\pi/2} \sin\Theta \cos\Theta\, d\Theta = \frac{Ec}{4} \cdot dS\, dt.$$

An amount of energy equal to $\frac{Ec}{4}$ thus falls upon 1 sq cm per second.

Sound amplification by reverberation

As already mentioned, a certain degree of reverberation is desirable because the intensity of the sound with a given energy of the source is thereby increased. In order to investigate this amplification we make use of the total absorbed energy calculated in the foregoing

$$\frac{cEA}{4} = \frac{cE}{4}\, \Sigma\, a_i\, S_i.$$

This energy in the state of equilibrium is equal to the energy P radiated from the source of sound. The density of the energy in the space, which determines the intensity of sound, is, therefore, at equilibrium

$$E = \frac{4P}{c\,\Sigma\, a_i S_i}.$$

Now, according to Sabine's formula

$$\frac{1}{\Sigma a_i S_i} = \frac{T}{0.16 V},$$

so that the density of energy

$$E = \frac{4}{0.16\, c} \cdot \frac{P}{V}\, T = 0.072\, \frac{P}{V}\, T.$$

The intensity of sound is therefore proportional to the energy per unit volume and to the time of reverberation. For very short reverberation times this formula is no more valid than Sabine's formula; the sound intensity does not become zero, but passes over to the intensity of the direct sound.

Deviations from Sabine's formula.

In the foregoing derivation the essential assumption was that the space was uniformly filled with sound. Actually this means that the absorption must be so small that a large number of reflections are necessary to cause the sound to die out. When this is not so, for example, when all the sound is absorbed so that no reflection at all takes place, deviations from Sabine's formula occur. This was shown in the discussion of Sabine's experiments.

In order to set up a refined theory of absorption we shall follow a ray of sound along its path. This ray of sound will cover paths of different lengths, $l_1, l_2 \cdots l_n$, between each two successive reflections at a wall, and sound will be absorbed at each point of reflection. For ease in calculation we shall use a mean value of l, the so-called mean free path of the sound. The calculation can best be followed by means of an example. Suppose we have a room with a free path of 20 metres and an absorption coefficient on all walls of 0.50. The sound covers a distance of 340 m per second, and $340/20 = 17$ reflections take place. At each reflection 0.50 of the original sound intensity remains, and therefore after n reflections 0.50^n. To become inaudible this intensity must fall to 0.000 001. From this we can determine n: $0.50^n = 10^{-6}$, $n = 20$. The total number of reflections is thus 20, and there are 17 reflections per sec. The reverberation time is therefore $20/17 = 1.18$ sec.

The mean free path l used in the above theory is dependent in the first place on the ratio V/A of the volume of the hall to the reflecting surface. If the number of reflections is sufficiently great the free path, independent of the shape of the room, becomes

$$l = \frac{4V}{A} \text{ (see small print for proof)} \quad \cdot \cdot \quad (4)$$

Therefore in the above example $4V/A = 20$ metres. Now that this ratio is known, we can also apply Sabine's formula to the example in question and we find

$$T = 0.16\,\frac{V}{\Sigma a_i S_i} = 0.16\,\frac{V}{0.5\,A} = 0.08 \cdot \frac{4V}{A} = $$
$$= 0.08 \cdot 20 = 1.6 \text{ sec.}$$

This differs considerably from the value of 1.18 sec calculated above.

It is therefore clear that the treatment given above differs from the calculation according to Sabine's formula. If we now repeat the expression in general terms, we must obtain a different formula for the reverberation. The reverberation time is now the total number of reflections N divided by the number of reflections per sec (n), and the latter number is the velocity of sound divided by the mean free path, thus

$$n = \frac{c}{4V/A} = \frac{cA}{4V}.$$

The total number of reflections may be found from $(1-\alpha)^N = 10^{-6}$, where α is the absorption coefficient, thus:

$$N = \frac{-6}{\lg\,(1-\alpha)} = \frac{-6\,\ln\,10}{\ln\,(1-\alpha)}.$$

Combining we obtain

$$T = \frac{N}{n} = \frac{N \cdot 4\,V}{cA} = \frac{24\ln 10}{c} \cdot \frac{V}{-A\ln\,(1-\alpha)} =$$
$$= 0.16\,\frac{V}{-A\ln\,(1-\alpha)} \quad \cdot \cdot \cdot \cdot \quad (5)$$

For small values of α this formula passes over into the original Sabine formula, namely

$$T = 0.16\,\frac{V}{\alpha A}.$$

In this derivation we have taken a uniform value of the absorption coefficient. It is clear that if this coefficient varies slightly over the walls, instead of $-A\ln\,(1-\alpha)$ we must take in the denominator $\Sigma - S_i \ln\,(1-\alpha_i)$.

The practical importance of equation (5) is that it shows that less absorbing material is necessary to obtain a definite (short) reverberation time than was to be expected according to the old formula, because at a definite value of α

$$-\ln\,(1-\alpha) = \frac{\alpha}{1} + \frac{\alpha^2}{2} + \frac{\alpha^3}{3} \ldots . > \alpha,$$

and therefore the denominator has the same value for smaller values of S_i.

Derivation of the formula for the mean free path

Let us first derive the formula for the special case of a sphere with the source of sound B at the wall and with total reflection (*fig. 6*). Each sound ray which leaves B at an angle α

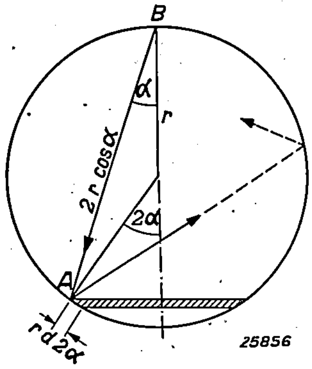


Fig. 6. Derivation of the mean free path.

with the diameter, will after reflection at a point A again make an angle α with the diameter through A. This will be repeated. Each ray keeps its original length of path. We may, therefore, take the average of all chords, starting from B, and will then obtain the correct answer. We now assume that the source of sound radiates according to Lambert's cosine law. In this case the amount of direct sound per unit of area is the same at all parts of the spherical wall. (This recalls the sphere of Ulbricht in which for the same reason every element of the surface is equally intensely lighted by the light which is radiated from any arbitrary element of the surface).

The number of sound rays from B which make an angle α with the diameter is equal to the ring-shaped surface $2\,\pi\,r \sin 2\,\alpha \cdot r\mathrm{d}\,(2\,\alpha)$. If we multiply this by the lengths of the chords $2r\cos\alpha$, we obtain $16\,\pi\,r^3 \sin\alpha\cos^2\alpha\,\mathrm{d}\alpha$. If we now let α vary from 0 to $\pi/2$ and divide by $4\,\pi\,r^2$, we have the average length of path of the sound rays from B. This is

$$l = \int_0^{\pi/2} \frac{16\,\pi\,r^3 \sin\alpha\cos^2\alpha\,\mathrm{d}\alpha}{4\,\pi\,r^2} = \frac{16/3\,\pi\,r^3}{4\,\pi\,r^2} = \frac{4\,V}{A} \quad \cdot \cdot \quad (4)$$

We can also derive formula (4) in a general way from a consideration of the kinetic gas theory. The derivation which follows is, however, only valid when we may assume that the space is uniformly filled with sound. We have previously derived that the energy falling upon 1 sq cm per second is $Ec/4$ in this case. For a gas in which n is the number of particles per cc the same derivation gives $nc/4$ for the number of particles which strike 1 sq cm per second. The number of particles which strike a surface A per second is then $ncA/4$. Let us consider a hypothetical surface which surrounds a volume V in the gas. Assume that the particles move in straight lines and do not collide with each other, and that the average path to the surface of each particle within the volume is l and its velocity c, then each particle remains l/c seconds in the volume V. If we now multiply the number of particles which strike the wall per second with this average time of remaining in the volume, it must give just the total number of particles in the volume. We obtain

$$\frac{n\,cA}{4}\cdot\frac{l}{c} = Vn \quad \text{of} \quad l = \frac{4\,V}{A}.$$

With a sufficient number of reflections, equation (4) is quite generally valid for the mean free path. If, however, the shape of the room or the position of the source of sound is not a very special one, so that the mean distance of the source of sound from the wall corresponds approximately with l, we shall also be able to use equation (4) for a small number of reflections, as was shown above for the spherical space.

Absorption by moisture of the air

In the foregoing we have spoken only of the absorption at the walls. For the higher frequencies there is also a volume absorption, that is, at very high frequencies energy is also absorbed from the sound waves by the air molecules themselves. Within the last few years it has been found that this absorption in the air is very dependent on the moisture in the air.

The influence of the water vapour seems to be based on the fact that it plays the part of a catalyst for the distribution of the energy over the possibilities of motion of the molecules. The absorption reaches a maximum at a definite small percentage of water vapour, which is dependent on the frequency. Above and below this percentage of water vapour the absorption decreases suddenly. If we indicate the absorption coefficient per metre for plane waves by m, a term $4\,mV$ occurs in the denominator of the formula for the reverberation time. This term is connected with the volume absorption. A derivation similar to that of equation (5) then gives

$$T = \frac{0.16\,V}{4\,mV - S\ln(1-a)}\,.$$

m is here a function of frequency and humidity. For frequencies below 1000 c/s m is so small that it may be neglected at all degrees of humidity, but above 5000 c/s the first term of the denominator may even become greater than the second. At frequencies above the limit of audibility the second term may be neglected in comparison to the first. For these frequencies the reverberation time in a reasonably large space becomes independent of the volume and the surface area of the space.

Reverberation time and intelligibility

From the preceding considerations we have seen that the sound intensity decreases exponentially after the source ceases to act. It is now plausible that the growth in intensity of sound from the moment the source begins to act is also exponential. We shall make use of this in order to explain the relation mentioned in the introduction between reverberation time and intelligibility.

Fig. 7 gives the growth and fading out of sound

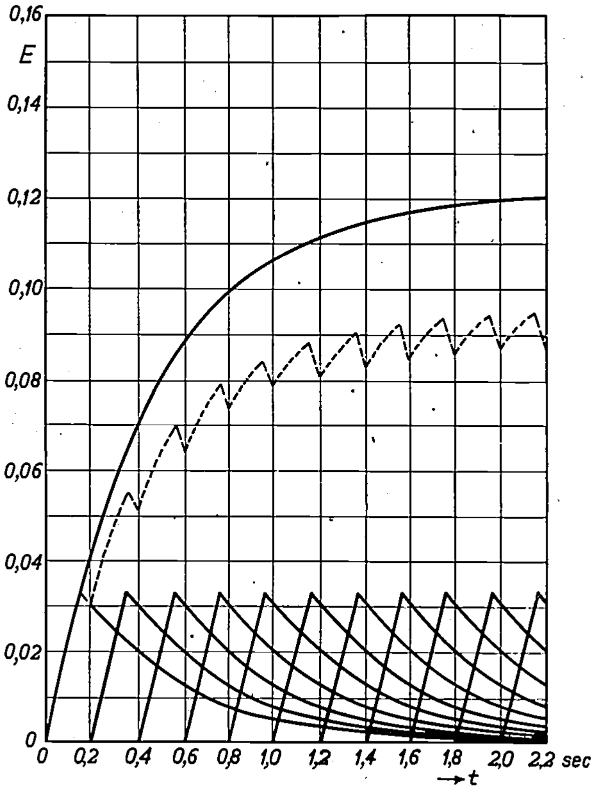


Fig. 7. Growth and fading of the sound in a space with a reverberation time of 6 sec. The sound consists of syllables which have a constant intensity for 0.15 sec and which are separated by intervals of silence of 0.05 sec. The intensity of the syllables projects only slightly above the average sound. The intelligibility is therefore poor.

in a space with a reverberation time of 6 sec. The sound consists of spoken syllables, each of which lasts 0.15 second and which are separated by a time interval of 0.05 second. In the interval between syllables it is quiet. In the lower part of the graph we see the intensity time curves for the separate syllables. Each of these curves consists of two pieces of an exponential curve. The dotted curve is the integrated intensity, or better, the energy density proportional to it. The continuous upper curve represents the intensity which is obtained by continuous radiation of sound. From the dotted curve it may clearly be seen that the individual syllables project only slightly above the average level of sound. It is therefore plain that the sound will be intense but not very intelligible. Let us now reduce the reverberation time of this space to 1 sec. by increasing the absorption (*fig. 8*). The intensity does not increase so rapidly, and there is sufficient fluctuation to make the syllables stand out clearly. *Fig. 9* gives the extreme case where there is practically no reverberation. The maximum intensity is that of the individual syllables. In this last case the speech is easily understood but the intensity is weak. We say of such a space that it is dead, and consider it in general unfavourable.

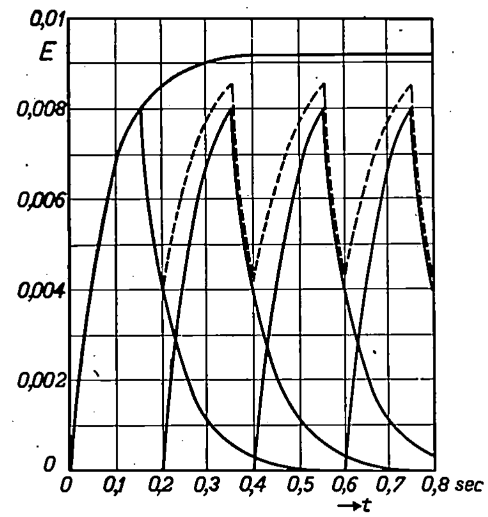


Fig. 8. Growth and fading of the sound in a room with a reverberation time of 1 sec. The individual syllables are well separated from each other. The intelligibility is good.

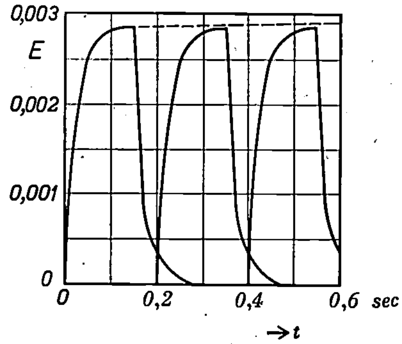


Fig. 9. Growth and fading of sound in a room with a reverberation time of 0.3 sec. The syllables are exaggeratedly separate from each other.

Final considerations

We shall briefly review the results given in the article.

The most important quantity for the acoustics of a hall is the reverberation time which must be 1 to 2 seconds according to the dimensions of the hall. This value of the reverberation time can be ensured in the design stage by providing a suitable relation between the volume of the hall and the area of the absorbing surface. The reverberation can be calculated by means of formula (1) or (5).

If a hall is very long or has an irregular shape, the calculation of the reverberation by formula (1) or (5) may give very inaccurate results. In such cases one is often concerned experimentally not with one definite reverberation time, but, for example, with two, as the recordings reproduced in figs. 2 and 3 show. The sound first decreases rapidly; the final fading out on a lower level then takes place more slowly. We have not gone deeply into such phenomena since they will not occur in a hall with good acoustics. For good acoustics it is necessary that the sound intensity be about equal everywhere in the room, that is, that the sound be distributed evenly throughout the whole space. This was, however, just the assumption which had to be made in order to derive Sabine's formula. If the acoustics of a hall are good, Sabine's formula is therefore also satisfied.

BARRETTERS

by J. G. W. MULDER.

621.316.72

In this article a study is made of the way in which the current regulating function of a regulator barretter is connected with the variation of the electrical resistance and the heat dissipation of a filament as functions of the temperature. The most satisfactory regulation curves can be obtained with an iron filament in an atmosphere of hydrogen at low pressure. Different forms of construction of Philips barretters are discussed. Finally several points are discussed which are important in the application of barretters to electrical installations. For the sake of illustration, the decrease in the current fluctuations in a tungsten lamp, a constant resistance and a carbon filament lamp are studied in connection with the loss of voltage in the preceding barretter.

Introduction

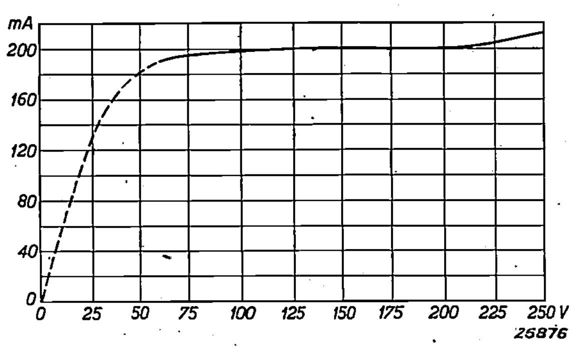
Barretters have been used in electrical engineering for many decades for the purpose of decreasing the current fluctuations which appear in electric circuits as a result of variations in the mains voltage or of changes in the load applied. An urgent need of such a current regulator was first experienced about 1900 when Nernst lamps were connected to power mains. The incandescent element of the Nernst lamp (thorium oxide) has a strongly negative temperature coefficient of the electrical resistance. If as a result of a small increase in voltage the current $I = V/R$ becomes somewhat larger, the resistance R decreases because of the corresponding increase of temperature, so that the current and the temperature continue to rise. If there were no resistance at all in series with the lamp, the current would continue to grow because of this instability until the element fused.

In order to avoid this unstable behaviour a resistance is connected in series with the Nernst lamp, which resistance varies with the temperature in exactly the opposite way to that in which the resistance of the element of the Nernst lamp varies. As early as 1899 it was found that an iron filament in an atmosphere of hydrogen was very suitable for this purpose. Iron was used because it has the highest positive temperature coefficient of the resistance, while hydrogen was originally chosen to protect the iron from oxidation. It was found later also that the heat conduction of the gaseous hydrogen is of importance.

As an illustration of the action of a current regulator the current voltage characteristics of two different barretters are shown in *fig. 1*. In example *a*), which may be considered a normal case, the current remains practically constant with a threefold increase of the voltage. In example *b*) there is even a falling back of the characteristic over a definite voltage interval, thus a decrease in the current with increasing voltage, and at the end of the regulation range the voltage is five times as great as in the beginning of that range.

In the course of years the Nernst lamp has been displaced by better sources of light. Barretters on the above-described principle have, however, continued to be used, and are finding continually wider application for the most divergent purposes. The iron filament lamp, as has already been mentioned in this periodical, is used as the series resistance of gas discharge lamps, where it fulfils exactly the same function as in the Nernst lamp. A barretter is, however, some-

a)



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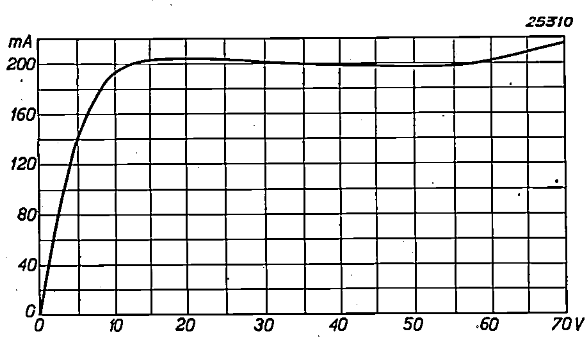


Fig. 1. Current through two barretters as a function of the voltage.
a) Normal type.
b) Type with exceptionally wide regulation range.

times also connected in series with ordinary incandescent lamps, where it is necessary to provide for large fluctuations of the mains voltage as on railways. There are, moreover, other applications in electrical engineering, such as for maintaining the constancy of the heating current of radio valves and of the charging current of batteries.

How is a flat current-voltage characteristic obtained?

If we assume that the filament has the same temperature over its whole length, we can easily calculate the relation between I and V, when the energy loss (radiation plus heat conduction plus convection) A, and the electrical resistance R as functions of the temperature T are known. We have the relations

$$\left.\begin{aligned} V \cdot I &= A(T) \\ V/I &= R(T) \end{aligned}\right\} \quad \ldots\ldots\ldots\ldots \quad (1)$$

A depends upon the nature and pressure of the gas and upon the external temperature; R is a property of the material of the wire.

The first equation expresses the fact that the state is stationary; the energy supplied is equal to that given off. The second is simply Ohm's law.

From these two equations it follows that

$$\left.\begin{aligned} &\text{a)} \quad V = \sqrt{A \cdot R} \\ &\text{b)} \quad I = \sqrt{A/R} \end{aligned}\right\} \quad \ldots\ldots\ldots\ldots \quad (2)$$

By calculating V and I for a number of temperatures we thus obtain an I-V curve. The following calculations are based on measurements of the heat dissipation of very thin wires in hydrogen gas by Busch [1]) and on resistance measurements of very pure iron by Potter [2]).

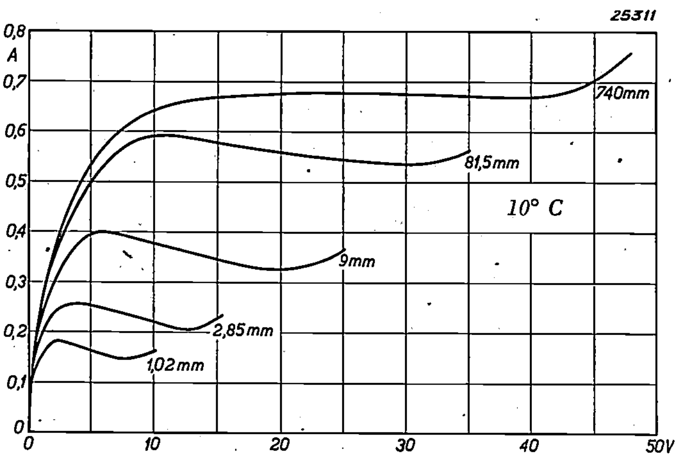


Fig. 2. Calculated I-V curves for an iron filament in hydrogen at different pressures. In a certain voltage interval there is a decrease of the current with increasing voltage.

[1]) H. Busch, Ann. Physik **64**, 401, 1925.

[2]) H. H. Potter, Proc. Phys. Soc. 49, 671, 1937.

The diagrams reproduced in *fig. 2* refer to a constant external temperature of 10° C and various pressures of the hydrogen. In *fig. 3* the characteristics are reproduced for a constant amount of hydrogen in the valve (25.8 mm Hg at 20° C) and three different external temperatures. The surprising result of these calculations is that the I-V curve for a given voltage reaches a maximum and then exhibits a falling back over a long range of voltages. The falling back of the characteristic is more pronounced the lower the external temperature.

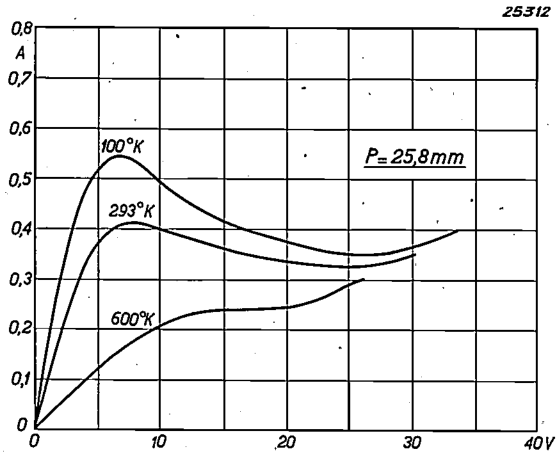


Fig. 3. Calculated I-V curves for a hydrogen pressure of 25.8 mm Hg at three different temperatures. With increasing external temperature the fall of the characteristic is less pronounced.

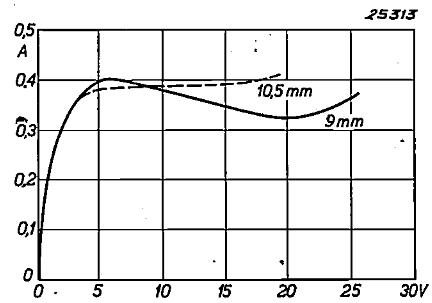


Fig. 4. Comparison of a calculated characteristic (fig. 2, 9 mm Hg) with a recorded characteristic under about the same conditions. The general shape is well confirmed. The descending portion, however, cannot be realized, because a uniform temperature is not established in this portion.

Fig. 4 shows, in addition to the curve for 9 mm pressure of *fig. 2*, an experimental curve for the same conditions. The general behaviour of the calculated curve is well confirmed; the falling back of the characteristic is, however, not present. The cause must be sought in the fact that a uniform distribution of temperature over the whole length of the wire, as was assumed in the calculation of the I-V curve, is not stable in the descending portion of the characteristic.

If we allow the temperature of a given element of the wire to rise slightly, not only the heat dissipation increases, but also the resistance and the supply of heat. In the descending portion of the I-V curve the supply of heat is found to increase more rapidly than the dissipation of heat, so that the temperature of the element continues to increase while the rest of the wire cools off when the voltage on the wire is kept constant.

The shape of the I-V curve is very sensitive to accidental factors. The parts of the wire which will become hotter, and how much hotter they will become, and the parts which will become cooler depend upon very slight irregularities in the state of the surface of the wire. The heat conduction along the wire is also important in this connection since it works to decrease the irregularity in the distribution of temperature along the wire. We shall not go more deeply into this very complicated problem.

What kinds of wire and what gases are suitable for use in barretters?

We shall now examine the conditions which must be satisfied if the current is to remain constant over a certain interval of temperature. The answer is implicit in equation (2).

By setting the first differential of the current with respect to the temperature equal to zero, we obtain:

$$\left.\begin{array}{l} R\,\dfrac{\mathrm{d}A}{\mathrm{d}T} - A\,\dfrac{\mathrm{d}R}{\mathrm{d}T} = 0; \\ \text{and therefore} \\ \dfrac{T}{A}\dfrac{\mathrm{d}A}{\mathrm{d}T} = \dfrac{T}{R}\dfrac{\mathrm{d}R}{\mathrm{d}T} \end{array}\right\} \quad . \quad . \quad . \quad . \quad (3)$$

In order to make the significance of this equation somewhat clearer, we can let A and R approach a certain working point T_0 by definite powers of the temperature:

$$A = A_0 \left(\frac{T}{T_0}\right)^{\alpha},$$

$$R = R_0 \left(\frac{T}{T_0}\right)^{\varrho},$$

and therefore

$$I = \sqrt{A_0/R_0}\left(\frac{T}{T_0}\right)^{1/2\,(\alpha - \varrho)}.$$

The requirement of a constant current thus comes down to the condition that $\alpha = \varrho$.

The exponent ϱ of the resistance is approximately equal to unity for ordinary metals. α should be about 4 when the heat dissipation takes place by radiation. With heat dissipation by convection, α may be smaller, but it always remains considerably greater than one. The result is that a constant current is only possible when care is taken *that the resistance increases with an extraordinarily high power of the temperature, and the heat dissipation increases with an extraordinarily low power of the temperature.*

The metals iron and nickel, which show exceptionally high temperature coefficients of the electrical resistance somewhat above the magnetic Curie points, may in the first place be considered as materials for the filament. The values of $T/R\ \mathrm{d}R/\mathrm{d}T$ lie in a suitable temperature range between 2 and 2.5.

In contrast to the heat radiation, which increases with the fourth power of the temperature, heat conduction and convection, especially with great differences in temperature between filament and gas, show a much slower increase with temperature. By using a light gas, such as hydrogen, the conduction can be made to dominate over the radiation and the lowest possible value of $T/A\ \mathrm{d}A/\mathrm{d}T$ is thus obtained.

In order to obtain some idea of the suitability of different materials and gases, the quantities $T/R\ \mathrm{d}R/\mathrm{d}T$ (continuous lines) and $T/A\ \mathrm{d}A/\mathrm{d}T$ (dotted lines) are shown for these materials and gases in *fig. 5* as functions of the temperature.

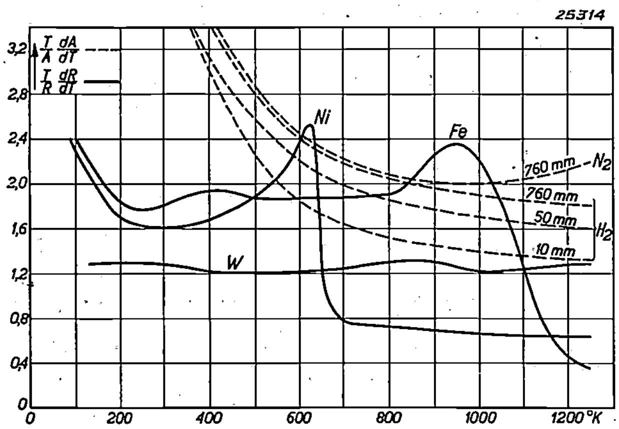


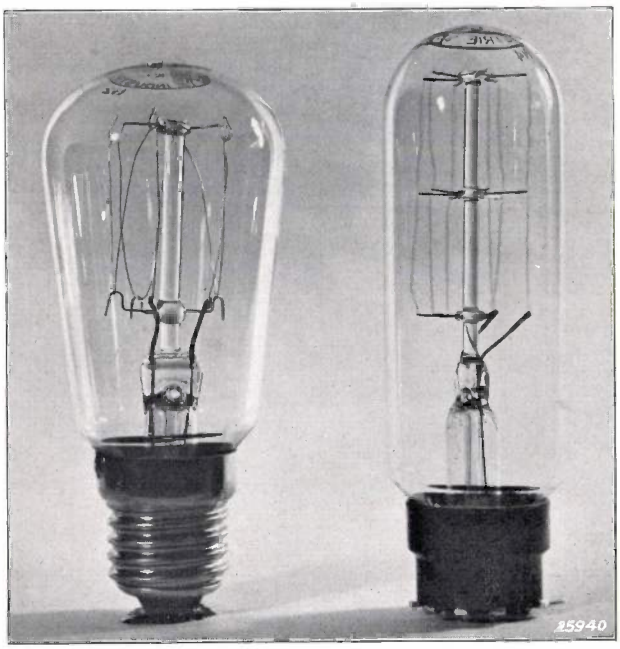
Fig. 5. Continuous lines: $T/R\ \mathrm{d}R/\mathrm{d}T$ values of different metals as a function of the temperature (R, electrical resistance, T absolute temperature). Dotted lines: $T/A\ \mathrm{d}A/\mathrm{d}T$ values of a very thin wire for different gases and pressures (A heat dissipation of a wire in this gas). In the range where a continuous line lies above a dotted one, the corresponding combination of metal and gas shows a descending characteristic. The greatest descending range is that of an iron wire in an atmosphere of hydrogen at a pressure of 10 mm.

If a continuous line and a dotted line intersect, the combination in question of filament material and gas produces a descending I-V curve. In the temperature range between 850° K and 1100° K the $T/A\ \mathrm{d}A/\mathrm{d}T$ value for nitrogen lies relatively close to the $T/R\ \mathrm{d}R/\mathrm{d}T$ value of iron. In this range therefore an iron filament in nitrogen at ordinary pressure would be suitable as a current regulator. In the case of most metals the resistance is about proportional to the temperature

($T/R\ \mathrm{d}R/\mathrm{d}T = 1$), which is too low for the production of a flat *I-V* curve. At very high temperatures the temperature coefficient of heat dissipation in hydrogen and other light gases is so low that it approaches the temperature coefficient of the electrical resistance of tungsten. It has actually been found possible to construct a barretter with tungsten in helium.

Some details of the construction of Philips' barretter

The desired favourable conditions for cooling are best realized by a straight wire. In order to avoid difficulty in construction, however, the wire is usually suspended in a zigzag form (*fig. 6a*) or in the form of a spiral (fig. 6*b*). This has practically no objectionable effect if care is taken that the surface of the bulb is not too small for the energy dissipated. If it is desired to obtain a long range of regulation the surface of the bulb must be at least several sq.cm/watt.



a) b)

Fig. 6. Barretters.
a) The filament is suspended in a zigzag form.
b) The filament has the form of a spiral.

With a given diameter of the tube and a given diameter and temperature of the filament, an increase in the heat dissipation can be attained by making the cross section of the filament and/or of the tube flattened instead of circular. This possibility is successfully utilized in regulators for heavy currents with relatively small dimensions. In such valves the heat dissipation is naturally often insufficient so that the regulation range is relatively short. *Fig. 7* gives an example of the way in which the regulation curve is improved when the circular wire is replaced by a flat strip.

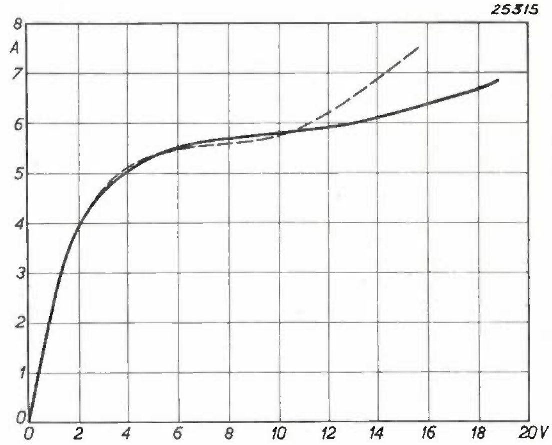


Fig. 7. *I-V* curve of a barretter, dotted line: filament with circular cross section; continuous line: flattened filament (thickness $^1/_{30}$ of breadth). The improvement in the conditions for cooling resulting from the flattening has resulted in an increase of the range of regulation.

Special measures were necessary in the construction of regulator valves for low currents. If such a valve with a filament of pure iron is connected to the alternating current mains, it is observed that the filament shrivels, as a result of the temperature fluctuations, and becomes so short that it quickly breaks. In order to prevent this happening, care must be taken that the transition point between the α-modification and the γ-modification of the iron is not passed through during the fluctuations in temperature. *Fig. 8* shows two iron filaments each of which has repeatedly been heated to red heat, that on the left just below, that on the

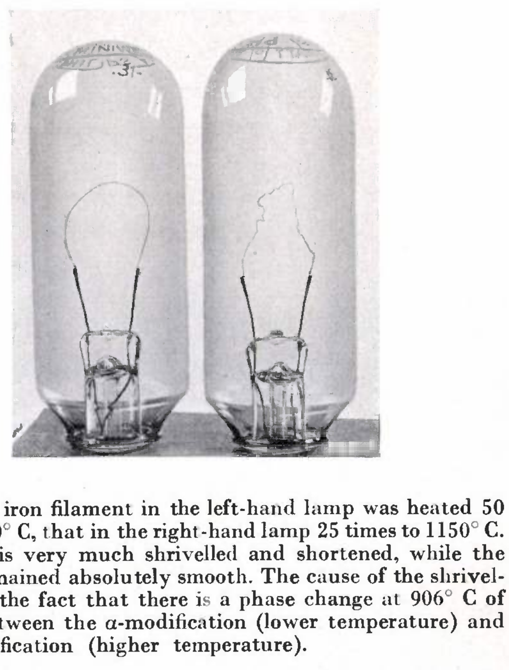


Fig. 8. The iron filament in the left-hand lamp was heated 50 times to 900° C, that in the right-hand lamp 25 times to 1150° C. The latter is very much shrivelled and shortened, while the first has remained absolutely smooth. The cause of the shrivelling lies in the fact that there is a phase change at 906° C of the iron between the α-modification (lower temperature) and the γ-modification (higher temperature).

right just above the transitionpoint (906 °C). The left-hand filament is quite smooth and bright, the right-hand one is badly shrivelled.

With filaments of a nickel-iron alloy, which has no transition phase, these difficulties are not encountered, and current regulators with low regulation current can be made for alternating current also. The relation between the maximum and the minimum voltage of the regulation range need only be slightly less favourable with these barretters than with those having pure iron filaments.

Circuits containing barretters

Since the current in the regulation range of the barretter lamp is fixed, a separate type of lamp must be used for each current. There is in principle no limitation of the current. It is, however, advisable to use several lamps in parallel for currents greater than 6A, because the heating up time of 6A barretters is several seconds, and increases approximately with the square of the current.

For different regulation ranges of the voltage a separate type of valve is also necessary. The voltage, however, need not be adjusted so accurately as the current; it is only necessary that the variations in voltage occurring fall within the regulation range of the valve. Barretters are made with voltages up to 240 volts in the middle of the regulation range. The connection of several lamps in series is not immediately possible. If the barretters do not have exactly the same current in the regulation range, then, when I is the current through both, the barretter lamp with characteristic a (*fig. 9*) will

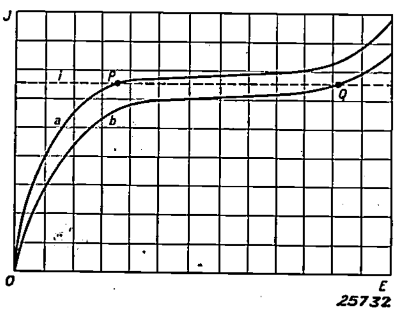


Fig. 9. Two barretter lamps of the same type, with slightly differing characteristics may have quite different working points P and Q at the same current. Because of this regulator lamps may not be connected in series.

adjust itself to the working point P, and the tube with the characteristic b to point Q. This destroys any possibility of regulation, because at the moment when lamp a assumes its regulatory function, b is already fully loaded and every further increase in the load damages barretter tube b.

Two applications of Barretters

As an example [3]) of the application of barretters let us try to decrease the current fluctuation of a tungsten lamp, a carbon filament lamp or a constant resistance on a variable mains voltage. To do this, we represent the V-I characteristic of the regulator at the working point by the equation

$$\frac{\Delta V_r}{V_r} = q\,\frac{\Delta I}{I} \quad \text{. . . . .} \quad (4)$$

For a constant resistance $q = 1$; for an absolutely flat characteristic $q = \infty$. We shall take the rather low value of $q = 10$ for the resistance tube, which is therefore not a very flat I-V characteristic. For the voltage V_{gl} on the load, for example an electric lamp, a similar equation holds

$$\frac{\Delta V_{gl}}{V_{gl}} = p\,\frac{\Delta I}{I}, \quad \text{. . . . .} \quad (5)$$

where p, for tungsten filaments for instance, is about equal to 2, and for carbon filament lamps is about 0.7. From (4) and (5) it follows that

$$\frac{1}{p}\,\frac{\Delta V_{gl}}{V_{gl}} = \frac{1}{q}\,\frac{\Delta V_r}{V_r} \quad \text{. . . . .} \quad (6)$$

Now the total variation in the voltage of the mains is

$$\Delta V = \Delta V_{gl} + \Delta V_r, \quad \text{or according to (6)}$$

$$\Delta V = \Delta V_{gl} + \frac{q}{p}\,\frac{V_r}{V_{gl}}\,\Delta V_{gl},$$

from which it follows that

$$\frac{\Delta V_{gl}}{V_{gl}} = \frac{\Delta V}{V}\,\frac{p\,V}{p\,V_{gl} + q\,V_r},$$

$$\frac{\Delta V_{gl}}{V_{gl}} = \frac{\Delta V}{V}\,\frac{1}{1 + \left(\frac{q}{p} - 1\right)\frac{V_r}{V}} \quad \text{. . .} \quad (7)$$

The relative voltage fluctuations on the lamp are thus decreased more, the greater the portion V_r/V of the mains voltage destroyed by the regulator. In *Table I* are given several numerical values for $q = 10$; $p = 2$, $p = 1$ and $p = 0.7$.

If a barretter is used which takes up half of the voltage ($V_r/V = 0.5$), the voltage fluctuations of an ordinary electric lamp, a constant resistance and a carbon filament lamp are reduced to 33, 18 and 13 per cent respectively.

[3]) N. A. Halbertsma, Elektrotechn. Z. 48, 512, 1927.

Table I.

V_r/V	$\frac{\Delta V_{gl}}{V_{gl}} : \frac{\Delta V}{V}$		
	$p = 2$	$p = 1$	$p = 0{,}7$
0.0	1.000	1.000	1.000
0.1	0.714	0.526	0.429
0.2	0.556	0.357	0.273
0.3	0.455	0.270	0.200
0.4	0.384	0.218	0.158
0.5	0.333	0.182	0.131
0.6	0.294	0.156	0.111
0.7	0.263	0.137	0.097
0.8	0.238	0.122	0.086
0.9	0.217	0.110	0.077
1.0	0.200	0.100	0.070

The considerations above refer only to stationary working states. When a starting current surge occurs, it is usually not decreased by the current regulator, and in some cases it is even reinforced, since the cold resistance of the barretter is very small.

In the stabilization of the heating current in some radio sets it is necessary to take precautions against these starting surges. In these sets there is a small lamp for lighting the tuning scale connected in series with the heating filament of the indirectly heated cathodes. This lamp has a much smaller heat capacity and might burn through in the time necessary for the indirectly heated cathodes to heat up. For the stabilization of this heating current a special stabilizing resistor has been constructed which has a semi-conductor as limiting resistance in series with the filament. The size of this resistance is about 2000 ohms at room temperature; when, however, it is raised to 300° C by the heat of the current, only 100 ohms of the resistance remain [4]).

[4]) Resistors of this material are also used as starting resistances for motors, and have already been described in this periodical by P. C. van der Willigen, Philips techn. Rev. 1, 205, 1936.

Upon switching on the set the current is limited by the resistance of the semi-conductor to a small portion of its running value. During the heating up of this resistance the iron filament also becomes warm and assumes its regulatory function so that the current never reaches too high a value.

The action of a semi-conductor in series with the regulator valve is made clear in *fig. 10*. The upper

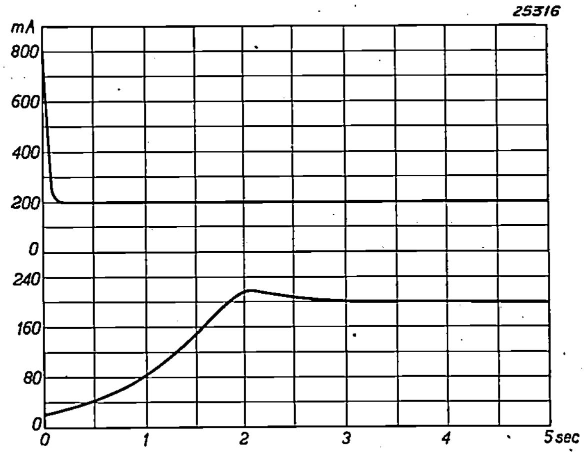


Fig. 10. Variation of the heating current in a radio receiver with receiver valves heated in series as a function of the time after switching on. Above: an ordinary barretter is in series with the filaments. Below: a regulator with a semi-conductor incorporated in it is in series with the filaments. The resistance of the semi-conductor has a very high value in the cold state, and therefore limits the starting surge of the current.

curve gives the variation of the current upon switching on the heating current without a limiting resistance, the lower curve shows the variation of the same when a starting resistance is used. In the first case the starting surge is about four times the running current, while the current reaches 10 per cent more than the running value at its maximum when a starting resistance is used, and reaches its stationary value only after several seconds. In this final state the greater part of the resistance of the valve is due to the iron filament, and only a small part is due to the semi-conductor, so that the regulatory action of the iron filament is manifested, and the current remains constant within narrow limits when the mains voltage fluctuates.

SECONDARY ELECTRON EMISSION

by H. BRUINING.

537.533.8

A discussion is given of the properties upon which the capacity for secondary emission of materials and surfaces is based. A special study is made of the choice of material in cases where, for the purpose of technical application, either a very large or a very small secondary emission is required.

Introduction

When a material, either a metal or a dielectric, is bombarded by electrons having a certain speed, a small portion of these electrons is reflected. The greater portion penetrates into the layer which is bombarded, and passes on its energy to the electrons present in that layer; the latter electrons may, when the direction of their motion is favourable, leave the bombarded surface. We are then concerned with secondary electrons. They can be observed when an electrode capable of capturing the electrons emitted is placed in the neighbourhood of the secondary emitting surface.

In all tubes in which surfaces are struck by electrons we are concerned with the phenomenon of "secondary emission". Among such are radio valves, cathode ray tubes and the like. There are a number of cases in which the secondary emission is a disturbing factor, so that efforts are made to suppress it, for example: the secondary emission of anode and screen grid in a tetrode, the secondary emission of the glass wall, the secondary emission of the grid of a broadcasting valve. On the other hand there is the possibility of making use of a high secondary emission; a well known application is the amplification of the electron current as in the electron multiplier and the secondary emission valve. A high secondary emission is further of importance in the fluorescing screens of cathode ray tubes and for the anode in the so-called dynatron. The technical application of secondary emission will be discussed in more detail in a subsequent article. The intention of the examples just given is only to make clear that it is important to be able to prepare substances which emit either very many or very few secondary electrons. We shall discuss in this article the laws of secondary emission so that it may be possible to make a choice of materials for technical applications.

Determination of the capacity for secondary emission

The capacity for emitting secondary electrons is always expressed in the number of secondary electrons, which are freed on the average by one primary electron. We shall indicate this ratio by the letter δ. The factor δ was determined in this laboratory for a large number of materials with the help of the apparatus shown in *fig. 1* [1]). It consists of a glass tube without cement or grease

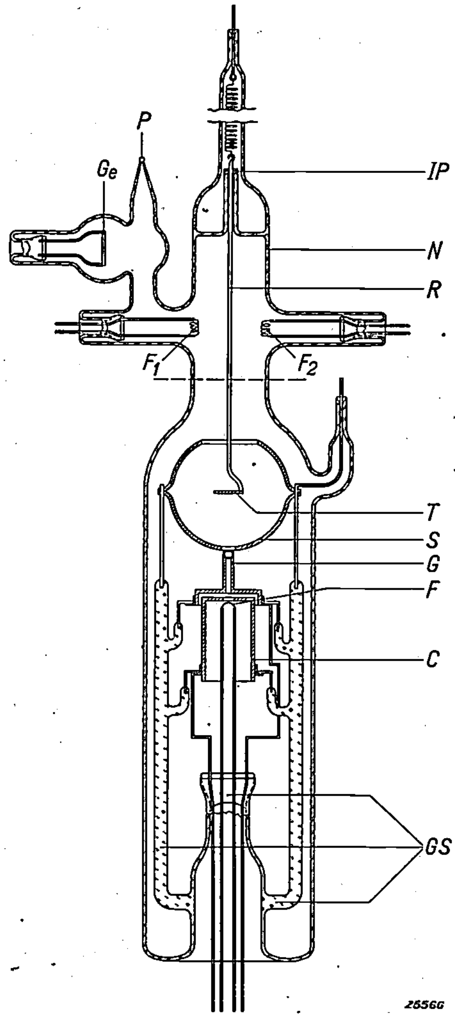


Fig. 1. Tube for the measurement of secondary emission of different materials.
F = cathode, source of primary electrons.
C = cylinder.
G = electron gun.
T = target whose secondary emission is to be determined.
S = sphere, collector for the secondary electrons.
GS = bridge upon which F, C, G and S are mounted.
R = rod.
IP = iron cylinder attached to R. By means of a magnet around IP the target can be drawn into the neck N of the tube.
N = neck of the tube.
F_1F_2 = tungsten filaments, by means of which the target T can be covered by sublimation with the substance to be examined.
Ge = plate from which getter can be evaporated.
P = point where the tube is sealed off.

1) See also H. E. Farnsworth, Phys. Rev. 25, 41, 1925.

joints. The tungsten filament F is the source of the primary electrons. This is surrounded by a cylinder C which concentrates the electrons on the opening of the canal G. The electrons passing through the aperture of C are gathered into a beam by the electron gun G which strikes the target T. The secondary emission of T is to be determined. The secondary electrons are collected on the sphere S, and the strength of the electron current to S is measured. For this purpose the potential of S must be slightly higher than that of T (10 volts is usually chosen as the potential difference). The target T can be covered with different materials. For this purpose it is attached to a rod R at the extremity of which a piece of iron IP is fastened which makes it possible to draw the target into the neck N of the tube by means of a magnet. The substances to be investigated are applied to heating spirals F_1, F_2, and can be deposited on the target by sublimation.

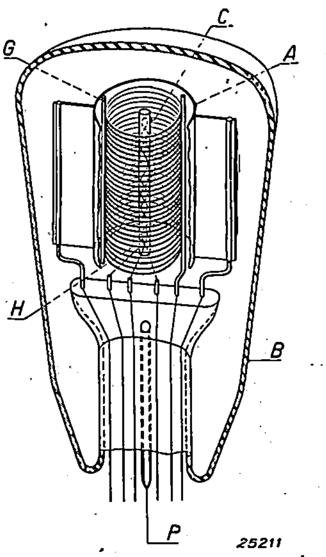


Fig. 2. Triode for measuring secondary emission.
C = cathode (indirectly heated, may also be tungsten wire).
H = heating filament of C.
G = grid.
A = anode.
B = wall of bulb.
P = connection to pump.

It is also possible, though with less accuracy, to determine secondary emission by means of a triode (*fig. 2*). The cathode C, heated by the spiral filament H supplies the primary electrons. The substance to be investigated is applied to the inner side of the anode A, while the grid potential Vg is chosen higher than the anode potential V_a, so that the secondary electrons are drawn from the anode to the grid G. A part si_k of the primary cathode current i_k will be captured immediately by the grid, and the current $(1-s)\, i_k$ will reach the anode. This latter will cause a current i_s of secondary electrons. If δ is the coefficient of the secondary emission of the anode surface, then

$$i_s = \delta\,[(1-s)\, i_k)] \quad \cdots\cdots \quad (1)$$

δ may be determined by measuring the grid current and the anode current separately. The anode current is

$$i_a = (1-s)\, i_k - i_s \quad \cdots\cdots \quad (2)$$

The grid current is

$$i_g = s\, i_k + i_s \quad \cdots\cdots \quad (3)$$

By eliminating i_k and i_s in the equations (1), (2) and (3) we find

$$\delta = 1 - \frac{i_a}{(1-s)\,(i_g + i_a)}.$$

According to the above equation the anode current $i_a = 0$ corresponds to $\delta = 1$. This is immediately clear. If every electron which strikes the anode frees on the average one secondary electron, then no mean charge is given to the anode and the anode current disappears. With many materials δ is greater than one in a certain range of the anode potential, and a negative anode current is the result.

The fact that a portion (si_k) of the primary electrons are captured by the grid is a disadvantage, since the factor s is difficult to determine accurately. In the arrangement shown in fig. 1 this disadvantage does not exist, and the method is more accurate.

Energy distribution of the secondary electrones

With the apparatus shown in fig. 1 it is also possible to determine the energy distribution of the secondary electrons. If the potential of the

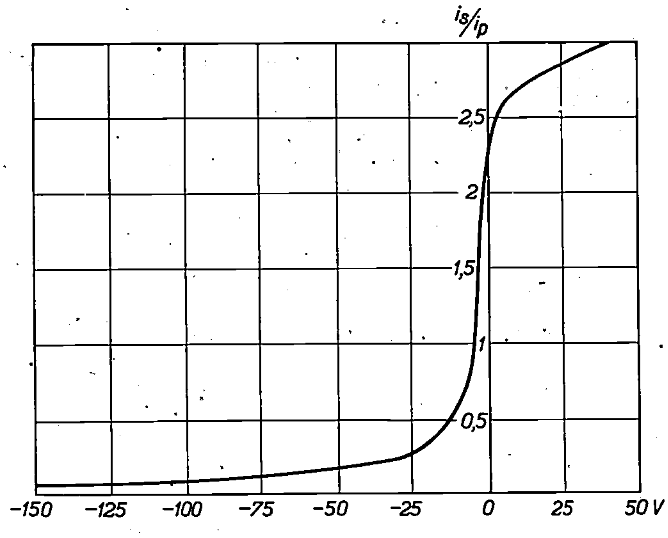


Fig. 3. Energy distribution of secondary electrons emitted by barium oxide.

sphere S is taken lower than that of T, all the electrons emitted by T cannot reach the sphere, but only those which are emitted with sufficient kinetic energy to overcome the applied counter potential. Thus if the potential of the sphere is for example 5 volts lower than that of the target, the sphere is struck only by secondary electrons with a kinetic energy greater than 5 electron volts. *Fig. 3* gives an idea of the results of such a measurement of the energy distribution obtained with barium oxide. The potential of the sphere S is plotted as abscissa, with the potential of T set equal to zero. The ratio of the secondary to the primary current i_s/i_p is plotted as ordinate. The intention of the figure in this case is to show that, when a secondary emission is desired, the electrode which must capture the secondary electrons must be at a higher potential than the secondary emitting layer itself. If, however, the desire is to avoid the capture of secondary electrons such a counter field must be applied that the transfer of electrons becomes impossible. It may be seen from the figure, for example, that in this case the secondary electron current decreases to 8 per cent of its original value when the potential of + 30 volts is reduced to — 30 volts with respect to the target T.

The relation between δ and the energy of the primary electrons

Figs. 4 and *5* show the dependence of the secondary emission δ on the potential difference V_p between the cathode and the secondary emitting target for different substances. It may be seen that δ as a function of V_p exhibits a maximum. The

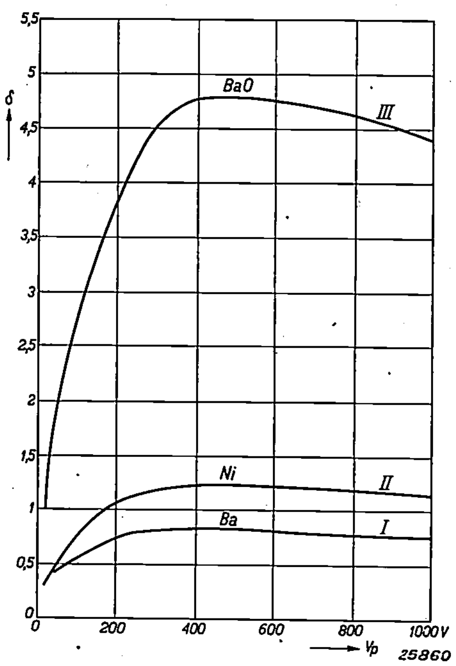


Fig. 5. Secondary emission of:
I barium deposited in a high vacuum
II nickel
III barium oxide

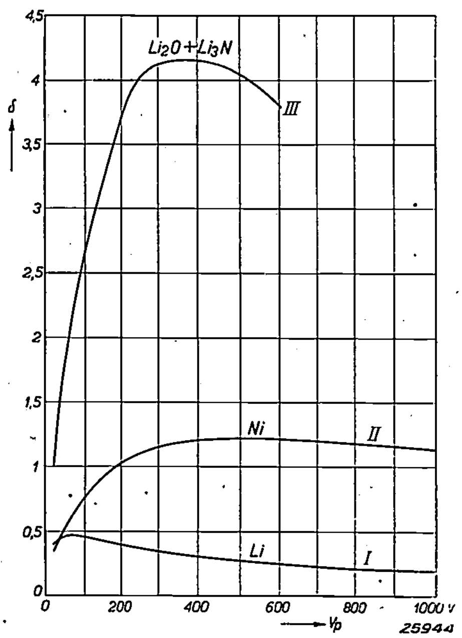


Fig. 4. Secondary emission of:
I lithium deposited by sublimation in a high vacuum.
II nickel.
III lithium deposited in the presence of residual gases.

occurrence of this maximum must be explained in the following way. With increasing potential the number of secondary electrons freed in the metal will increase, since the available energy present per primary electron becomes greater. On the other hand the primary electrons will penetrate deeper into the material with increasing velocity, so that the secondary electrons are freed at greater depths, and are therefore more strongly absorbed before they reach the surface. At a large value of V_p the absorption apparently dominates to such a degree that the curve shows a maximum. By means of an experiment which will not be described here [2]) we have estimated the average depth at which the secondary electrons are freed to be at least 20 layers of atoms in the case of nickel at V_p = 500 volts.

Substances with a high secondary emission

In this section we shall discuss the substances

[2]) H. Bruining, Physica 3, 1046, 1936.
For other investigations of this subject by the writer the reader is referred to:
H. Bruining and J. H. de Boer, Physica 4, 473, 1937.
H. Bruining and J. H. de Boer, Physica 5, 17, 1938.
H. Bruining, J. H. de Boer and W. G. Burgers, Physica 4, 267, 1937.

which can be used when a high secondary emission is required. The requirement made of these materials is that they be able to give a secondary current which is at least three times the primary current, in other words, that the factor δ be at least 3. We shall compare the various surfaces with regard to the factor δ in their behaviour when bombarded with primary electrons with an energy of 150 electron volts, since 150 volts is a suitable working voltage for the acceleration of the primary electrons.

The metals most commonly used in discharge tubes do not have a high secondary emission. As may be seen from fig. 4, for nickel $\delta = 0.94$ when $V_p = 150$ volts, so that it is impossible in this way to multiply electrons, that is to obtain more secondary than primary electrons. Other metals also of which electrodes may be made, such as copper, iron, molybdenum and tungsten, are useless for the same reason.

The secondary emission of all these metals is practically the same. This is not to be wondered at, since the work necessary to bring an electron out of the metals is of the same order of magnitude (4—5 volts) for all these materials. The question immediately arises as to the behaviour of metals with lower work functions. Known representatives of this group are the metals of the alkali metal group, Li, Na, K, Rb and Cs, and the alkaline earth metals, Be, Mg, Ca, Sr and Ba. Much less energy is necessary to cause an electron to leave these metals, as may be seen from the photoelectric and thermionic emission. It has often been assumed that the secondary emission is analogous in its behaviour, so that the factor δ of the alkaline earth metals should be larger than that for instance of Ni and Mo. *It is found, however, that metals with a low work function, at least in the V_p range with which we are concerned, have a lower secondary emission than the metals with a high work function.*

In figs 4 and 5 the variation of δ as a function of V_p is given for Li and Ba (curve *I*). Curves *II* give, for the sake of comparison, the variation of δ for nickel. This latter curve is reproduced in all figures up to *fig. 11*. The measurements on the metals Li and Ba, which readily form compounds with residual gases (oxides, nitrides) have been carried out with the greatest care. Every effort was made to obtain a good vacuum. If such measures are not taken, the results obtained are quite different, as may be seen from curves *III*. Curve *III* in fig. 4 refers to Li which was sublimed in a tube in which residual gases were still present, while curve *III* in fig. 5 refers to a layer of Ba which had been exposed to dry oxygen for some time and thus consisted mainly of BaO [3]).

The following conclusions may be drawn from these experiments.

1. *Secondary emission is determined by quite different factors than those which determine photoelectric and thermionic emission.*
2. The compounds of the electropositive elements form suitable surfaces for secondary emission, but the pure elements themselves do not.

Layers of the alkali halides deposited by sublimation have a high secondary emission. In *tables I* and *II* results are given which were obtained in an extensive series of experiments.

Table I. Secondary emission of metals at $V_p = 150$ volts.

Metal	δ	Work function in volts
Li	0.45 0.55	2.28
Cs	0.55	1.81
Be	0.52	3.16
Mg	0.90	2.42
Ba	0.63	2.11
Al	0.86	2.26
Cu	0.90	4.30
Ni	0.94	5.03
Fe	0.97	4.77
Mo	1.00	4.15
W	0.75	4.52

For practical use the secondary emitting substance must not only possess a good factor δ, but this value, moreover, must not change during emission. Many compounds with a high secondary

Table II. Secondary emission of compounds of electropositive elements at $V_p = 150$ volts.

Surface	δ
Li, sublimed in poor vacuum	3.25
NaCl	4.0
KCl	4.45
RbCl	4.15
CsCl	4.55
Cs, oxidized in dry oxygen	3-4
MgO	2-6 [4])
BaO	approx. 3
Al, covered with a thin layer of oxide	2.1

[3]) The contention is often encountered in the literature that elements with a low work function are good secondary emitters. This idea is based upon results of measurements on (probably) oxidized surfaces.

[4]) With this surface, which consists of small particles, the factor δ is very much increased by the external applied field. This is more or less the case with all the compounds mentioned here, see fig. 3 for example.

emission have the property of decomposing during emission, and forming metal agglomerates, which naturally causes a fall in the value of δ. In this respect the most satisfactory results up to the present have been obtained with MgO whose δ value remains very constant with time. The alkali halides are useless in this connection.

Mechanism of secondary electron emission

In connection with the above-described results it is important to find out whether compounds of metals with a less pronounced electro-positive character also have a larger δ than the metals from which these compounds are derived. From *table III* it may be seen that this is not the case. We shall now discuss what we consider to be the reason for this.

Table III. Secondary emission of compounds of metals with a high work function at V_p = 150 volts.

Compound	Secondary emission	Secondary emission of the pure metal
MoS_2	0.9	1.00
MoO_2	~ 1	1.00
WS_2	0.77-0.85	0.75
Cu_2O	0.99-1.05	0.90
Ag_2O	~ 1	1.05

As already mentioned, a primary electron passes on its energy to the electrons present in the material, and therby loses its kinetic energy along its path through the material. The kinetic energy is not passed on to a single electron; it is lost by "stages"; several electrons receive an impulse from the primary electron.

In order to find out what happens to these electrons we shall first consider a simpler case and recall experiments by Franck and Hertz. They studied what happens when the atoms of a gas or metal vapour are bombarded by electrons. In this case also the kinetic energy of the impinging electron is passed on to electrons of the material. It is found, however, that the latter electrons cannot take up every arbitrary amount of energy, but only very definite amounts. For example, the electrons of mercury atoms are unable to take up smaller amounts of energy than 4.8 electron volts. This is shown by the fact that electrons with a smaller energy suffer practically no loss of speed. Electrons which have passed through a somewhat greater potential difference than 4.8 volts, however, suffer a very great decrease in speed and their energy is for the greater part transformed into the well known ultra violet radiation of mercury vapour. If a beam of electrons falls upon a solid metallic compound a similar phenomenon takes place. There is a minimum amount of energy which can be taken up by the electrons. This amount is indicated by ε_1 in *fig. 6*. It will make a great difference whether this energy ε_1 is larger or smaller than the energy u which the electrons need in order to leave the metal (the so-called work function).

In the case of compounds of the alkali metals $\varepsilon_1 > u$ (fig. 6*a*). All electrons which have received an impulse from the primary electrons are thus accelerated enough to overcome the work function, and a large part of these electrons will actually leave the material. In the case of compounds of metals which are not strongly electro-positive $\varepsilon_1 < u$, and only the electrons which have received more energy than ε_1, for instance ε_2, from the primary beam, can take part in secondary emission. The factor δ is hereby considerably decreased.

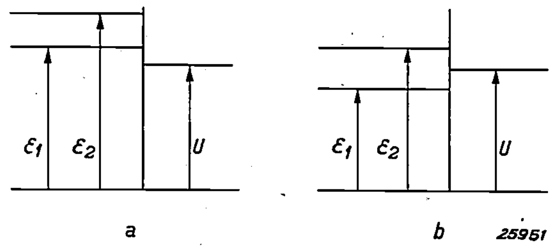


Fig. 6. a) The work function u of compounds of strongly electro-positive metals is lower than the minimum amount of energy ε_1, which can be passed on to the electrons in the solid substance.
b) With compounds of not strongly electro-positive metal $\varepsilon_1 < u$.

Materials and surfaces with only slight secondary emission

In the following we shall discuss substances and surfaces which have only a slight secondary emission. According to table I the metals lithium and beryllium must be especially considered in this connection. If these metals are used in tubes with a very high vacuum, they would indeed serve very well, especially at large values of V_p (of the order of magnitude of 1000 volts for example). During the experimental manipulations, however, there will always be residual gases present (from the oxide cathode, for instance) which combine with the lithium or beryllium and give in that way a substance with a large factor δ. In most practical cases a different kind of surface must therfore be sought.

In any case the choice will depend on the particular requirements made in practice. Thus for example for the grid of a transmitting triode (tetrode or pentode), it is necessary that the secondary emission does not become too great over the whole range of V_p [5]). The metal must, moreover, have a high melting point. Zirconium and titanium are suitable in these respects. Both metals still have more or less electro-positive characteristics, but they are covered with a very thin film of oxide when exposed to the air. Upon heating the oxide film is dissolved in the metal [6]), so that in a very simple way a metallic surface is obtained with a relatively low δ. The size of δ is shown in *fig. 7*. The secondary emission of a molybdenum or tungsten grid can also be decreased by covering it with a layer of ZrO_2, and heating it to a high temperature during evacuation. A zirconium surface is probably obtained by reduction of the ZrO_2.

Another method of keeping the secondary emission low is found in a suitable macroscopic structure of the surfaces. Up to now we have discussed only smooth surfaces. If a secondary electron has overcome the work function in the case of a smooth surface, it no longer encounters any material hindrances in its further motion (*fig. 8*). With a rough surface with a labyrinth-like structure the situation is different. In such a case there is the possibility that the electrons, after

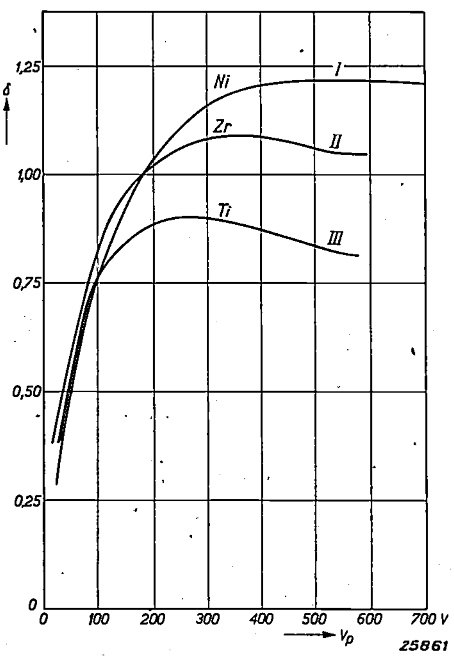


Fig. 7. Secondary emission of
I nickel
II zirconium
III titanium

they have left the material, will again be captured by a projection of the rough surface.

There are numerous methods of preparing such "labyrinth" surfaces. The surface may for example be composed of very small particles forming a

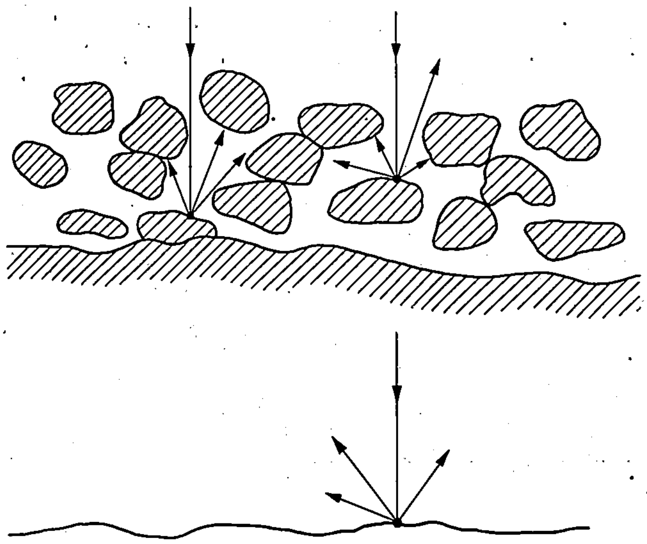


Fig. 8. Rough and smooth surfaces. In the case of the rough surface a portion of the secondary electrons are captured by the material itself, with the smooth surface they can escape unhindered.

porous layer such as a layer of carbon black [7]). All carbon layers, however, do not have a porous structure. Smooth layers of carbon can also be prepared, for instance by painting a plate with aquadag, which gives a layer built up of overlapping flakes of graphite. *Fig. 9* gives a comparison of the factor δ as a function of V_p for a smooth and a rough carbon surface. From this figure it may clearly be seen that aquadag gives a greater secondary emission than carbon black.

It is possible to obtain other substances besides carbon in this "black" form; the same may be done with many metals. One method is by subliming the metal in question in an atmosphere of one of the rare gases. The atoms in the vapour state collide with the atoms of the gas and therefore cover longer paths before they condense on the wall. The metal will then form conglomerates in the gas which will be deposited as such on the surface. These conglomerates may reach a size of several millimetres. The surface then appears black. Metal deposited in a vacuum usually has a mirrorlike surface. *Fig. 10* gives an example of the difference in the secondary emission of silver deposited respectively in a vacuum and through an atmosphere of 1 mm of argon.

[5]) If this factor is too large as in the case of most of the metals used for this purpose such as Ni, Mo, W and the like the grid current as a function of the grid voltage begins to show a negative coefficient of direction. See also H. G. Boumeester, Philips techn. Rev. 2, 115, 1937.

[6]) This is not the case with the metals mentioned in table I.

[7]) In this section we discuss only conducting materials. With non-conducting materials, making the surface rough may cause an increase in δ.

The black layer formed is not stable to temperature in the case of many metals. During degassing at a high temperature the black modifications often have the tendency to sinter (baking together of the small particles to larger ones), whereby the labyrinth structure disappears. The surface then becomes grey. The sintering temperature is much lower than the melting point, but runs parallel to it when the metals are compared among themselves. In general use must be made of metals with a melting point higher than 2000° C. In this respect the black forms of tungsten and molybdenum are particularly suitable. Carbon black also, because of the high melting point of carbon, has a high sintering temperature, but has the disadvantage of often giving off much gas and of dissolving rapidly in nickel at about 1000° C.

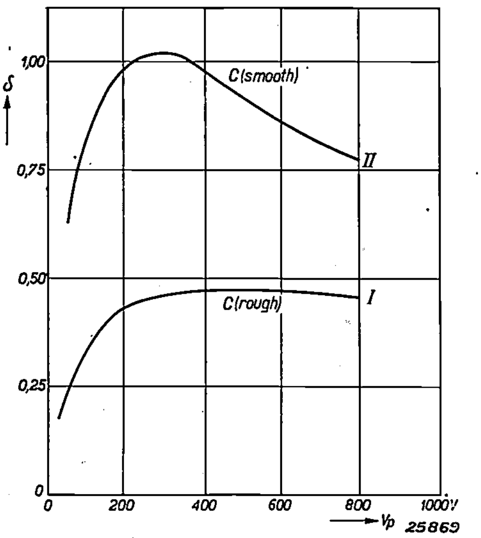


Fig. 9. Secondary emission of:
I aquadag
II carbon black

The behaviour of secondary emitting surfaces in a radio valve

If it is desired to use a surface with very high or very low secondary emission in a radio valve, the coefficient of the secondary emission must remain constant during use. The electrodes of a radio valve are not only exposed to a bombardment by electrons, but are usually also gradually covered with barium atoms which evaporate from the hot oxide cathode.

Since metallic barium, as has appeared from the experiments discussed above, has a low coefficient of secondary emission, a surface with high secondary emission will in general be spoiled by being covered with barium atoms. It may therefore be necessary to avoid having barium atoms strike the secondary emitting surface.

It is also undesirable for a surface with only a slight secondary emission to be covered with barium. There is a possibility that a gradual oxidation will take place, and therefore that the coefficient of secondary emission will increase considerably. If we allow the oxide cathode of a radio valve whose anode is covered with carbon black to burn for 100 hours without applying an anode voltage, we obtain a variation in the secondary emission similar to that of curve *AB* in *fig. 11*. If an anode voltage of 100 volts is now applied, with an anode current of 40 mA, the factor δ falls again during the course of 100 hours to its original low value. The adsorbed barium atoms apparently disappear from the surface. This effect may be explained by a migration of barium atoms along the surfaces of particles of the carbon black toward greater depths of the carbon layer. The speed of this migration is much increased by an increase of temperature. When a carbon black surface of suitable structure is bombarded by electrons strong local increases of temperature occur which were estimated at about 300° C in the case in question.

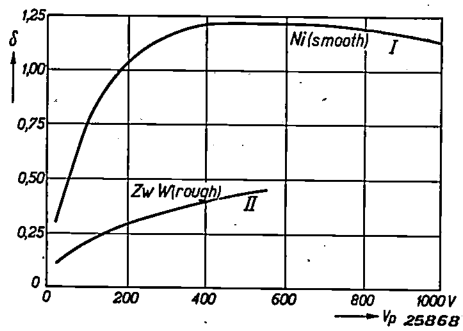


Fig. 10 Secondary emission of
I silver, deposited in a high vacuum
II silver deposted through 1 mm of argon

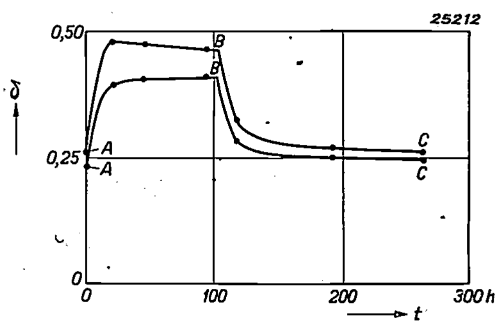


Fig. 11. Variation of the secondary emission of carbon black opposite an oxide cathode.
AB : $V_a = 0$; no electrons reach the anode; secondary emission rises with the time.
BC : $V_a = 100$ volts; electrons strike the anode; secondary emission falls to the original value.

AN ALTERNATING CURRENT DYNAMO WITH A FLAT CHARACTERISTIC FOR BICYCLE ILLUMINATION

H. A. G. HAZEU and M. KIEK.

621.313.322 : 629.118.3

A description is given of an alternating current dynamo in which the voltage at the terminals depends only slightly on the number of revolutions. This dynamo is intended for the feeding of a bicycle lamp, and is so constructed that when the bicycle is pushed at a walking speed it is adequately lighted, while at high speeds the lamp still has a satisfactory life.

Construction of the bicycle lamp dynamo type no. 7405

For the supply of a bicycle lamp alternating current dynamos are used not only of the type in which the permanent magnet is the stator and the winding the rotor, but also dynamos in which the winding is stationary and the permanent magnet rotates. The latter construction has been chosen for the Philips bicycle dynamo no. 7405, since, among its other advantages, no brushes are necessary, and therefore the possibility of poor contacts is excluded. For the generation of the necessary voltage at low speeds these dynamos must have a relatively large number of poles.

The field generated by the winding tends to demagnetize the steel magnet, and the demagnetization depends upon the load. In order to keep the resulting decrease in voltage as small as possible, we have used the special magnet steel "ticonal", whose magnetic induction at the working point is only slightly influenced by an alternating field. The so-called reversible permeability which determines this property is equal to only about 2. The induction B thus varies only slightly with the field strength H, and is not much greater in the metal than in air. The permanent magnet then need have no actual poles, but may simply be constructed in the form of a smooth, thin-walled cylinder magnetized to have eight poles.

As may be seen in the opened dynamo shown in *fig. 1,* the cylindrical magnet rotates within eight laminated pole shoes which run parallel to the axis and are joined above to form eight U-shaped magnetic circuits. Above the magnet a coil is wound about these projections in such a way (cf. *fig. 2*) that the energy fluxes through the projections which lie within the coil are opposite in direction to those through the projections which lie outside it.

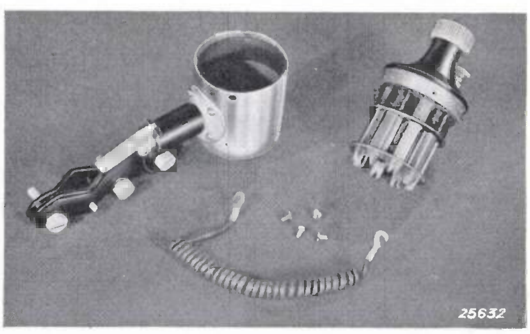


Fig. 1. Photograph of an opened bicycle dynamo type no. 6405.

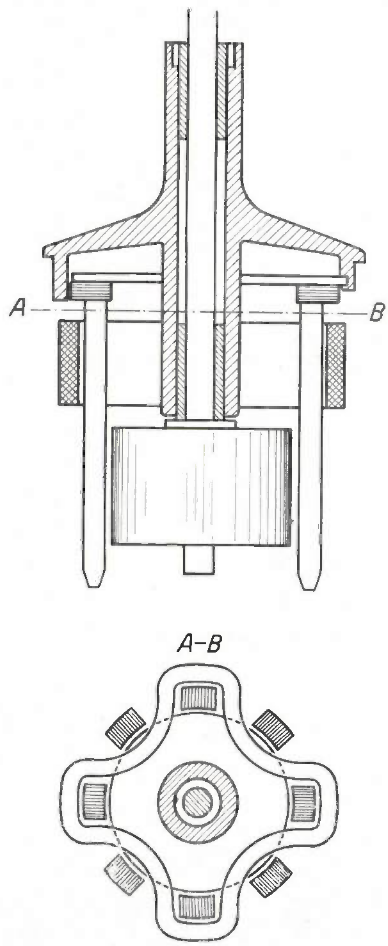


Fig. 2. Sketch of the construction of the bicycle dynamo. Above, a horizontal cross section, and below, a cross section through *AB*.

Method of obtaining a flat characteristic

In designing a dynamo suitable for a bicycle lamp it is desirable that it be so constructed that it is not only able to supply the necessary current of 0.5 to 1 A with the usual voltage of 4 to 6 volts at the normal speed of a bicycle of 13 to 15 km (8 or 9 miles) per hour, but that the lighting of the bicycle be adequate when it is pushed by hand at a speed of about 4 km ($2^1/_2$ miles) per hour. Moreover at a speed of 20 to 25 km (14 or 15 miles) per hour which for instance is the normal speed of tandems the voltage must not rise to such an extent as to cause the lamp to burn out.

If the leakage of the magnetic field, the losses and the field generated by the currents in the armature itself are neglected, the voltage of such an alternating current machine is directly proportional to the speed of its rotation, as is indicated in *fig. 3,* curve *a*. The light intensity of

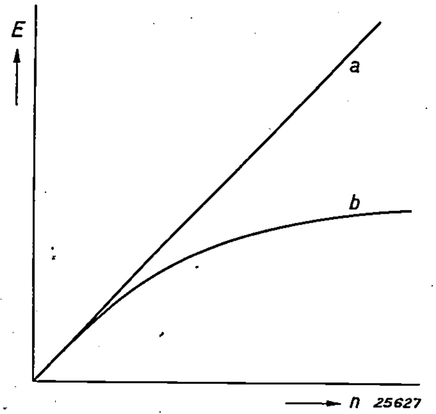


Fig. 3. Variation of the voltage at the terminals E of a dynamo with speed of rotation n. Curve a, non-loaded, and b, load of one or more lamps.

the lamp would then increase very rapidly with the speed of revolution, because a 1 percent increase in voltage on the lamp means about 3.5 per cent increase in the light intensity. However the requirement is made of the bicycle dynamo designer, that the construction of this alternating current dynamo be such that, upon being loaded with one or more bicycle lamps (for instance with both a front and a rear light as is now required by law in the Netherlands), a loss in voltage will occur, such that the voltage at the terminals depends upon the speed of rotation in the manner shown in curve b, fig. 3. In order to obtain such a relation the loss in voltage must increase more rapidly than the speed.

We shall prove in the following that a behaviour like that given in curve b is that of every dynamo, and we shall discuss the measures which may be taken to bring it about that the current becomes constant at as low a speed as possible.

If w be the number of turns of the winding, Φ the flux enlosed per turn, then the EMF generated

$$E = -w \frac{\mathrm{d}\,\Phi}{\mathrm{d}t}.$$

If we assume that the flux alternates sinusoidally as a result of passing the magnet poles, we may write:

$$E = -w\,j\,\omega\,\Phi, \quad \cdots\cdots \quad (1)$$

The flux Φ is generated partly by the permanent magnet and partly by the current I in the winding. This second part will be approximately proportional to the current, so that:

$$\Phi = \Phi_0 + C\,I \quad \cdots\cdots \quad (2)$$

It follows further from (1) that, when we introduce the total resistance R of the winding and the lamp:

$$I = \frac{E}{R} = -\frac{wj\omega}{R}\,\Phi \quad \cdots\cdots \quad (3)$$

By eliminating Φ from equations (2) and (3) we finally obtain:

$$I = \frac{-w\,j\,\omega\,\Phi_0}{R + wj\omega C}, \text{ therefore } |I| = \frac{w\,\omega\,\Phi_0}{\sqrt{R^2 + w^2\,\omega^2\,C^2}} \quad (4)$$

At a small speed of rotation the inductive resistance $w\,\omega\,C$ is still small with respect to R, and the current therefore increases proportionally to the speed. At a high speed, $w\,\omega\,C \gg R$ finally, and the current supplied becomes constant.

Equation (4) shows the essential point in the construction of a bicycle dynamo. The product $w\,\omega\,C$ must become large with respect to R for the lowest possible riding speed.

We shall consider the three factors of this product in detail.

The number of turns in the winding w

If there is a definite amount of space available for the windings of the coil, when w is increased, not only $w\,\omega\,C$ increases, but the internal resistance rises even more rapidly and R with it. Therefore an optimum number was found to be about 200; with this the dynamo has an internal resistance of about 4 ohms.

The frequency ω

The frequency ω is given by the product of the number of poles and the speed of rotation. A large number of poles is thus an advantage. In this respect the design of the dynamo in question is very advanced. A magnet with eight poles and having the small dimensions desirable in a bicycle dynamo could only be realized by using a magnet steel which, as a result of its low permeability at the working point, is only very slightly influenced by the demagnetizing field. Such a steel is the above-mentioned "Ticonal".

The constant C

The constant C indicates the flux through the coil which is generated by a current of 1 A. When the dynamo is constructed with no magnetic leakage, so that all the lines of force of the coil must go through the magnet, the flux is not very great because the magnet has a rather low permeability. For this reason the leakage is expressly made great by having the windings on projections of laminated iron which extend considerably beyond the winding and then run parallel to each other over quite a long distance and at a relatively small distance from each other (cf. figs. 1 and 2). The magnetic field of the winding can then be closed across these projections without the magnet being coupled.

Influence of eddy current

Since large losses of voltage must occur if the characteristic is to be flat, it might be thought that eddy currents would have a favourable influence on this factor. This is actually found to be the case; the loss of energy due to this cause is, however, not allowable in practice.

As previously explained in this periodical [1]), the eddy current losses may be represented by a resistance depending on the frequency, which is in parallel with the winding, and therefore also in parallel with the lamp. The loading resistance R is hereby reduced and the current will, according to equation (4), become larger and already reach a constant value at a lower frequency. In order to obtain a strong effect by this means, however, the parallel resistance due to the eddy currents must be small with respect to the external loading resistance. This results in the fact that the greater part of the energy is used to no purpose, and the dynamo drives too heavily. This may cause slipping.

This possibility of improving the shape of the characteristic therefore cannot be utilized. On the contrary, eddy currents must be avoided as much as possible. For this purpose the whole magnetic circuit is constructed of transformer stampings of less than $^1/_2$ mm thick.

Data of Bicycle Dynamo Type No. 7405

This dynamo has an internal resistance R_i of 4.1 ohms. At a speed of 14 km (9 m. p. h.) the normal voltage of 6 volts is delivered with 0.5 A, and at 38 km (24 m. p. h.) the voltage rises only to 7 volts. The no-load voltages are then however 24 and 57 volts respectively. In *fig. 4* the no-load

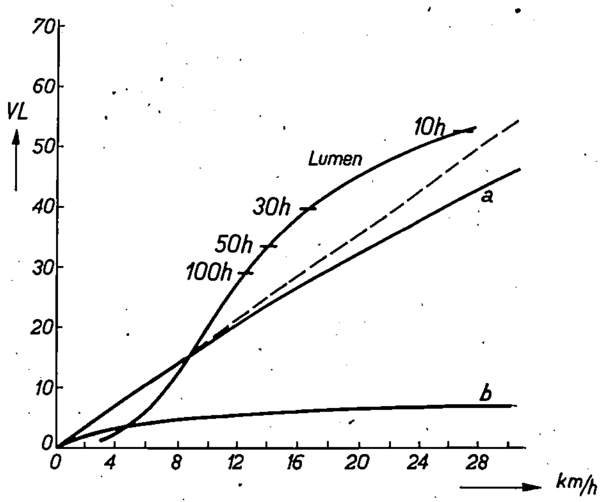


Fig. 4. Voltage in volts and light flux in lumens as functions of the riding speed. Curve *a* gives the variation of the no-load voltage and *b* the terminals voltage with the load of a lamp for 6 volts and 0.5 A. In the light flux curve the length of life of the lamp corresponding to the light flux also is indicated.

characteristic is given (a) and the variation of the voltage at the terminals (b) as a function of the riding speed when loaded with an ordinary bicycle lamp taking $^1/_2$ A at 6 volts. We see from this that at a speed of 9 km (6 m. p. h.) the terminals voltage is still 5 volts, and that it decreases only to 4 volts at a speed of 6 km (6 m. p. h.), so that even at the lower speed the bicycle is adequately lighted. Further it may be seen from the no-load characteristic that the eddy current losses are relatively small. If there were no losses at all, the no-load characteristic would be straight (dotted), so that at 30 km (20 m. p. h.) the voltage would be 54 volts. This value is decreased by only 8 volts to 46 volts, due to the eddy currents, as may be seen from curve (a).

Moreover fig. 4 (c) also shows how the light flux in lumens varies with the speed, and the length of life of the lamp is also indicated for several values of the light flux. It may be seen that at the normal speed of 15 km per hour (10 miles) a light flux of 36 lumens is delivered and the lamp has a satisfactory life. At the very high speed of 27 km per hour (18 m. p. h.) the light flux increases only to 53 lumens, and the life of the lamp is 10 hours.

In *fig. 5* the so-called external characteristic of the dynamo is given, that is the terminal voltage R_k, the current I and the energy delivered W as a function of the external resistance R_u at a

[1]) J. W. Köhler, Philips techn. Rev. 2, 194, 1937.

constant speed of 13.5 km per hour ($8^1/_2$ m. p. h.). From the short circuit current of 0.58 A at

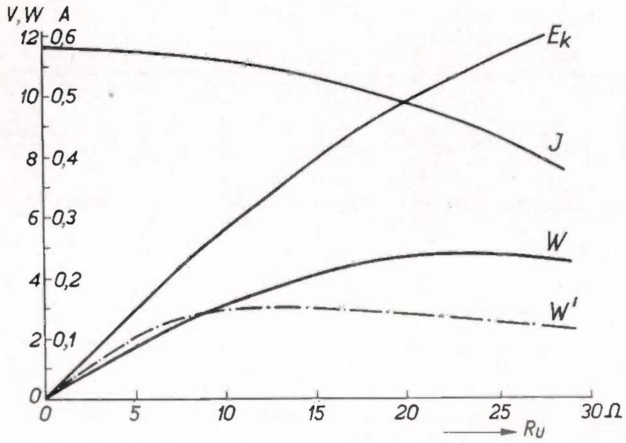


Fig. 5. Terminals voltage E_k, current I and power delivered W as dependent on the external resistance R_u; the so-called external characteristic of the dynamo. The power delivered W' is indicated by a dot-dash line for a dynamo provided with a winding of 140 turns, in contrast to the final construction with 200 turns.

$E_k = 0$, the current I decreases with increasing R_u, and the voltage increases. The power delivered reaches a minimum between 20 and 30 ohms external resistance. For the sake of comparison an indication is also given (dot-dash line) of the way in which the power delivered W' varies when 140 turns of the same wire are used instead of 200, the number which was used in the final model. As was to be expected the maximum power for 140 turns occurs at a smaller external resistance.

If a lamp of 12 ohms for 6 volts and 0.5 A is used, the power delivered lies below the maximum, but a very flat characterisitic is then obtained, as is shown in fig. 4. When the dynamo is used care must be taken that the lamps in parallel with each other together take up a current at the terminal voltage supplied which is equal to the current indicated for the dynamo. If it is desired to use a lamp for 6 volts and 0.05 A for the rear light (required by law in the Netherlands), a head light for 6 volts and 0.45 A must be used when the dynamo is for 6 volts and 0.5 A.

Besides the original model for 6 volts and 3 W, dynamos constructed on the same principle for 8 volts and 4 W and for 10 volts and 4.5 W are now supplied. In *fig. 6* a bicycle dynamo type no. 7405 with accompanying head light is shown.

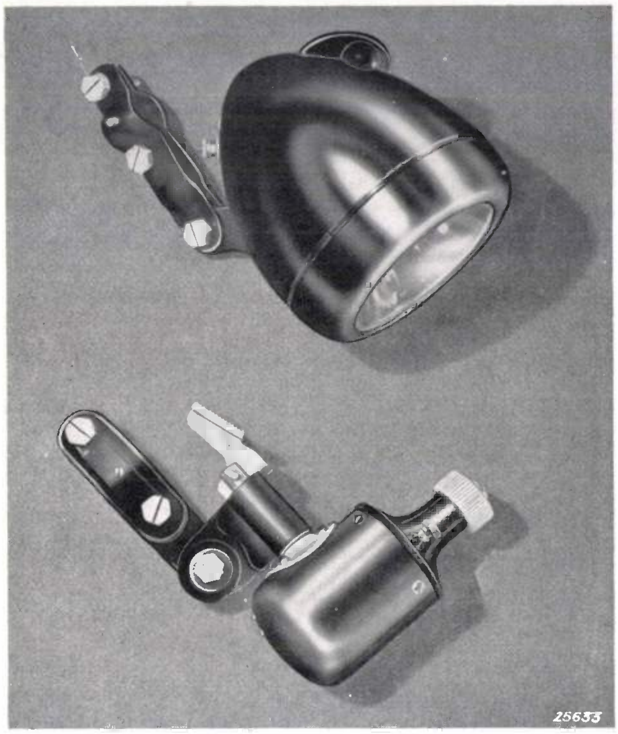


Fig. 6. Photograph of a bicycle dynamo type No. 7405 with the accompanying head lamp.

THE USE OF THE "PHILORA" HP MERCURY DISCHARGE LAMP IN AN ENLARGING APPARATUS

by J. A. M. VAN LIEMPT.

621.327.3 : 771.44 771.318.5

When a "Philora" HP mercury discharge lamp is used as the source of light in an enlarging apparatus, the exposure time can be considerably reduced, so that it is possible to work with silver chloride paper and with plenty of yellow light while adjusting and developing.

An ordinary electric lamp is usually used as a source of light in an enlarging apparatus. The power of this lamp must be limited to several hundred watts at the most because of the dimensions of the apparatus and with the heat development of the lamp.

Since when ordinary silver chloride contact paper is used the exposure times are relatively long, much use is now made of bromide paper, which is so much more sensitive that the exposure time can be shortened about 100 times. This is a great advantage especially in repetition work. The disadvantage of this paper, however, lies in the fact that it is somewhat more expensive and that in the dark room, the yellow light must be replaced by the weaker orange or light red light, under which it is more difficult to judge contrasts, especially in the darker parts of the picture. For these reasons experiments have been carried out by the court photographer P. Ziegler in The Hague, by the firm of Boomen in Amsterdam as well as by the writer of this article with the HP mercury discharge lamps which radiate relatively shorter wave lengths, and especially with the HP 300 lamp as a source of light. Normal contact printing paper was used. Since these lamps are supplied with the discharge tube placed at the geometrical axis of the lamp, the lamp must be placed perpendicularly to the axis of the tube of the enlarging apparatus. There would otherwise be a disturbing black spot in the middle of the positive as a result of the fact that the lamp radiates only a small amount of light in the direction of its axis.

It was found that the exposure time necessary to produce a good print can be about $^1/_{500}$ of that required by an ordinary electric lamp when the time is recalculated for the same energy, and about $^1/_{200}$ of that necessary with an ordinary electric lamp, recalculated for the same light flux. This means actually that when a mercury lamp of the type HP 300 (75 W) is used instead of an ordinary lamp of 200 W, and with normal silver choride contact printing paper, exposure times are just as short as those otherwise attained only with bromide paper. There is in addition the great advantage that one no longer need use light red light, but may again work with yellow light. As a source of yellow light, sodium light is highly to be recommended [1]). With this source it is possible to have a high intensity of light in the dark room for the making and developing of enlargements. The fact that the picture may be judged under normal intensities of illumination is very favourable to the result. With the much weaker red illumination it is impossible to distinguish the finer contrasts and the picture gives too hard an impression (and too dark in the shadow parts), so that one is always inclined to stop the development too soon.

The small heat development of the HP 300 mercury lamp, which is a favourable feature in any case, makes it possible in speed work, such as is often necessary in press photography, to enlarge the still wet negative, which has only been fixed but not washed, with less risk than when ordinary electric light is used where there is danger of the gelatine film becoming soft and blurring.

On point must still be noted. The cheaper enlarging apparatus has a lens which is only moderately achromatic. In many cases this will present no difficulties, but with special negatives such as those of line drawings and the like when printed on soft paper, there will be some difficulty.

When mercury light is used the possible achromatism of the lenses is somewhat more pronounced, because the composition of the light deviates from that of white light. It is therefore advisable to use a well corrected lens such as those supplied by reliable optical manufacturers.

If it is found that the lens already in use is unsatisfactory, improvement may be obtained by making the stop smaller. As a rule, for those special cases a decrease to *f*. 9 is enough. This of course is accompanied by some loss of light, but the exposure time remains so short that the advantage of the use of a mercury lamp is still very considerable.

[1]) For the value of the "Philora" sodium lamp in photography see Philips techn. Rev. 2, 24, 1937.

THE EXAMINATION OF THE MACRO-STRUCTURE OF MATERIALS AND PRODUCTS WITH THE HELP OF X-RAYS. IV

by J. E. DE GRAAF.

539.26534.844

In a purely mechanical manufacturing process the differences between the products produced will in general be slight. It is usually sufficient to examine samples only, since an irregularity in the manufacturing process seldom occurs.

As soon, however, as the changeable human being begins to take an important part in the manufacture of the product, every article must in principle be checked, at least when high quality is important. This subsequent control is naturally somewhat unsatisfactory, so that attempts are made by means of work tests to determine the "value" of the worker. In such tests maximum performance only is usually determined, since the worker approaches such a test quite differently than he approaches his routine work; and in any case great deviations from the test "value" may be expected as a result of psychological and physical influences and the like. In precise work considerable differences have more than once been found between work done at night and that done in the daytime, or between articles made on Monday and those made on Tuesday. By continually controlling production with the help of X-rays it is possible both to detect too dangerous mistakes promptly, and to take measures toward removing the causes of mistakes (by improving the illumination for instance). The very presence of such a control often has a strong preventive influence, which is manifested not only in better work, but also in better care of implements and tools.

One of the operations to which the above specially applies is electric welding. Mistakes often occur through nonchalance or ignorance, which mistakes are different for every welder. These mistakes must then be analyzed in order to be able continually to point out to the welder his own particular failings.

The diagnosis of mistakes in electric welding [1]

In the first place a distinction can be made between mistakes in the welding process and mistakes in the preparation of the weld. Among the latter, for example, belong an incorrect form of electrode and dirty plates.

In the second place a distinction can be made between, I, gas enclosures (circular dark spots with sharp out-lines and heavy blackening), II, slag enclosures (usually vague irregular limits) and III, flaws of a linear form such as cracks and imperfect joints (contrast closely dependent on the direction of the rays [2]). Several examples of these defects in butt welding are given below together with an account of some of their causes.

With other welding rods than the used PH-50 (cf. Philips techn. Rev. 2, 135, 1937) the illustrated faults only show unimportant changes.

A. Mistakes in the welding operations

Group I: Gas enclosures.

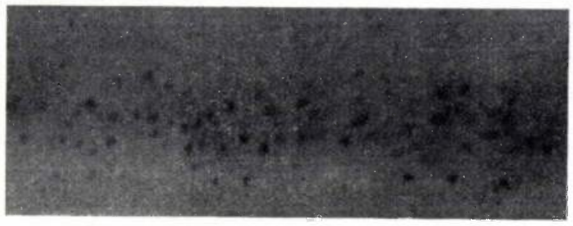


Fig. 1. Gas enclosures.

Characteristics: circular dark spots often in groups with relatively sharp outlines and heavy blackening; their diameter is seldom more than 2 mm.

Causes: too long an arc (too little protection from air); too heavy current (especially where the welding has been done with the last short end of the welding electrode, which often stands red-hot).

Group IIa: Slag enclosures entirely within the molten metal.

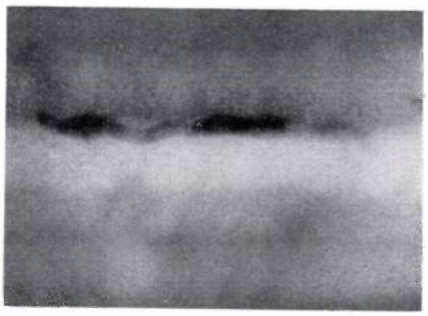


Fig. 2. Running out [3]) of the slag on a flat plate.

[1]) As in the precading articles (Philips techn. Rev. 2, 315, 350 and 377, 1937). The x-ray photographs are reproduced in original size.

[2]) Cf. Philips techn. Rev. 2, 351 (1937).

[3]) Under the term "running out" is meant the fact that the liquid flux (cf. fig. 16 of the article by J. Sack onWel-

Characteristics: a heavy line, sometimes several mm wide with irregular curved outlines and moderate to heavy blackening.

Causes: incorrect waving motion; the electrode passes over the already molten slag and the drops of iron fall to both sides of a slag channel.

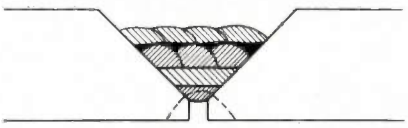


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Fig. 3. Diagramatic representation of slag enclosures in the V shaped weld groove. Footnote [4]) fig. 12 explains the signifacation of the dotted line.

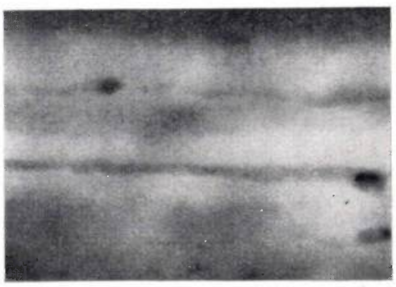

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Fig. 4. Running out of the slag in the grooves between the convex surface of the first layer of metal deposited (cf. fig. 3).

Characteristics: narrow, vague, oscillating lines, often of moderate blackness; interruptions often occur (sometimes with the ends bent outward).

Causes: due to too weak current or too limited waving motions the surfaces of the underlying layer have become too convex and grooves are therefore formed between them. (Other narrow grooves, such as those at the bottom of a V-joint may lead in this way to slag enclosures which have their own special characteristics).

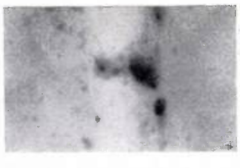

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Fig. 5. Crater enclosure.

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ding and Welding Rods in Philips techn. Rev. 2, 129, May 1937) does not cover the material deposited (the weld) but runs out over the cold surface of the work ahead of the weld.

Characteristics: a half-moon shaped spot with the width of the weld, low degree of blackening and sharpness; often accompanied by gas bubbles.

Causes: after an interruption of the arc, welding is recommenced at or before the old crater, instead of beyond it; slag runs into the hollow of the crater when another layer is applied on top of it.

Group IIb: Slag enclosures between deposited metal and base metal.

As soon as there is any suspicion of the presence of this kind of defect, X-ray photographs must be taken with the rays incident along the original boundary of the base metal, for instance along the wall of the welding groove. This type of defect may result in dangerous cracklike faults (imperfect joining).

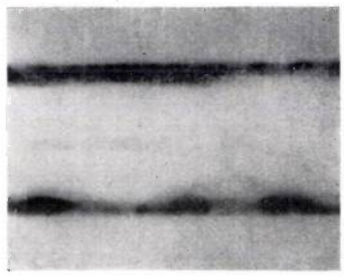

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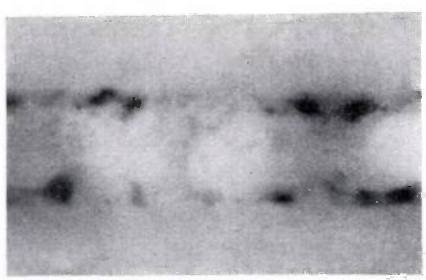
Fig. 6. The supply of heat to the groove is continually too low.

Characteristics: often a fairly broad line with a straight sharp boundary at the wall of the groove, extending over long distances.

Causes: the welding electrode was not directed sufficiently into the angle between the weld and the wall of the groove, so that the slag in that angle was able to solidify. This is often due to the postion of the welder, when it is such that he is unable to see into the angle.

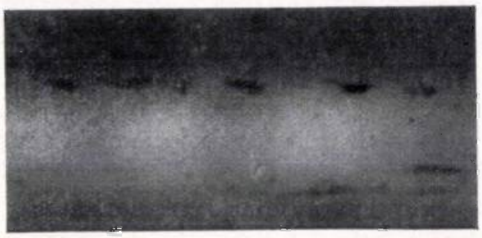


25203

Fig. 7. The supply of heat is very irregular and often insufficient.

Characteristics: enclosures limited by a straight line along the wall of the groove and having very irregular shapes and many interruptions.

Causes: too long an arc, which in the waving motion jumps over too quickly to the wall of the groove and heats the angle between the weld and the wall irregularly.

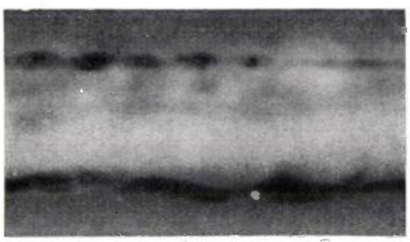


25204

Fig. 8. Triangular enclosures.

Characteristics: triangular enclosures bounded by a straight line along the wall of the groove and sharply outlined, at fairly regular intervals (alternately along both walls of the groove).

Causes: due to too rapid advance of the weld, a zigzag layer was formed; between the wall of the groove and the angles or curves of the weld, slag is enclosed during the welding of the following layer.



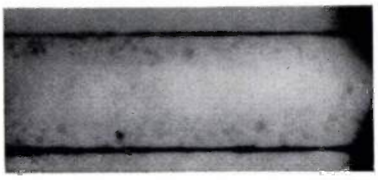
25205

Fig. 9. Undermining.

Characteristics: enclosures lying outside the original limits of the weld, and having a wavy outer boundary and a vague irregular inner boundary; usually accompanied by the fault of fig. 6.

Causes: too long an arc (as in fig. 6) which melts too much metal from the wall of the groove; in depositing the following layer, the undermined spot is usually filled with slag instead of with metal.

Group IIIa: Imperfect joints.



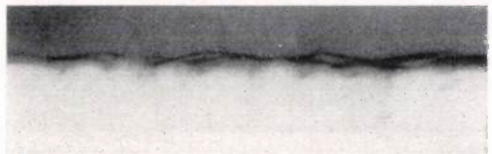
25206

Fig. 10. Imperfect joints.

Characteristics: straight lines, usually fine and without side branches lie along the walls of the groove. They are almost always accompanied by slag enclosures.

Causes: The electrode metal is not fused together with the base metal because of inadequate supply of heat to the latter (for instance because the arc was not sufficiently directed upon the wall of the groove just as in the case of the defect in one layer only in fig. 5; in fig. 9 the defect lies higher up on the wall than in fig. 5).

Group IIIb: Cracks.



25207

Fig. 11. Cracks.

Characteristics: fine, often branched lines; they are usually smooth in the base metal, while in the metal deposited or along the wall of the groove they are usually irregular; these defects are often difficult to discover.

Causes: because of immoderate or irregular shrinking the stresses occurring are too great to be relieved sufficiently by flowing of the metal.

B. Faults in the preparation

25208

Fig. 12. Faulty cutting out of the groove for a counter weld [4]).

Characteristics: a line close to the outer surface of the counter weld, somewhat irregular and often broken, with vague boundaries and slight blackening.

Causes: the cutting out of the root was done with a chisel which made too sharp a groove. The slag ran out in this groove as in fig. 3.

4) After the V-shaped weld groove has been welded full, in the case of thick plates the portion indicated by a dotted line in fig. 3a (the "root") is cut out, since this portion is often filled with slag and badly deposited metal.

25209

Fig. 13. Faults in the filling of too narrow holes.

Characteristics: on a stereo-photograph which gives a special impression of the enclosures, the slag was found to be enclosed in the form of a spiral.

Causes: in the too narrow hole the slag could not be sufficiently well removed during the filling of the hole with molten metal with a spiral motion.

25210

Fig. 14. Fault in the method of filling the seam.

Characteristics: at regular distances in the seam occur collections of vague linear or pointed enclosures of slight darkening.

Causes: a section of the seam about 30 cm in length was completely filled, then a following section, etc. The rough end of one finished section was not cut out smoothly before the beginning of the next.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS' GLOEILAMPENFABRIEKEN

1251: J. van Niekerk and F. Franken: On the influence on chicks of very large amounts of antirachitic vitamine of animal origin (Landbouwk. T. **49**, 742 - 748, Oct. 1937).

The influence of the administration of large quantities of vitamine D on the growth and death of chicks is studied. The limitation of growth and the death rate are found to be proportional to the degree of the excessive concentrations of vitamine D.

These results are only observed with high overdosage of the vitamine. In connection with the small concentration in which vitamine D appears on the market, possible mistakes need cause no anxiety as to a poisoning effect. If the administration of vitamine D to chicks which have had too much is stopped, they continue to grow normally. Finally methods are given for the quantitative determination of the poisonous effect of antirachitic preparations with chicks as test animals.

1252*: R. Houwink: Die Viskosität in konzentrierter Lösung (Oesterr. Chem. Z. **40**, 472 - 475, Nov. 1937).

The behaviour of concentrated solutions, which is of such fundamental significance in colloid chemistry problems and in the treatment of the subject of highly polymerized materials is discussed in this article (Cf. **1199**).

*) An adequate number of reprints for the purpose of distribution is not available of those publications marked with an asterisk. Reprints of other publications may be obtained on application to the Natuurkundig Laboratorium, N.V. Philips' Gloeilampenfabrieken, Eindhoven (Holland), Kastanjelaan.

1253: M. J. O. Strutt und A. van der Ziel: Die Ursachen für die Zunahme der Admittanzen moderner Hochfrequenz-Versterkerröhren im Kurzwellengebiet (El. Nachr. Techn. **14**, 281 - 293, Sept. 1937).

From measurements carried out on high-frequency amplifier valves, input losses, output losses and reaction admittance are found to increase considerably with the frequency for short waves. In contrast to the usual conception that the cause of this must always be sought in the transition times of the electrons, it follows from many measurements that $^1/_2$ to $^1/_3$ of the input losses and practically all of the output losses and reaction admittance must be ascribed to induction effects inside and outside the modern high-frequency amplifier valve of normal type (AF 3 and AF 7 for instance). A general theory about the influence of induction effects on the admittances is proposed for different types of valves. The formulae found agree with the measurements, which is not the case with the formulae for the influence of the transition times of the electrons on the admittances of the amplifier tubes.

1254: E. M. H. Lips: The construction of castings with relation to the "stress-free" requirement (Metallbewerkung, **4**, 289 - 291 and 309 - 312, Oct. 1937).

Non-uniform cooling of a casting causes casting stresses. A description is given of what is meant by non-uniform cooling, and in what way uniform cooling, and thus absence of stresses in the casting,

25209

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1255: W. Elenbaas: Über eine Kombination der hydrodynamischen Theorie des Wärmeübergangs und der Langmuirschen Theorie (Physica **4**, 761 - 765, Oct. 1937).

Langmuir's theory of the transfer of heat by wires of various diameters to gases at various pressures is combined with the hydrodynamic theory, especially that of Nusselt.

Langmuir's theoretical curve, with a suitable choice of parameters, agrees with the experimental curve of Nusselt. By combination of the two theories the transfer of heat in all diatomic gases can be calculated with the help of a single constant.

1256: J. A. M. van Liempt and J. A. de Vriend: Testing focal plane shutters (Physica **4**, 811 - 827, Oct. 1937).

A description is given of the method by which a focal plane shutter can be thoroughly tested with the help of a cathode ray tube with respect to shutter time, width of the slit, motion of each of the blinds, time of exposure and efficiency.

1257: J. L. Snoek: Volume magnetostriction of iron and nickel (Physica **4**, 853 - 862, Oct. 1937).

The volume magnetostriction is measured by means of an aluminium dilatometer, in which the thermal expansion of the test rod and of the filling liquid due to the magnetocaloric effect are compensated by the expansion of the dilatometer. This is only exactly the case at a definite temperature, but corrections can be applied for other temperatures.

At room temperature accurate measurements have been carried out on pure iron and nickel. As was to be expected, the slope of the line which represents the specific elongation as a function of the magnetic field is scarcely influenced by the state of internal stress of the test rod. There is, however, considerable parallel displacement. The change in the effect with temperature is not greater than the experimental errors.

1258: C. J. Bakker: On the efficiency of the production of artificial radio-active substances by slow neutrons (Physica **4**, 863 - 870, Oct. 1937).

Solutions of various salts were irradiated with neutrons from a radium-beryllium preparation in order to produce artificial radioactive elements. The efficiency in the production of radio-chlorine with a half life of 35 min was 1%, radio-bromine, half life 18 min: 16%, radio-bromine, 4 hours: 7%, radio-bromine, 36 hours: 0.6%, radio-iodine, 35 min: 3% and radio-manganese, 2.5 hours: 6%. Minimum values were determined for the capture cross section for slow neutrons.

1259: E. M. H. Lips: Härtemessungen an Gefügebestandteilen (Z. Metallk. **29**, 339 - 340, Oct. 1937!.

The measurement of the hardness of structural components is described (of Philips techn. Rev. **2**, 179, 1937).

1260*: C. J. Dippel: Reversible and irreversible phenomena in the adsorption of vapours on salt surfaces deposited by sublimation (Chem. Wbl. **34**, 676 - 678, Oct. 1937).

The molecular relation in the adsorption of caesium and iodine on sublimed calcium fluoride is not constant. This may be explained, in connection with other investigations, by means of the assumption that this relation is actually the product of a pure adsorption relation and a relation of the areas of the surface.

Under the influence of the adsorption of caesium the available portion of a calcium fluoride surface may become greater (opening of space between lamellae) due to a kind of swelling. This is partially reversible and partially irreversible, since a portion of the substance adsorbed in held by irreversible bonds. This hypothesis also offers an unforced explanation of the adsorption hysteresis found in the adsorption of iodine.

1261: A. E. van Arkel and J. H. de Boer: Supplement to "Chemische Binding" (31 pages; Centen, A'dam 1937).

In this supplement to the book „Chemische Binding als electrostatisch verschijnsel" published in 1929, the authors have summarized some of the most important points regarding which theories have been developing rapidly in recent years. Among other features, the formation of complexes and the phenomenon of condensation under the influence of van der Waals' forces are discussed.

1262: M. J. O. Strutt: Measures des constantes caractéristiques de quelques penthodes haute-fréquence pour des fréquences de 1,5 à 300 mégacycles par seconde (Onde él. **16**, 553 - 577, Oct. 1937).

This article is a French version of **1243**.

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RELATING TO THE PRODUCTS, PROCESSES AND INVESTIGATIONS OF
N.V. PHILIPS' GLOEILAMPENFABRIEKEN

EDITED BY THE RESEARCH LABORATORY OF N.V. PHILIPS' GLOEILAMPENFABRIEKEN, EINDHOVEN, HOLLAND

VOLTAGE REGULATION OF DIRECT CURRENT GENERATORS BY MEANS OF TRIODES

by N. A. J. VOORHOEVE and F. H. DE JONG. 621.313.13 : 621.385.3 : 621.316.722

The variations in the voltage from a direct current dynamo can be diminished by regulating the excitation current in relation to the terminal voltage. A very accurate regulation can be attained by using as excitation current the anode current of a triode amplifier valve whose input voltage is influenced by the variations of the terminal voltage.

A triode regulator working on this principle is described in this article. By means of various examples attention is drawn to the very effective characteristics of this arrangement.

A direct current generator delivers a voltage which exhibits certain variations depending on the load and on accidental circumstances. These variations can be diminished by regulating the excitation current in relation to the terminal voltage. This can be done by varying a resistance in the excitation circuit of the generator or of the separate exciter.

For rapid and accurate regulation it is desirable that the necessary variation of the resistance be effected by a device which functions as nearly as possible without time lag. Electronic tubes therefore offer an ideal means of carrying out this regulation.

In principle such regulation can be carried out with an arrangement such as that given in *fig. 1.* The circuit functions in the following way. When the terminal voltage v_b of the generator falls, due for instance to a variation in the load, the voltage at the grid of the triode T becomes less negative with respect to the cathode, so that the anode current increases. The anode current flows through the excitation circuit of the generator. The increase of the excitation current works against the decrease of the terminal voltage.

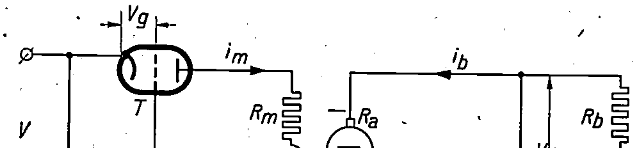

Fig. 1. Diagram of the circuit for regulation of the voltage with a direct current generator by a triode in series with the field winding.

For very accurate regulation it will in general be necessary not to conduct the variable terminals voltage directly to the grid of the regulating triode, but through a voltage amplifier. This voltage amplifier M is shown schematically in fig. 1.

Theory of the triode regulator

The most important requirements demanded of a voltage regulator are the following:

1. The regulation must be accurate. The ratio in which the voltage fluctuations are decreased by means of the regulator may be considered a measure of the accuracy.
2. The regulation must take place rapidly. Every regulation device has a certain lag, so that upon very sudden variations in load the voltage fluctuations are not diminished immediately.

If, for example the load on the generator with voltage regulation is suddenly removed, the voltage then varies as shown in *fig. 2.* It decreases suddenly and then increases gradually again to slightly more than its original value. According to the nature of the circumstances this adjustment to the final value may proceed exponentially or with a damped oscillation. In both cases we may characterize the speed with which the voltage approaches its final value by a time factor in a power of e, and this factor, the damping constant, forms a measure of the speed of regulation.

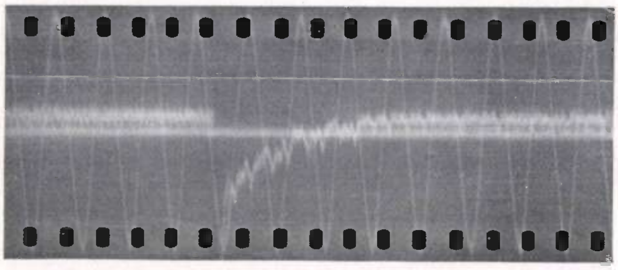

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Fig. 2. Variation of the terminals voltage upon sudden application of the load.

In order to examine the possibilities offered by the triode regulator in these two respects, we shall study the action of the apparatus mathematically. The following symbols will be used (see fig. 1):

L_a, R_a self-inductance and resistance of the armature circuit of the generator,
L_b, R_b self-inductance and resistance of the load applied,
i_b, v_b current and voltage supplied by the generator,
pv_b the portion of v_b conducted to the regulator,
m amplification factor of voltage amplifier M,
i_m, v_m excitation current and voltage,
L_m, R_m self-inductance and resistance of the excitation circuit,
μ, v_g, R_i amplification factor, grid voltage and internal resistance of the regulator triode,
v_{g0}, i_{m0}, i_{b0} values of v_g, i_m, i_b corresponding to an arbitrarily chosen stationary state.

The following equations can be set up. For the armature circuit:

$$(L_b + L_a)\frac{\mathrm{d}i_b}{\mathrm{d}t} + (R_b + R_a)\, i_b = \mathrm{EMF} \quad (1)$$

With slight variations of the excitation current the EMF will vary linearly with i_m:

$$\mathrm{EMF} = v_0 + c_1\, i_m \quad (2)$$

The current i_m comes from the triode circuit. In this circuit:

$$(i_m - i_{m0}) = \frac{\mu}{R_i}(v_g - v_{g0}) + \frac{1}{R_i}(v_m - v_{m0}),$$

where:

$$v_m - v_{m0} = -R_m\,(i_m - i_{m0}) - L_m\frac{\mathrm{d}i_m}{\mathrm{d}t}.$$

Thus:

$$\mu\,(v_g - v_{g0}) = (R_i + R_m)\,(i_m - i_{m0}) + L_m\frac{\mathrm{d}i_m}{\mathrm{d}t}. \quad (3)$$

The voltage fluctuations in v_g originate from fluctuations in the voltage of the generator by voltage division and amplification in M.

$$v_g - v_{g0} = -mp\cdot(v_b - v_{b0}) =$$
$$= -mp\left[c_1\,(i_m - i_{m0}) - R_a\,(i_b - i_{b0}) - L_a\frac{\mathrm{d}i_b}{\mathrm{d}t}\right]. \quad (4)$$

If we eliminate $v_g - v_{g0}$, i_m and $\frac{\mathrm{d}i_m}{\mathrm{d}t}$ from equations (1) to (4) we find for i_b a homogeneous linear differential equation of the second order:

$$P\frac{\mathrm{d}^2 i_b}{\mathrm{d}t^2} + Q\frac{\mathrm{d}i_b}{\mathrm{d}t} + R\left(i_b - \frac{S}{R}\right) = 0, \quad (5)$$

in which:

$$P = L_m\,(L_b + L_a),$$
$$Q = L_b\,(R_i + R_m + mp\mu c_1) + L_a\,(R_i + R_m) +$$
$$+ L_m\,(R_b + R_a),$$
$$R = R_b\,(R_i + R_m + mp\mu c_1) + R_a\,(R_i + R_m),$$
$$S = (v_0 + c_1\, i_{m0})\,(R_i + R_m) +$$
$$+ mp\mu\, c_1\,(v_0 + c_1\, i_{m0} - R_a\, i_{b0})$$
$$= (v_0 + c_1\, i_{m0})\,(R_i + R_m) + mp\mu\, c_1\, v_{b0}.$$

As is known, the general solution of differential equation (5) can be written in the form of the sum of two powers of e:

$$i_b - S/R = A_1\, e^{-\alpha_1 t} + A_2\, e^{-\alpha_2 t}. \quad (6)$$

If this expression is introduced into equation (5) the following quadratic equation is obtained for α_1 and α_2:

$$P\,\alpha^2 - Q\,\alpha + R = 0.$$

The solutions are:

$$\alpha_1, \alpha_2 = \frac{Q}{2P} \pm \sqrt{\frac{Q^2}{4P^2} - \frac{R}{P}}, \quad (7)$$

We shall see later from examples that the square root in equation (7) can be real as well as imaginary. In the first case the current will approach its final value exponentially according to the formula below; in the second case in the form of a damped oscillation. The following formulae are valid for the two cases respectively:

$$i_b = \frac{S}{R} + A_1\, e^{-\left(\frac{Q}{2P} + \sqrt{\frac{Q^2}{4P^2} - \frac{R}{P}}\right)t} + A_2\, e^{-\left(\frac{Q}{2P} - \sqrt{\frac{Q^2}{4P^2} - \frac{R}{P}}\right)t} \quad \text{(8a)}$$

and:

$$i_b = \frac{S}{R} + e^{-\frac{Qt}{2P}} \left(A_1' \cos t \sqrt{\frac{R}{P} - \frac{Q^2}{4P^2}} + \right.$$
$$\left. + A_2' \sin t \sqrt{\frac{R}{P} - \frac{Q^2}{4P^2}}\right), \quad (8b)$$

where A_1' and A_2' as well as A_1 and A_2 are arbitrary constants. P, Q, R and S are positive according to definition. From this it follows that the exponents of the powers of e of equation (8) are always negative; instability is thus out of the question.

The speed of regulation is determined by different quantities depending on whether we are concerned with equation (8a) or (8b).

The first case, exponential adjustment, occurs when $Q^2 > 4PR$. In this case the speed of regulation is determined by the smaller of the two exponents of the powers of e of equation (8a). This is:

$$a_2 = \frac{Q}{2P}\left(1 - \sqrt{1 - \frac{4PR}{Q^2}}\right).$$

In some practical cases Q^2 will be much greater than $4\ PR$, so that we may approximate the term $\sqrt{1 - 4\ PR/Q^2}$ by $1 - 2\ PR/Q^2$. We then obtain:

$$a_2 = \frac{R}{Q}.$$

The second case, $Q^2 < 4\ PR$, gives a damped oscillation according to equation (8b), the damping constant (β) is:

$$\beta = \frac{Q}{2P}.$$

Examples in the following will show that a_2 as well as β may easily be equal to 3000 sec^{-1}, for instance, and the regulation then takes only $^1/_{3000}$ sec.

Further discussion of the result

In the case of an oscillatory adjustment the following value is found for the damping constant $\beta = \frac{Q}{2P}$:

$$\beta = \frac{L_b m p \mu c_1}{2 L_m (L_b + L_a)} + \frac{1}{2}\frac{R_i + R_m}{L_m} + \frac{1}{2}\frac{R_b + R_a}{L_b + L_a} \quad (9)$$

and for the frequency of oscillation: $\sqrt{\frac{R}{P} - \frac{Q^2}{4P^2}} = \omega$:

$$\omega^2 = \frac{R_b (R_i + R_m + m p \mu c_1) + R_a (R_i + R_m)}{L_m (L_b + L_a)} - \beta^2. \quad (10)$$

In the case of a dead beat adjustment ω^2 is negative; the damping is then determined by:

$$a_2 = \beta - \sqrt{-\omega^2}.$$

In general the loading resistance has an inductive component ($L_b > 0$); in this case β and ω^2 as functions of the amplification factor m vary qualitatively as shown in *fig. 3a*. For a sufficiently high amplification ω^2 will become negative; the damping constant, which at first increases with m, then becomes smaller again and approaches the final value R_b/L_b, which no longer depends upon the generator, but only on the loading impedance. If, however, the load consists of a purely ohmic resistance, then β and ω^2 vary according to fig. 3b; an oscillatory adjustment cannot now be made exponential by increasing the amplification, and the damping of the oscillation is not affected in this case.

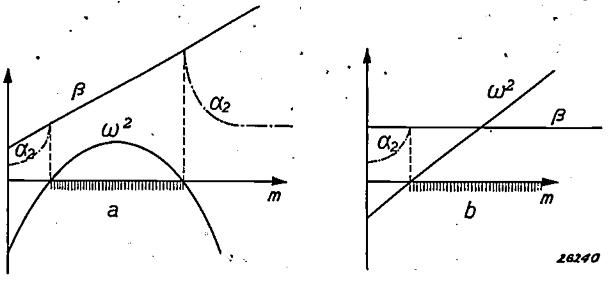

Fig. 3. Variation of a_2, β and ω^2 as functions of the amplification m. In the cross hatched area of the abscissa an oscillating adjustment must be made; outside of the area the adjustment is exponential.
a) The loading impedance has self-inductance.
b) The load is a purely ohmic resistance.

In order to estimate the accuracy of the regulation we must find out how the final value of

$$v_b = i_b R_b, \text{viz. } v_b = \frac{S}{R} R_b,$$

depends upon the loading resistance R_b (the self-inductance L_b has no influence on the final value). If R_{b0} and R_b are two different values of the loading resistance, then the difference between the corresponding generator voltages is:

$$v_b - v_{b0} = - \frac{(v_0 + c_1 i_{m0}) \dfrac{R_a (R_b - R_{b0})}{R_b R_{b0} (1 + R_a/R_{b0})}}{\dfrac{m p \mu c_1}{R_i + R_m} + 1 + \dfrac{R_a}{R_b}} \quad (11)$$

If voltage regulation is not applied the same expression is obtained with $m = 0$:

$$(v_b - v_{b0})_z = - \frac{(v_0 + c_1 i_{m0}) \dfrac{R_a (R_b - R_{b0})}{R_b R_{b0} (1 + R_a/R_{b0})}}{1 + \dfrac{R_a}{R_b}} \quad (12)$$

The accuracy of regulation was defined above by

$$N = \frac{(v_b - v_{b0})_z}{(v_b - v_{b0})},$$

and from (11) and (12), therefore:

$$N = \frac{\dfrac{mp\mu\, c_1}{R_i + R_m} + 1 + \dfrac{R_a}{R_b}}{1 + \dfrac{R_a}{R_b}} .$$

In most practically occurring cases $mp\mu c_1 \gg R_i + R_m$ and $R_a \ll R_b$, so that by close approximation:

$$N = \frac{mp\mu\, c_1}{R_i + R_m} . \qquad (13)$$

It follows from equation (13) that the accuracy of regulation increases proportionally with the voltage amplification.

Example:

In the case of a direct current dynamo of 460 volts, 2.5 kW with a triode regulator the following values were observed:

m	$= 360$	R_a	$= 7.2$ ohm
p	$= 1$ resp. 0.1	R_b	$= 585$ ohm
c_1	$= 1700$ ohm	R_i	$= 360$ ohm
$L_a = L_b$	$= 0.155$ H	R_m	$= 1675$ ohm
L_m	$= 82.5$ H		

For $p = 0.1$ a damped oscillation and for $p = 1$ a dead beat adjustment is found. With the help of the values given above the following values can be calculated for $p = 0.1$:

$P = 82.5\,(0.155 + 0.155) = 25.6$

$Q = 0.155\,(360 + 1675 + 0.1{\cdot}360{\cdot}6{\cdot}1700) + 0.155\,(360 + 1675) + 82.5\,(585 + 7.2) = 106500$

$R = 585\,(360 + 1675 + 0.1{\cdot}360{\cdot}6{\cdot}1700) + 2\,(360 + 1675) \cong 216{\cdot}10^6$

and for $p = 1$:

$P = 25.6$, $Q = 620\,000$, $R \cong 216{\cdot}10^7$.

Fig. 4. Philips voltage regulator type 4142.

From this it follows that:

for $p = 0.1$: $\beta = 2080\ \mathrm{sec}^{-1}$; $N = 180$,
for $p = 1$: $\alpha_2 = 4200\ \mathrm{sec}^{-1}$; $N = 1800$.

The lag is therefore only several ten thousandths of a second, and the voltage fluctuations in the last case are reduced to practically $^1/_{2000}$. This shows that a very rapid and very accurate regulation may be obtained by the use of triodes.

Practical construction of the triode regulator

The Philips voltage regulator, type 4142, is reproduced in *fig. 4*, the diagram of the circuit in *fig. 5*. The method of functioning agrees in the main

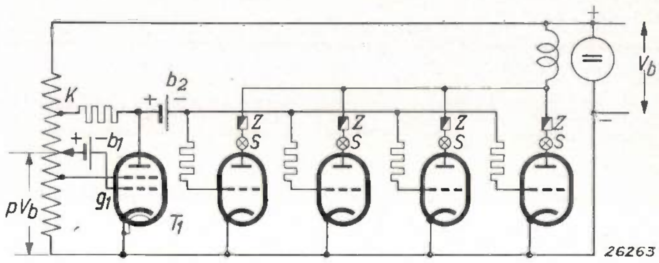

Fig. 5. Diagram of a voltage regulator with triodes.

with the principle illustrated in fig. 1 and it can easily be examined by means of the diagram given in fig. 5.

The voltage amplifier M is a resistance amplifier. The tetrode T_1 serves as amplifier valve. The regulation voltage pv_b is taken from the potentiometer P and conducted *via* the grid battery b_1 to the control grid g_1 of the tetrode. When the terminal voltage v_b increases the anode current of the tetrode also increases, and the anode voltage decreases in consequence. This variation in voltage is transmitted *via* a second grid battery b_2 to the grid of a number of triodes in parallel, which supply the excitation current. The anode voltage of the triodes, like that of the tetrode, is supplied by the generator. In order, however, if necessary to use the tetrode with a higher anode voltage, point K (anode voltage tetrode) has an extra external lead.

The following points may be noted in regard to the arrangement:

1. The regulator works with several valves connected in parallel. This gives increased security during action. If one of the valves develops a defect the regulation proceeds practically unchanged with the remaining valves.
2. In the anode circuit of each regulator triode there is a small pilot lamp s and a fuse z. The pilot lamp is extinguished as soon as the anode current of the triode becomes abnormally small

due to a defect. The fuse burns through, when the current becomes abnormally great due to a short circuit.

3. The grids of the regulator triodes are connected to the grid battery through high resistances. This prevents the total anode current (excitation current) from becoming too great in case of a short circuit between grid and cathode.

Some practical applications of the triode rapid regulator

The special characteristics of the triode rapid regulator, namely, very high accuracy of regulation and very rapid regulation, make it suitable for use where a very constant voltage is essential.

It may be used in laboratories, schools, institutions for the calibration of instruments and in towing tanks for ship design. We shall briefly describe the last two applications, where an expensive accumulator battery needing much upkeep can be replaced by a much less expensive converter installation with a triode regulator needing little attention.

Use of the triode regulator in a calibrating station

For the calibration of electrical meters the following is the usual method.

Separate alternating current generators driven by a common direct current motor are used to feed the voltage and current coils. In order to be able to measure at different phase displacements the voltage generator is provided with a stator which may be turned about its axis.

During a calibration process the voltages and the frequency of both generators must remain strictly constant; this requirement is met when the direct current motor is connected with a source of voltage which has practically a constant voltage. A battery of accumulators is commonly used as source of voltage; the triode regulator makes it possible to replace the battery by a converter with direct current dynamo whose terminals voltage is kept strictly constant by the regulator.

A triode regulator for such an installation is shown in *figs. 6* and 7.

The converter consists of a three-phase synchronous motor coupled with two direct current generators of 22 kW, which feed the three wire system of 2×125 volts for the calibrating units.


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Fig. 6. Converter consisting of a three-phase synchronous motor and two direct current generators of 22 kW.

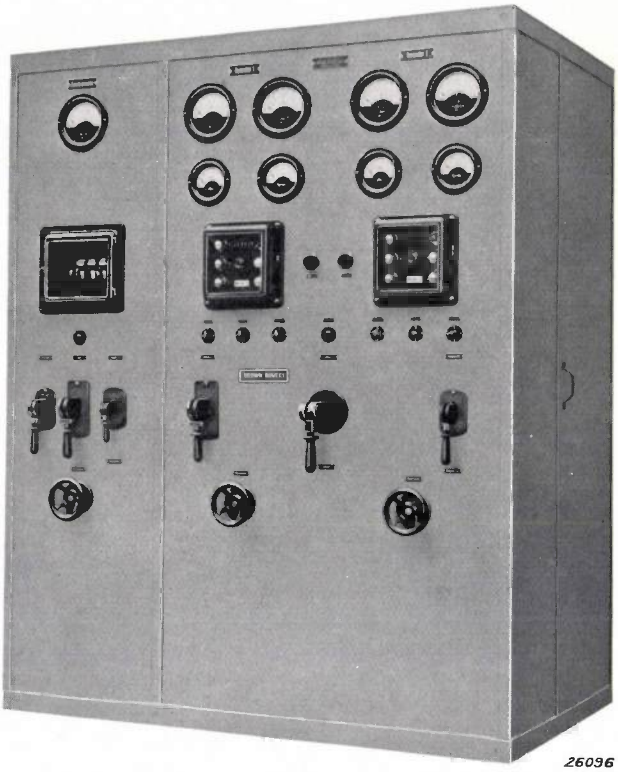
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Fig. 7. Switch board for converter of fig. 6 with two built-in Philips voltage regulators type 4142.

Fig. 6 shows the converter, fig. 7 the switch box with Philips triode rapid regulators type 4142 built in.

The values $R_m = 1055$ ohms and $c_1 = 330$ ohms are obtained by measurement from the generator.

The triode regulator is so adjusted that $p = 1$, $m = 500$; the regulator triodes have an amplification factor $\mu = 6$ and an internal resistance of 1500 ohms. There are five triodes connected in parallel, so $R_i = 300$ ohms. From this it follows that:

$$N \cong \frac{6 \cdot 500 \cdot 330}{1055 + 300} \cong 730.$$

The fall in voltage from no load to full load without regulation is 6 volts, with regulation the fall in voltage is decreased to $6/730 \cong 0.008$ volts.

The terminal voltage is 125 volts; the deviation from the average terminals voltage with regulation is therefore only $\pm$ 0.03 per cent. It has been proved by measurements that in practice, upon sudden variations in the load, this very high accuracy of regulation is actually attained. After longer times the terminal voltage will in general show somewhat greater variations, because the grid battery voltages also exert an influence on the terminal voltage. The percentage change of v_b is equal to the percentage change of the battery voltage. With good quality batteries these changes proceed very slowly, so that it is practically possible to keep the terminals voltage constant within 0.1 per cent.

Application in the case of a towing tank

A triode regulator for the converter installation of a towing tank was supplied to the Netherlands Shipbuilding Testing Station in Wageningen. The towing truck used to draw models of ships through the tank runs on four wheels. Each wheel is driven by a direct current motor of about 20 H.P. The motors are excited separately from direct current mains of 220 volts. The armatures of the motors are connected in series with a maximum armature voltage of 145 volts per motor. The running speed of the motors is regulated by varying the armature voltage, and for this purpose the armatures, in series with each other, are fed from a circuit with a variable voltage from 8 to 600 volts.

Fig. 8. Switch board of converter for 220 volts direct current with built-in rapid regulator.

Since the speed of the towing car during a test must be strictly constant, it is necessary that the voltages of the two direct current circuits be kept accurately constant. This is done in the following way.

Fig. 9. Switch table for 72 kW Ward-Leonard converter with built-in rapid regulators.

The circuit of 220 volts is fed by a converter with a synchronous three-phase motor and a compound dynamo whose terminal voltage is regulated by a triode regulator. For feeding the variable 600 volts circuit, which supplies the energy to the motors of the car, a battery of accumulators is usually employed. In order to avoid this expensive and elaborate installation in the testing station,

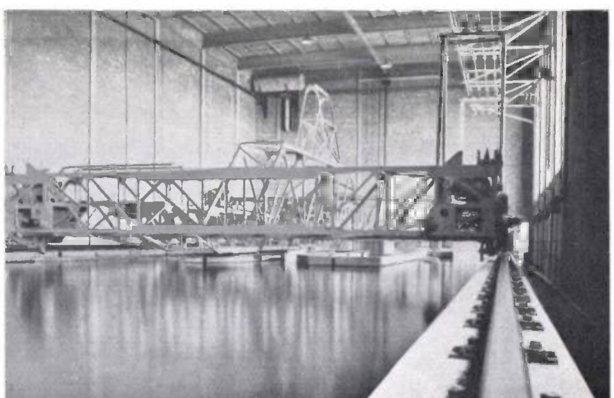

Fig. 10. Truck for towing models of ships at the testing station in Wageningen. The truck runs on two rails over a tank 10.5 m wide, 5.5 m deep and 160 m long. It is driven by four motors each of 20 H.P. fed with a voltage kept constant by means of a voltage regulator.

the battery is replaced by a Ward-Leonard converter, consisting of a direct current dynamo of 72 kW with a 1.5 kW exciting dynamo supplied from the 220 volts mains, both driven by a synchronous three-phase motor.

The terminals voltage of the direct current dynamo is kept strictly constant — at all values used between 8 and 600 volts — by a triode rapid regulator which is included in the field circuit of the exciting dynamo.

Fig. 8 show the switch board belonging to the 220 volts converter with Philips triode regulator; *fig. 9* shows the control table for the Ward-Leonard converter; the triode rapid regulator may be seen in front. Since the number of regulator triodes had to be doubled in this case, they are housed in two similar cabinets. In *fig. 10* may be seen the towing truck which runs on two rails over a tank 10.5 m wide and 160 m long.

The installation has already been working for several years. The saving obtained by the installation of converters with triode rapid regulators instead of batteries of accumulators has in this case amounted to £ 1650.—.

THE BEHAVIOUR OF AMPLIFIER VALVES AT VERY HIGH FREQUENCIES

M. J. O. STRUTT and A. VAN DER ZIEL. 537.545 : 621.3.029.6 : 621.936.645.31

The properties of amplifier valves may be characterized by four admittances (reciprocals of impedances): the slope, the output admittance, the input admittance and the reaction admittance.

These admittances are in general complex. The real part is constant at broadcasting frequencies, and can therefore be considered as a resistance; the imaginary part is proportional to the frequency and may be considered as a condenser.

With ultra short waves there are deviations from this behaviour. These deviations may be ascribed, first, to the transition times of the electrons between the electrodes of the valve, and second, to self-inductances and mutual inductances between different conductors.

In this article the behaviour is studied of the four admittances in the case of ultra short waves, and the causes of the behaviour observed in each of the four cases are discussed individually.

Introduction

The properties of an amplifier valve are in general dependent on the frequency of the alternating voltage which is amplified by the valve. For some of the quantities the dependence on frequency is obvious. This is so, for example, in the case of the capacitive conductivity between different electrodes of the valve. Other quantities, such as the slope, are, in the region of broadcasting wave lengths (> 200 m), practically independent of the frequency and exhibit deviations from this behaviour only at very high frequencies. In the latter case less obvious phenomena play a part, and it is a discussion of these phenomena and their consequences which forms the main subject of the article.

The action of a radio valve may be described for sufficiently small alternating voltages between cathode and grid, and between cathode and anode, by the following equations respectively:

$$\left.\begin{aligned} i_a &= A\,v_g + B\,v_a \\ i_g &= C\,v_g + D\,v_a \end{aligned}\right\} \quad . \quad . \quad . \quad . \quad (1)$$

The first equation expresses the fact that the current i_a at the output side (anode) of the valve has a linear relation to the alternating voltage v_g at the input side and v_a at the output side. The second equation gives an analogous relation for the current at the input side. The quantities A, B, C, D are determined by the various valve constants and have in general complex values.

We see from the equation that the quantities A to D indicate the ratio of a current to a voltage.

They are therefore reciprocals of impedances, in other words, admittances. We shall indicate them further as characteristic admittances of an amplifier valve.

The significance of the characteristic admittances may be understood by first short circuiting the output side of the valve and then repeating the process with the input side. In the first case $v_a = 0$ and

$$i_a = A\,v_g; \quad i_g = C\,v_g.$$

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$$i_a = A\,v_g; \quad i_g = C\,v_g.$$

C is called the input admittance because it gives the relation between the input current and the input voltage, and A is the slope (transconductance) of the valve.

In the second case $v_g = 0$ and

$$i_a = B\, v_a; \quad i_g = D\, v_a.$$

B is called the output admittance and D the reaction admittance.

For a systematic study of the influence of the frequency on the properties of an amplifier valve, it is best to find out how each of the four admittances depends upon the frequency. Before doing this, however, we shall first call attention to several general considerations.

Cold and hot admittances of an amplifier valve

The input terminals of an amplifier valve consist of cathode and control grid, the output terminals of cathode and anode. When the cathode is not heated, there is no direct connection between the terminals in either case, at least when we may disregard insulation leakages or subtract their influence on the admittances. When in this article we speak of a definite admittance we shall always mean *the value which remains after the insulation leakages and the dielectric losses measured on the cold valve have been subtracted.* In the case of a cold valve the admittances will in the first place be made up of capacitive conductivities between the electrodes. In the second place inductive and ohmic resistances of the leads may play a part. These impedances are in series with the capacitive impedances between the electrodes, and are comparatively much smaller at broadcasting frequencies, so that we may neglect their influence as a approximation. The substitution diagram of the cold valve [1]) is then represented by *fig. 1a.*

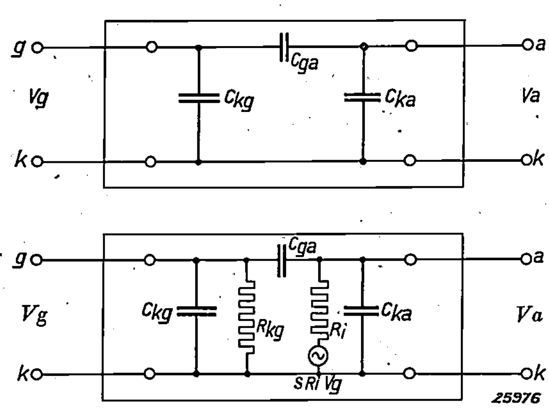

Fig. 1. a) Substitution diagram of a cold amplifier valve. The input admittance is determined by C_{kg}, the output admittance by C_{ka}, slope and reaction are given by the capacity C_{ga}.

b) Substitution diagram of a heated amplifier valve. Currents flow to the grid and to the anode, both are modulated by, and in phase, with the grid voltage. This is expressed by a resistance R_{kg} across the input terminals and a resistance R_i with an E.M.F. of the amount SR_iv_g in series across the output terminals.

When a valve is put into use, grid and anode no longer remain insulated from the cathode, but electrons can pass over from the cathode to the other electrodes. For the input and output admittances, C and B, this means that ohmic conduction is now possible, and thus certain resistances are now in parallel with the condensers. Moreover, the magnitude of the input capacity will change somewhat due to the presence of electrons in the space between cathode and grid, a fact which was noted previously in this periodical [2]). The influence on the slope A and on the reaction admittance D is more difficult to indicate by elements of an electrical circuit. Actually the reaction is scarcely changed, while the slope receives a large ohmic component S_0, due to the fact that the electron current to the anode is greatly influenced by the grid voltage in intensity and phase. The substitution diagram of the valve in the working state is reproduced and explained in fig. 1*b*.

Summarizing, we may now express the four characteristic admittances of an amplifier valve for not too high frequencies as follows:

$$\left.\begin{aligned}
&\text{Slope: } A = j\omega C_{ga} - S_0.\\
&\text{Output admittance:}\\
&\qquad B = j\omega\,(C_{ka} + C_{ga}) + 1/R_i.\\
&\text{Input admittance:}\\
&\qquad C = j\omega\,(C_{kg} + C_{ga}) + 1/R_{kg}.\\
&\text{Reaction admittance:}\\
&\qquad D = j\omega\, C_{ga}.
\end{aligned}\right\} \quad \ldots \quad (2)$$

In practice the circuit becomes somewhat simpler because of the fact that the control grid is usually made so strongly negative that practically no electrons pass between cathode and control grid. The resistance R_{kg} then becomes infinitely large and $1/R_{kg}$ cancels.

Causes and nature of the dependence of the admittances on the frequency

At wave lengths below the region from about 20 to 40 metres equations (2) for the characteristic

[1]) In this form the substitution diagram holds for a triode. It can, however, also be used for valves with more electrodes, if we assume that all the electrodes except the control grid and the anode are short circuited with respect to the cathode for alternating voltages. Then the symbols C_{kg} and C_{ka} must be taken to represent respectively the capacity between the control grid and all the electrodes connected with the cathode and the capacity between the anode and all the electrodes connected to the cathode.

[2]) C. J. Bakker. Some properties of receiving valves at short waves. Philips techn. Rev. 1, 171, 1936.

admittances are found to be no longer valid. Two reasons for this may be pointed out. One of these has already been noted. We have assumed that the inductive impedances of the leads are small with respect to the capacitive impedances between the electrodes. With increasing frequency, however, the capacitive impedances decrease and the inductive impedances increase, so that the assumption mentioned must be incorrect below a certain wavelength. With the ordinary dimensions of amplifier valves the inductive and capacitive impedances reach the same order of magnitude for wavelengths of a few metres.

The second reason is the finite transition time of the electrons. It has been shown previously [2]) in this periodical that a considerable increase in the size of the real part of the input admittance occurs for frequencies the oscillation time of which is no the transition time of the electrons between cathode and control grid. Further it is clear that in this frequency range the slope will also be affected by the transition time, because the anode current will be later in phase than the grid voltage due to the transition time, so that instead of the real ratio S_0 the ratio between the two must be a complex quantity. For the output admittance and the reaction admittance the influence of the finite transition speed is of less importance, as will be shown later.

In order to be able to deal with all these phenomena in a uniform way, we shall study the general nature of the dependence of an admittance on the frequency.

When the influence of the frequency on the admittances is determined by capacities, inductances and transition times, the frequency can appear only in the formulae in the combination $j\omega$. Except in this combination the imaginary j cannot occur in the formulae for the admittances. We may develop an arbitrary admittance Y in a Taylor series as follows:

$$Y = Y_0 + (j\omega)\, Y_1 + (j\omega)^2\, Y_2 + (j\omega)^3\, Y_3 + \dots$$

In this formula the quantities Y_0, Y_1, etc. are real. If we now divide Y into a real and an imaginary part:

$$Y\text{real} = Y_0 - \omega^2 Y_2 + \dots$$
$$Y\text{imag.} = j\omega(Y_2 - \omega^2\, Y_3 + \dots)$$

We may let the real part of the admittance be equal to the reciprocal of a resistance. The part of this reciprocal resistance which depends on the frequency is then proportional to the square of the frequency in the first approximation. The imaginary part of Y can be considered as being caused by a capacity. The part of this capacity dependent on the frequency is then in the first approximation also proportional to the square of the frequency.

This approximation represents with sufficient accuracy the behaviour of radio valves for frequencies up to 60 megacycles/sec (5 metres). In other words, when the characteristic admittances are given as in equation (2), we find that the capacities and reciprocal resistances are not constant, but each one contains a term, in addition to the constant, which is proportional to the square of the frequency.

The following is a discussion of the four admittances individually.

The slope A

The measurement of the slope is carried out by applying a low alternating voltage to the grid of the amplifier valve to be measured, and by placing between grid and anode an admittance whose real and imaginary parts are variable. By adjusting this admittance so that the alternating current in the anode circuit is zero, which is always possible, the measurement of the slope is reduced to the determination of a complex admittance. It is found that at phase angles between 0° and 90° a detuned circuit can be used as admittance. The measurement of the modulus and the phase angle of this circuit offers no difficulties.

For low frequencies the slope of a valve may be considered real. The imaginary component $j\omega C_{ga}$ (see equation (2)) is very small, especially in screen grid valves, and may be entirely neglected in comparison with the real component of the slope due to the electron current.

With short waves the electron current will need a certain amount of time to pass from the cathode to the anode. A simple estimation of the transition times may be obtained by considering the case of a homogeneous field between two flat plates a and b at a distance d from each other, which have respectively the potentials V_a and V_b with respect to the cathode. The speed v of the electron is determined at every point between the plates by the law of the conservation of energy:

$$\frac{m}{2}\, v^2 = eV.$$

The transition time of the electron between a and b follows from this:

$$t_{ab} = \int_0^d \frac{\mathrm{d}x}{v} = \sqrt{\frac{m}{2e}} \int_0^d \frac{\mathrm{d}x}{\sqrt{V}}\,.$$

Now for a homogeneous field

$$\frac{\mathrm{d}x}{d} = \frac{\mathrm{d}V}{V_b - V_a}$$

and therefore

$$t_{ab} = \sqrt{\frac{m}{2e}}\ \frac{d}{V_b - V_a} \int_{V_a}^{V_b} \frac{\mathrm{d}V}{\sqrt{V}} = \frac{2d\sqrt{\dfrac{m}{2e}}}{\sqrt{V_b} + \sqrt{V_a}}\,.$$

The value of e/m is 17.6×10^{14} cm^2 sec^{-2} volt^{-1}.

Suppose for example that $V_b = 0$, $V_a = 250$ volts, and $d = 0.5$ cm, which would about represent the condition between the suppressor grid and the anode of a pentode; one then finds:

$$t_{ab} = \frac{2{\cdot}0.5\,\sqrt{1/35.2}{\cdot}10^{-7}}{\sqrt{250}} = 1.07{\cdot}10^{-9}\ \text{sec.}$$

In this way the transition times between all the electrodes can be calculated. The total transition time t_{ka} is obtained by adding. The existence of the transition time, as has already been mentioned, results in the slope taking on a complex value at wave lengths of a few metres, where the frequency is no longer very small with respect to $1/t_{ka}$.

Instead of the real and the imaginary component, we may also consider the modulus and the phase of the slope as functions of the frequency and this description is simpler in this case because it has been found by experiment that the modulus of the slope remains practically unaltered, and the phase angle φ_t due to the transition time changes proportionally with the frequency. It is found that the transition times between different electrodes do not all contribute proportionally to the phase shift.

For a triode it may be calculated:

$$\varphi_t = -\ \omega(0.36\ t_{kg_1} + 2/3\ t_{g_1a}) \quad \cdots\cdots \quad (3a)$$

and for a pentode:

$$\varphi_t = -\ \omega(0.36\ t_{kg_1} + t_{g_1g_2} + t_{g_2g_3} + 2/3\ t_{g_3a}) \quad (3b)$$

In these formulae t_{kg_1} is the time of transition of the electrons between cathode and control grid, $t_{g_1g_2}$ the time between control grid and screen grid, $t_{g_2g_3}$ the time between screen grid and suppressor grid, t_{g_3a} the time between suppressor grid and anode.

The fact that the coefficient of ωt_{kg_1} is less than one may be explained as follows. If a sinusoidal alternating voltage v_g on the control grid would only determine how much alternating current left the cathode, the current $S_k v_g e^{-j\omega t_{kg_1}}$ would pass through the control grid, and this current is thus behind v_g in phase. But in addition the transition time varies periodically with the time, due to the alternating voltage on the voltage on the control grid, and due to this effect there is also an alternating current through the control grid. This latter current is found to be about 90° in phase ahead of the grid voltage. The result is that the resulting alternating current passing through the grid has a phase angle smaller than ωt_{kg_1}. Calculation gives $\varphi = 0.36\ \omega t_{kg_1}$.

The fact that the coefficient of ωt_{g_3a} is less than one may be explained in the following way. When an electron passes through the third grid, there is an equal positive image charge on the third grid. When the electron moves towards the anode, this image charge flows gradually from the third grid to the anode, so that when the electron arrives on the anode the whole image charge has also just arrived there. In the anode lead, therefore, current is flowing as long as the electron is between the third grid and the anode. If current only flowed in the anode lead when the electron had arrived on the anode, the coefficient of ωt_{g_3a} would have the value one, but since current is flowing as long as the electron is on the way from the third grid to the anode, the coefficient will be smaller than one. An accurate calculation shows that the coefficient must have the value $^2/_3$ when the electron passes the third grid at a low speed. Finally, in order to understand why in the equation for the pentode the terms $\omega t_{g_1g_2}$ and $\omega t_{g_2g_3}$ appear with the coefficient one, we must assume that the electrodes shield each other so well that the current excited by a change in the control grid voltage only begins to flow in the anode lead when the electron is between the suppressor grid and the anode. The transition time between control grid and suppressor grid is then entirely manifested in the phase shift.

The transition time, however, is not the only cause of a phase shift between the anode current and the grid voltage. It is obvious that the self-inductance of the cathode lead also has a similar effect. When there is an alternating voltage v_g between the grid and the external terminal of the cathode, the following current flows through the cathode lead:

$$i_k = S_k\, v_g.$$

This current experiences an inductive reactance in the lead which corresponds to a inverse e.m.f.

$$i_k\, j\omega\, L_k = j\omega\ S_k L_k v_g.$$

The voltage between the grid and the cathode itself thus becomes

$$v_g' = v_g\ (1 - j\omega\ S_k L_k) \quad \cdots\cdots \quad (4)$$

and when $\omega S_k L_k \ll 1$, this means that the anode current lags a phase angle $\omega S_k L_k$ behind the external applied modulation voltage.

The total phase shift is obtained in fair approximation by adding the two results. Thus for a triode:

$$\varphi = -\ \omega\ S_k L_k - \omega\ (0.36\ t_{kg_1} + 2/3\ t_{g_1a}) \quad \cdot \quad (5a)$$

and for a pentode:

$$\varphi = -\ \omega\ S_k L_k - \omega\ (0.36\ t_{kg_1} + t_{g_1g_2} + t_{g_2g_3} + 2/3\ t_{g_3a}) \quad \cdot\ \cdot \quad (5b)$$

Equations (5a) and (5b) were tested by measurements on a pentode of the type AF7. The values measured at a wave length of 9.1 m are:

as pentode with anode current of 3 mA:

$$|S| = 2.69 \text{ mA/V}; \ \varphi = -22°;$$

as triode with 4 mA anode current:

$$|S| = 3.82 \text{ mA/V}; \ \varphi = -6°.$$

In the measurement as a triode, anode, suppressor grid and screen grid were interconnected; in the measurement as a pentode, screen grid and suppressor grid were connected to the cathode through large condensers. For a very low frequency we found $S_0 = 2.7$ mA/V for the pentode connection, and $S_0 = 3.8$ mA/V for the triode connection.

It may be noted in the first place that the modulus of the slope at 9.1 m is equal to the static value within the limits of experimental error. The phase angles measured must now be compared with the values which can be calculated from equations (5a) and (5b). We found:
$S_k = 3.8$ mA/V and L_k about 5×10^{-8} H, and the following values:

Triode		Pentode	
t_{kg_1}	=	t_{kg_1}	$= 0.77 \cdot 10^{-9}$ sec,
t_{ga}	=	$t_{g_1g_2}$	$= 0.26 \cdot 10^{-9}$ sec,
		$t_{g_2g_3}$	$= 0.49 \cdot 10^{-9}$ sec,
		t_{g_3a}	$= 0.75 \cdot 10^{-9}$ sec,
S_kL_k	=	S_kL_k	$= 0.19 \cdot 10^{-9}$ sec.

From this it follows at 9.1 m wave length:

	measured	calculated
as triode	— 6°	— 7.5°
as pentode	— 22°	— 20.5°

This good agreement was also found with measurements on other valves, so that the explanation given for the phase angle of the slope may be considered complete. At the same time this supplies proof that the dependence on frequency of the phase angle of the slope is well represented by formulae (5a) and (5b).

The significance of the phase angle of the slope

The phase angle of the slope is important in all those cases in which a portion of the amplified voltage is fed back to the input circuit in order to diminish its damping or to excite oscillations. If the feed back voltage is not in phase with the input voltage, not only is the damping of the input circuit affected by the feed back, but also its resonance frequency. A large phase angle of the slope in the case of an oscillator may therefore unfavourably affect the frequency stability with variations of the supply voltage.

The phase angle of the slope may also be important in cases where feed back is not deliberately employed, for example, in high frequency amplifiers. Due to the capacity between control grid and anode an (undesired) back coupling may also occur in this case. The phase angle of the slope can then increase the tendency to oscillation of a high frequency stage.

The output admittance B

The measurement of the output admittance is in principle as follows. The output side of a valve (cathode, anode) is connected in parallel to a circuit whose resonant impedance is known. The circuit is then detuned by the capacitive part, and damped by the ohmic part of the output admittance. From the detuning the output capacity can be determined, and by tuning the circuit anew, and determining the new damping, the output parallel resistance is obtained.

As a result of such determinations it is found that the output capacity hardly changes with the frequency, while the real part of the output admittance

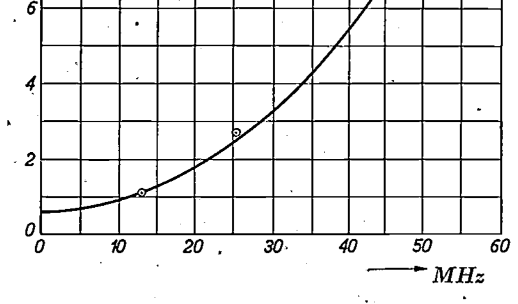

Fig. 2. The real part of the output admittance of a valve EF6 plotted against the frequency. Points measured. The curve drawn is given by the formula:
$\frac{1}{R_i} = (0.60 + 0.30 \cdot 10^{-2} \nu^2)$ micrombos, where ν is in megacycles per sec.
It may be seen that the points fall very well on the second order curve. For low frequencies E is equal to the internal resistance of the valve which is of the order of magnitude of 2 MΩ.

mittance as a function of the frequency above 10 megacycles/sec (30 m) increases considerably, namely proportionally to ω^2, as may be seen from the results of measurement on a pentode EF6 given in fig. 2. The output parallel resistance of the valve EF6 is 2 MΩ at low frequency, and at 6 megacycles/sec (5 m) it is only 5000 Ω.

In order to find out whether the transition times of the electrons in valves of a type like EF6 are important for the output admittance on short waves, we shall consider what happens in the space between suppressor grid and anode. We apply a low alternating voltage v_a to the anode, and wish to determine the alternating current i_a to the anode caused hereby. When we consider the internal resistance as infinitely great, only direct current passes through the suppressor grid. Between the suppressor grid and the anode, however, an alternating current is found to act because of the fact that the transition time varies with time. It is clear that the admittance i_a/v_a corresponding to this is proportional to the direct current passing through the suppressor grid. If thus the admittance contains a real part, and therefore represents a damping, this must also be proportional to the current.

This has been tested by experiment. The dimensions of the third grid and of the anode are exactly the same in the pentodes AF3 and AF7. The output damping of the two valves also varies only slightly. At 8.0 m the output damping of AF3 is 7.7 micromhos, and of AF7 is 8.7 micromhos. The difference between the anode currents, however, is much greater. With AF3 the anode current at the working point is 8 mA, and with AF7 only 3 mA. From this it may be seen that the transition time of the electrons between the third grid and the anode cannot play an important part in the anode parallel resistance.

There must, therefore, be another cause. The conductance which corresponds to the statically measured internal resistance is only about 0.5 micromhos for both valves, and thus need not be considered. The greater part of the output damping comes from the self-inductances and mutual inductances of the leads and the capacities between the electrodes of the valve. If an alternating voltage is applied to the anode, alternating currents flow through the capacities between anode and suppressor grid, anode and screen grid, etc. to screen grid and suppressor grid, and then through the leads to earth. Due to the mutual induction of these leads with the cathode connection, a slight alternating voltage is caused between the cathode lead in the valve and the cathode terminal, and this alternating voltage also acts between control grid and cathode. In an amplified form it is returned to the anode in the valve, and this may cause damping. Measurements of this effect have been carried out with an AF3 valve in which the capacity C_{g_3a} between anode and third grid, and the mutual inductance M_{kg_3} between the connections to third grid and to cathode are principally concerned. From the above reasoning the following formula may be set up:

$$\frac{1}{R} = \omega^2 \, S \, C_{g_3a} \, M_{kg_3} \, ,$$

where S is the slope of the tube.

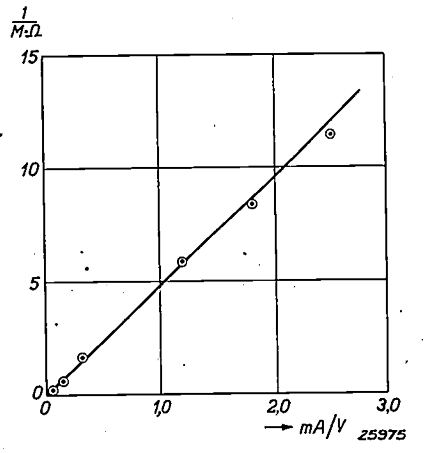

Fig. 3. Values of $1/R$ as a function of the slope S in mA/V for the valve AF3. As required by the theory the measured points (circles) lie on a straight line. Wave length 8.0 m.

In fig. 3 the values of $1/R$ for the valve AF3, measured at a wavelength of 8.0 m, are plotted against the slope S of the valve. From the linear relation between $1/R$ and S, and under the reasonable assumption that $C_{g_3a} = 3.5\ \mu\mu\text{F}$, the value 24×10^{-9} H follows for M_{kg_3}. This is a probable order of magnitude. This also explains why valves AF3 and AF7 have approximately equal anode parallel resistances, because they have the same values of C_{g_3a} and M_{kg_3}, while the slopes also differ very little at the working points.

The input admittance C

The measurement of the input admittance is carried out on the same principle as that used for measuring the output admittance.

Fig. 4 gives the result of such a determination. The real part of the input admittance is plotted in double logarithmic coordinates against the frequency. This curve shows that the input parallel resistance so defined varies proportionally with the square of the wave length up to very short waves, which proves that higher powers of ω play practically no part in the development of the series.

The real part of the input admittance may here also be ascribed to two essentially different causes, and consists of two terms $1/R_t$ and $1/R_{LC}$, which come respectively from the transition time of the

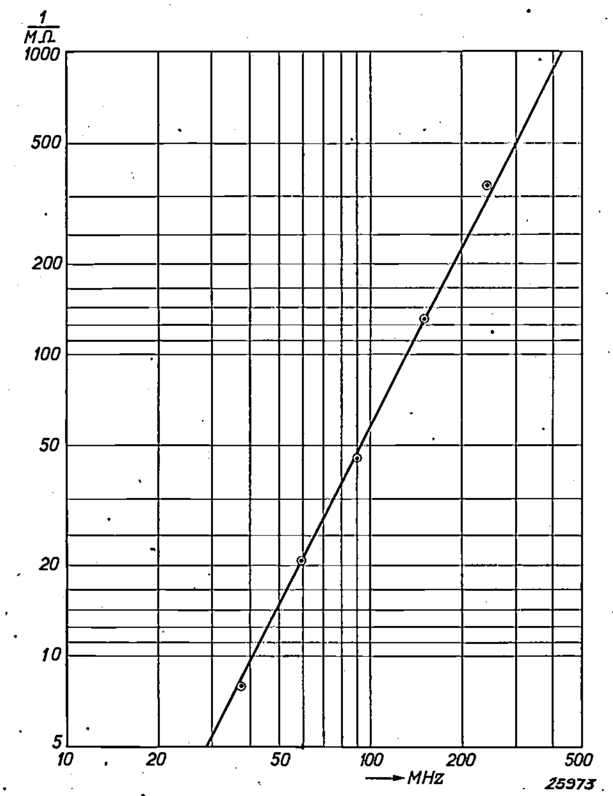

Fig. 4. Input parallel resistance of a pentode as a function of the frequency. The frequency in megacycles/sec is plotted horizontally, and the input damping in micromhos vertically. It may be seen that the input parallel resistance varies as the square of the frequency.

electrons, and from the inductances in the valve together with the capacity between grid and cathode:

$$\frac{1}{R_{kg}} = \frac{1}{R_t} + \frac{1}{R_{LC}}.$$

For the transition time part, the following formula has already been derived in this periodical [2]):

$$\frac{1}{R_t} = \frac{1}{20} f \cdot S_k \cdot (\omega t_{kg})^2 \quad . \quad . \quad . \quad . \quad (6)$$

In this formula t_{kg} is the transition time of the electrons between cathode and control grid, and f is a factor, usually of the order of 2, which depends upon the transition time of the electrons between control grid and anode (or screen grid, when we are concerned with tetrodes or pentodes).

Upon consideration of the influence of the different inductances on the real input admittances it is found that the self inductance L_k of the cathode lead plays the most important part. For the contribution to $1/R_{kg}$ of this self inductance, the following formula is found:

$$\frac{1}{R_{LC}} = \omega^2 S_k L_k C_{kg_1}, \quad . \quad . \quad . \quad . \quad . \quad (7)$$

where C_{kg_1} is the capacity between control grid and cathode.

Equation (7) may be derived as follows. As showns above (equation (4)), the voltage between cathode and grid differs in phase from the external applied modulation voltage and is:

$$v_g' = v_g\,(1 - j\omega S_k L_k). \quad . \quad . \quad . \quad . \quad (4)$$

This voltage is connected with the capacity C_{kg_1} between grid and cathode. The current to the grid is therefore

$$i_g = v_g\,(1 - j\omega S_k L_k)\, j\omega C_{kg_1},$$

from which it follows that an input admittance is given of

$$\frac{i_g}{v_g} = j\omega C_{kg_1} + \omega^2 S_k L_k\, C_{kg_1},$$

in agreement with equation (7).

Separation of the influences of transition times and inductances on the input admittance

Formulae (6) and (7) for the input parallel resistance caused by the transition times and inductances correspond closely. In order to demonstrate this agreement even more clearly let us consider the oscillation time T_{kg_1} of the alternating current for which the self-induction of the cathode lead and the capacity between cathode and grid are in resonance. This time of oscillation is given by:

$$T_{kg_1} = 2\,\pi \sqrt{L_k\, C_{kg_1}}$$

and by inserting this value into equation (7) it follows that

$$\frac{1}{R_{LC}} = \frac{1}{4\,\pi^2}\, S_k\, (\omega T_{kg_1})^2,$$

which formally resembles very much the transition time effect:

$$\frac{1}{R_t} = \frac{f}{20}\, S_k\, (\omega t_{kg_1})^2.$$

As already mentioned the resonances in the valve usually appear at wave lengths between 1 and 2 m, therefore with oscillation times of 3 to 6×10^{-9} sec. Now the transition times of the electrons between cathode and grid are just of the same order of magnitude so that the two causes of damping will be practically equally important.

An experimental separation of the two influences on the input damping is possible by comparing the damping of a single system with that of two

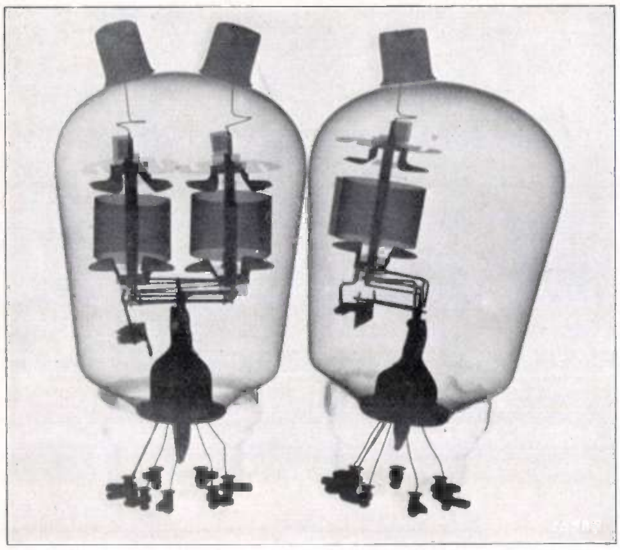

Fig. 5. X-ray photograph of a single and a double pentode (type EF5) for separating the two causes of the input damping on short waves.

systems connected in parallel with common leads (see X-ray photographs *fig. 5*).

Upon doubling the valve, the slope is doubled and so is the $1/R_t$ value, since the other quantities in equation (6) remain unaltered. The induction part (equation (7)) is, however, quadrupled, because it is not only proportional to the slope, but also to the capacity C_{kg_1} which is also doubled. If we call the input parallel resistance of the single and of the double valve R_e and R_d respectively, then:

$$\frac{1}{R_e} = \frac{1}{R_{LC}} + \frac{1}{R_t},$$

$$\frac{1}{R_d} = \frac{4}{R_{LC}} + \frac{2}{R_t},$$

and from this follow the two input parallel resistances R_{LC} and R_t:

$$\frac{1}{R_{LC}} = \frac{1}{2R_d} - \frac{1}{R_e},$$

$$\frac{1}{R_t} = \frac{2}{R_e} - \frac{1}{2R_d}.$$

Table I gives some results, obtained in this way.

Table I. Measurements of the input damping of three test valves at a wave length of 7 m.

Valve No.	distances between electrodes k-g_1	g_1-g_2	S_k	R_t	R_{LC}	$\frac{1}{R_{LC}} : \frac{1}{R_t}$
	mm	mm	mA/V	Ω	Ω	
1	0.10	0.30	18	60000	600	100
2	0.20	0.76	2.3	27000	46000	0.60
3	0.35	2.10	1.9	11000	39000	0.28

With increasing distance between the electrodes the share of the transition times in the total damping becomes steadily greater because the transition time increases and the capacity decreases.

Valve No. 1 has in connection with its very steep slope an unusually long cathode which has a relatively great capacity with respect to the control grid; for this reason, and due to the steepness of the slope, the damping resistance R_{LC} is much smaller here than in ordinary amplifier valves.

Significance of the input impedance

In the case of an amplifier stage, in which a tuned circuit is included in the grid connection, the input parallel resistance acts across the circuit, which is thereby damped and the amplification may be considerably decreased. When several similar amplifier stages are connected in series, the input resistance of the following valve always acts across the circuit connected to the anode. The output resistance of the valve is connected in parallel with this. This latter resistance, however, is very much greater than the input resistance at all frequencies, so that it has practically no influence on the damping. The greatest possible resistance in the anode connection is thus the input parallel resistance R_{kg}, and the greatest possible anode alternating voltage thus becomes:

$$v_a = i_a R_{kg} = v_g S R_{kg}.$$

At very high frequencies the product SR_{kg} may become less than one; no amplification is then possible. This shows how important it is to have the input resistance as high as possible. Because of this for very short waves valves have been designed with very small dimensions and very short connections to the different electrodes. In this way the influence not only of transition times but also of induction is avoided as much as possible. The measurements shown in fig. 4 refer to such a valve. Considering the fact that this valve has a slope of 1.5 mA/V, it follows from fig. 4 that at a wave length of 1 m an amplification by a factor 2 can still be attained.

The reaction admittance D

The reaction admittance can be measured by applying a known alternating voltage to the anode, and connecting a tuned circuit of known impedance between control grid and cathode. From the measurement of the alternating voltage across this circuit the reaction admittance can be derived.

Measurements of the reaction admittance show no important difference between a cold and a warm tube. This shows that the electrons are scarcely concerned with this admittance. It is found more over that this admittance remains purely imaginary even at very high frequencies, and in agreement with our general considerations it can be represented by

$$D = j\omega C_{ga}' = j\omega (C_{ga} - K\,\omega^2), \quad \cdot \cdot \cdot \quad (8)$$

where K is usually positive. K depends upon the self- and mutual inductions of the leads, and upon the valve capacities.

In *fig. 6* the value of C_{ga}' measured for a pentode AF3 is plotted against the frequency. A curve of the second order has been drawn as well as possible through the points measured; it may be seen that the agreement is good.

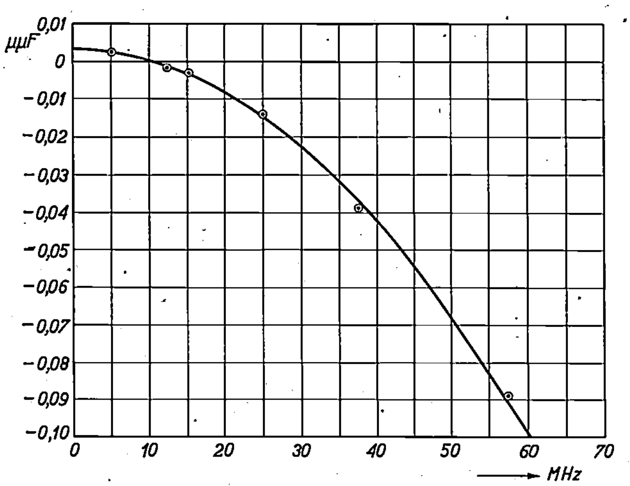

Fig. 6. The capacity C_{ga} of an AF3 valve plotted as a function of the frequency. Points measured. Curve drawn is given by the formula: $C_{ga}' = (0.0030 - 0.285 \times 10^{-4}\nu^2)$ μμF, where ν is in megacycles per sec. It may be seen that the measured points fall satisfactorily on the second order curve in agreement with formula (8)).

The variable part of the reaction admittance may be explained as follows. When a high frequency alternating voltage v_a acts on the anode, currents flow to the different electrodes through the capacities C_a of the valve. Due to the self inductances and mutual inductances of the leads, alternating voltages are found to act on these electrodes, and due to the capacity C_g between these electrodes and the grid, currents flow to the grid. If we call the average inductance (self-induction or mutual induction) of the leads concerned in this process M, the current may to the first approximation be written as follows:

$$i_g = v_a \cdot j\omega C_a \cdot j\omega M \cdot j\omega C_g = v_a \cdot j\omega\, (-C_a M\, C_g \omega^2),$$

which corresponds to equation (8).

That this representation is correct was proved by the fact that it was possible, by introducing a suitably chosen mutual inductance between the connections to the screen grid and to the anode, to compensate the term $K\omega^2$ completely. The arrangement of the circuit for this purpose is drawn in *fig. 7.*

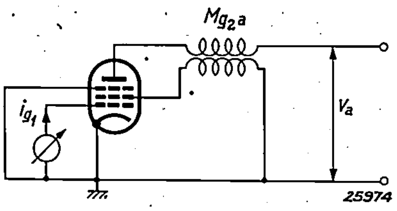

Fig. 7. Compensation of the reaction on ultrashort waves by means of a mutual induction M_{g_2a} between anode lead and screen grid lead. M_{g_2a} is chosen so that the ω^2 term in formula (8) is entirely compensated.

Equation (8) is an approximation which is only valid outside the region where resonances in the valve become pronounced. In this region the reaction admittance may become very large. In order to make this plain a substitution diagram is given in *fig. 8* for the case in which the self-inductance L_{g_2} of the connection to the screen grid is in resonance with the capacity C_{g_2} between the screen grid and the other electrodes. In this case a very high alternating voltage is found to act on the screen grid, so that, due to the capacity between screen grid and control grid, a very high current will flow to the control grid.

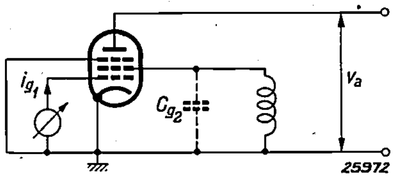

Fig. 8. The appearance of resonances in the valve at high frequencies. The self inductance L_{g_2} of the screen grid lead is tuned with the capacity C_{g_2} of the screen grid to the frequency of v_a. This is expressed in a strong current i_{g_1}, i.e., in a high conductivity from anode to the first grid.

The reaction admittance is undesirable in a high frequency amplifier valve and must be kept as small as possible. It is by the reaction admittance that a part of the amplified voltage is conducted back to the grid. If this back coupling becomes too great, oscillation can occur in certain circumstances.

A CAR RADIO

by J. W. ALEXANDER.

621.396.62 : 629.113

The most important demands made on a motor car radio set are: small dimensions, high sensitivity, shielding against external interferences, ease of installation, very great ability to withstand shocks, low current consumption and large output of sound. In this article it is shown how these partially conflicting requirements can be satisfied by constructing a set in which use is made of parts especially designed for this purpose, including such as radio valves, loudspeaker and arrangement for operation.

Introduction

In recent years there has been a steadily increasing interest in radio sets for the motor car. These receiving sets are built specially for use in a car and are capable of reproducing the programmes of the ordinary broadcasting stations. A receiving set such as that ordinarily used at home cannot be used in a car for various reasons. The circumstances under which the sets work, the consequent difficulties and their solutions will be discussed in the following, from which it will also appear why a car radio set differs so much from a radio set for use at home.

In geneıal there may be said to be five points which exert an influence on the construction of car radio sets. These points are connected with:
1) Aerial,
2) Loudspeaker,
3) Supply,
4) Interferences,
5) Mechanical construction.

The aerial

An aerial for a car radio set for obvious reasons can only have a limited height in comparison with an ordinary house aerial; because of this the voltage excited in the aerial by the transmitter field will be relatively small. The effective height (cf. Philips techn. Rev. 2, 216, 1937), which in the case of a house aerial is about 1 m on the average, is only a few centimetres in the case of a car radio. The result is that the sensitivity of the motorcar set must be several times as great as that of a house set, in order to be able to receive ordinary broadcasting stations satisfactorily.

In addition, the signal entering the set will be further considerably weakened due to the cable connection between aerial and set. This connection must be well shielded in order to prevent the electrical equipment of the car from interfering with reception. If the aerial has a capacity C_a to earth, and the connection cable a capacity C_k, a voltage occurs between the input terminals of the set which is proportional to $\frac{C_a}{C_a + C_k}$. From this it may be seen that the capacity of the cable must be as small as possible and the capacity of the aerial as large as possible.

However, the development of the motorcar aerial has proceeded in a direction such that the capacity of the aerial has become steadily smaller. Formerly the aerial usually took the form of a strip of wire gauze introduced under the roof covering. These aerials had a capacity of about 150 cm and higher. The introduction of the steel top has made this form of aerial impossible, and various other forms have appeared, such as the aerial under the car in the form of a rod or wire, the aerial on the top, and at present also the aerials of constant or variable length in the form of rods which are fastened to the side of the car or on the bumper. The capacity of these aerials is only of the order of 30 cm, while the capacity of a house aerial is 200 cm on the average.

Due to the low intensity of the signal received it is necessary as was stated above, to increase the sensitivity of the receiving set many times. The sensitivity of a receiving set is defined in the following way. A high frequency voltage with a low frequency modulation of 30 per cent and 400 cycles acts in the aerial circuit. This voltage is transformed by the apparatus into low frequency energy, which, for measurement of sensitivity, is not conducted to the loudspeaker, but to a suitably chosen ohmic resistance. 1 W is usually chosen as the normal value of this low frequency power for automobile radios. The sensitivity of the set is then the number of microvolts necessary in the aerial to obtain this low frequency energy of 1 W in the above-mentioned output resistance. For automobile radio sets the sensitivity is of the order of some microvolts; for a house radio for the same low frequency energy a signal of 50 to 100 microvolts in the aerial is necessary, so that it may be said that an automobile radio set is about 20 times more sensitive than an average house set.

The increase of the sensitivity is limited by the noise occurring in the receiver itself (see Philips techn. Rev. 2, 136, 1937). As is shown in the article quoted, because of the thermal fluctuation of the electricity in a conductor with resistance R, a voltage appears such as that given by the following formula:

$$\overline{V^2} = 4\ kTR\ \Delta\nu,$$

where $\overline{V^2}$ represents the contribution to the mean square of the voltage for the frequency interval between ν and $\nu + \Delta\nu$, k Boltzmann's constant and T the absolute temperature. A tuned LCR circuit may in general, in the frequency region in the neighbourhood of its tuning point, be considered as a resistance of the value L/CR. For the width of the frequency region around the tuning frequency, which is amplified by the apparatus, we may take as an approximation 6000 c/s. The value R/L, which determines the shape of the resonance curve of the input circuit, is, under normal circumstances, 120 000 c/s. At a value of C of 50 $\mu\mu$F of the tuning condenser, and for a temperature of 300° K we therefore find:

$$\overline{V^2} = 4 \cdot 1.37 \cdot 10^{-23} \cdot 300 \cdot \frac{1}{50 \cdot 10^{-12} \cdot 12 \cdot 10^4} \cdot 6 \cdot 10^3$$

$$\overline{V^2} = 16 \cdot 10^{-12},$$

so that the square root of the mean square of voltage is:

$$\sqrt{\overline{V^2}} = 4\ \mu\text{V}.$$

The noise thus gives as much low frequency energy as a signal voltage on the first circuit of 12 μV with an average modulation of 30 per cent.

From this it follows that there will be a signal strength below which the noise becomes disturbing; at higher signal strengths the ratio of signal to noise is more favourable. This noise effect can appear sooner in a car radio set than in a house set, because the car set is extremely sensitive and the circumstances can change so completely while the car is moving. When the car is in the open, the transmitter field will be much stronger than when the car is passing through a narrow street between high buildings, or, even worse, when it is within the metal structure of a bridge where it is practically surrounded by a Faraday cage. This difference in field strength is practically removed by the automatic volume regulation which provides that the amplification of the apparatus shall depend upon the strength of the signal received. The amplification is greatest with the weakest signal, and decreased with increasing strength of the signal. In this way the acoustic energy delivered remains almost constant. When the circumstances are very unfavourable (Faraday cage) the amplification becomes higher, so that the noise effect begins to be clearly heard.

The loudspeaker

In a moving motorcar the loud speaker must give more output than in a room in a house. While in the room the music is sufficiently loud when the level of sound is 50 phons (see Philips techn. Rev. 2, 54, 1937), it can scarcely be appreciated in an automobile whose own level of sound varies between 45 and 65 phons. Therefore a much higher level is necessary depending on the construction of the car, the running speed, the wind, the nature of the road surface, etc.

In addition to the necessity of establishing a high level of sound because of the presence of other sounds, the absorption of sound inside the car also plays a part. In this factor, as is known from acoustics, the upholstery of the car is very important, especially with respect to the absorption of high tones. Further the degree of absorption depends upon the number of persons in the car, while the great influence of open windows, especially during running, will immediately be clear.

Thus for a car radio a loudspeaker is necessary which, in spite of its small dimensions, has a high acoustic yield. This is achieved by making a small magnet of a special kind of steel ("Ticonal"), which gives a very high field strength.

Power Supply

Like all other radio sets the car radio needs various different voltages, namely cathode, anode and grid voltage. As a source of energy in the automobile there is the accumulator for the ignition and the other electrical equipment. The cathodes can be connected directly to the accumulator, a much higher voltage than 12 volts (6 volts in some cars) is necessary for the anode voltage, so that an arrangement is necessary for transforming the direct voltage of 12 volts into a direct voltage of about 250 volts. In the beginning of car radio development use was made of separate batteries for the anode voltage, later rotating converters were also used. These installations, which occupied much space and were very expensive, have been practically superseded by the vibrator converter, the principle of which has already been described in this periodical (see Philips techn. Rev. 2, 346, 1937). In the case there discussed, however, a direct voltage of

100 - 250 volts had to be converted into 100 - 250 volts alternating voltage. In the converter for the car radio the construction is some what different, since one begins here with a direct voltage of 6 volts, while the energy which must be handled by the vibrator is much smaller than in the case considered previously. The principle is given in *fig. 1*. By means of an electromagnet (M) a spring

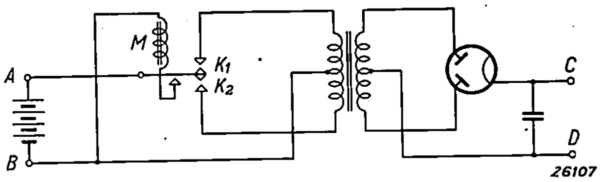

Fig. 1. Diagram of the supply of a motorcar radio set by means of a vibrator converter. The 6 or 12 volts accumulator is connected between the terminals A and B, while the anode voltage of about 250 volts is taken off between C and D. M is the electromagnet which keeps the vibrator moving; the contacts K_1 and K_2 are alternately made and broken.

is kept vibrating, and alternating closing contacts (K_1 and K_2). A direct current flows for a time (as long as the contact is closed) alternately through the two halves of the primary winding of the transformer. These currents cause alternately in the two halves of the secondary winding of the transformer, transformed currents which are rectified and smoothed by means of the rectifier valve and the condenser, until there is a direct current of about 250 volts between the terminals C and D.

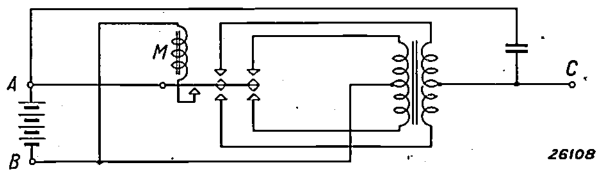

Fig. 2. Construction of the vibrator converter with a second set of contacts, whereby the current direction is reversed simultaneously in the primary and in the secondary winding of the transformer. This makes the use of a rectifier unnecessary. The anode voltage is taken off between the terminals A and C.

In the newest type (*fig. 2*) the rectifier valve is replaced by an extra set of contacts on the vibrator, which are in the secondary circuit and can be moved synchronously with the contacts in the primary circuit. The anode voltage is thereby obtained between the points A and C. This construction has the advantage of greater efficiency and of needing less space because of the omission of the rectifier valve. The whole apparatus thus becomes smaller.

Accumulator supply means in the end that the dynamo which changes the accumulator is more heavily loaded by the car radio. The current consumption of the set must therefore be kept as low as possible. One way of doing this is by keeping the consumption of heating current low.

In addition to this requirement, the valves for the car radio set must also satisfy several other requirements, which together have laid the foundations for a whole new series of valves, the so-called E-series.

These valves are intended for a supply voltage of 6 volts. This is, however, not a constant voltage, but in unfavourable cases may run up to 8 volts, depending on the state of the accumulator, on the dynamo and on the running speed. Because of this, in the supply systems described, not only the cathode voltage but also the anode and grid voltages increase so that the valves are extra heavily loaded. This has been taken into account in the construction of the valves.

Care must not only be taken to provide for a low heating current consumption (with these valves 200 mA), but also for small dimensions. This requirement is satisfied by a very compact construction; the total height of these valves is 90 mm and the greatest diameter 32 mm. In addition by means of careful construction, such as the introduction of extra supports, care has been taken that the valves are very resistant to the shocks and vibrations which they experience in the set while the automobile is running.

Interferences

A very important difference between a radio set for use at home and a motorcar is that the latter is situated in the immediate neighbourhood of a very strong source of interference. These interferences are caused by the ignition system, the ignition works on the following principle (*fig. 3*).

The accumulator is connected via the primary winding of a transformer T (the ignition coil) in series with an intermittent contact K. Due to

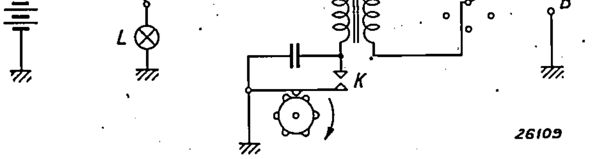

Fig. 3. Diagram of the ignition system. K is the "make and break" contact. The transformer T is the ignition coil, and V is the distributor through which the current is led to one of the sparking plugs B. L is the lighting circuit of the car which also is connected to the accumulator.

this, on the secondary high voltage side of the transformer, surges of the order of 10 000 volts occur, which are conducted over a distributor V successively to the different sparking plugs B. There are thus sudden variations in voltage in the primary as well as in the secondary circuit. If the variation in voltage is divided up into a Fourier frequency spectrum all frequencies are found to be present. With increasing frequency the amplitude increases up to about 30×10^6 cycles per sec and decreases again for higher frequencies. These voltages are not only present on the high and low voltage connections of the ignition system, but since the latter are connected with the accumulator, they also act on all cables connected with it, such as those for the lights (L). Not only do these voltages act on the connections, but also on all other metal parts such for example as the chassis, since this is used as return connection. To reduce these interferences it was customary to connect resistances of the order of 10 000 ohms in the sparking plug line. This decreases the interference very much. However, in the design of the ignition system no account had been taken of a possible later introduction of these resistances, and difficulties arose. The method of complete shielding has also been applied, in which the whole ignition system was placed in an electrically closed space. This is very expensive and complicated, since practically every kind of automobile is constructed differently.

A closer study of the interference has indicated a very much simpler solution: When interferences occur in the case of an ordinary radio, they usually appear in the following manner. The set when tuned to a definite transmitter, is sensitive practically only to voltages of the same frequency as the transmitter. In general an interference will include a whole frequency range. The apparatus will pick out only that frequency to which it is tuned, and this will cause a voltage on the first circuit which is amplified further by the apparatus just like a signal. If we now connect a tuned circuit in series between aerial and earth, it will only be able to diminish the interference at the signal frequency; thus the signal disappears together with the interference.

In a car radio set the situation is more complicated. If here we connect a tuned circuit in parallel with the set the signal disappears but the interference continues. The interference thus occurs because of frequencies other than that of the signal. This is possible here because:

1) the distance between the source of interference and the receiver is so small,
2) the source of interference is so unusually strong.

In a car radio voltages with frequencies greater than those of the broadcasting band enter the receiver with high intensity and then cause interferences due to non-linear phenomena.

For removing such interferences, therefore, the set must be so constructed that the frequencies outside the broadcasting range cannot penetrate into it, while an extra provision is necessary for the frequencies within the broadcasting range. The first is attained by:

1) Making the apparatus itself electrically completely closed (apart from holes for ventilation). This is done by giving the set a metal jacket and closing the necessarily removable parts (for instance for the changing of valves) with springs,
2) Providing the connections entering the set, such as aerial and supply connections, with filters, so that the aerial connection only passes frequencies of the broadcasting range.

In order to keep out the remaining interferences with frequencies in the broadcasting range, it is usually sufficient to introduce a condenser between the chassis and the accumulator side of the ignition coil. Further, in installing the set, care must be taken that the aerial is as far as possible away from interference-bearing conductors, such as for example lamp wiring and battery. In this respect an aerial on the top is to be preferred to one under the car.

Although the ignition system causes the most serious interferences, it is not the only cause of interference. The dynamo for charging the battery also gives appreciable interferences. The introduction of a condenser between the dynamo and the chassis gives satisfactory improvement.

There is an additional group of interferences which do not always appear but which may nevertheless sometimes be very disturbing. The so-called "axle interferences" belong to this group of specific motorcar radio interferences. In a car with rear drive the front wheels turn about an axle which is directly connected with the chassis. Between the axle and the bearing of the wheel there is a thin layer of grease which insulates the wheel from the axle. Especially with older cars there is a possibility that, depending on the shocks experienced by the wheel, more or less good electrical contact will be made between wheel and axle at this point. Crackling interference noises result from this varying contact. The remedy is either a permanent insulation, or a permanent contact which can be obtained by means of a pressure spring between the revolving hub of the wheel and the stationary extremity of the axle. In automobiles with front-wheel drive it is obvious that these springs must be introduced on the rear wheels.

Another incidental interference, the "brake interference" may occur due to the generation of static electricity by the friction of a brake shoe on a brake drum. The remedy for this is the correct adjustment of the brake shoes.

In some cases a "tyre interference" may be experienced which appears to be due to frictional phenomena between the tyre and the road. This interference depends upon the nature of the road surface and appears chiefly with particular kinds of tyres. On a macadam road it is much less disturbing than on concrete roads. Better earthing of the wheel by means of the above-described spring sometimes brings improvement. For this group of interferences also an aerial is to be preferred which is as far as possible away from the source of interference, i. e., on top of the car or on the side.

Mechanical construction

In its mechanical construction a radio set also differs to an important degree from a house set. The requirements made of an auto- mobile radio in this respect are of course quite different. We shall go into several of these:

1) Since the apparatus is installed in the motorcar at a point where the space is very limited as a rule, it must be made as small as possible (*fig.* 4). This is done by using the smallest possible number of parts, each of which is as small as possible, and by placing them as close together as possible, as may be seen from the photograph in which one of the side walls has been cut away (*fig.* 5). In the middle of the foreground two valves have been removed in order to show the part of the apparatus lying behind them. As may be seen from the photograph, the loudspeaker is built into the apparatus. The least amount of space is occupied in this way, and in most cases this apparatus can easily be installed under the dashboard. If still better quality is desired, or when it is necessary to set up the loudspeaker in some other place, as may be the case in a motorbus, a model with a separate loud speaker (*fig.* 6) may be chosen. The loud speaker of this type is larger and more sensitive than is possible with a loud speaker built into the set. For a motorbusses there is also the possibility of connecting a microphone to the set. By means of a reversing switch the conductor may then interrupt the radio broadcast and use the set and the loudspeaker hanging above the driver in order to address the occupants of the bus.

2) The apparatus must be able to be operated from a particular seat, for example from the driver's seat. This is only possible when the parts necessary for operation are fastened on or under

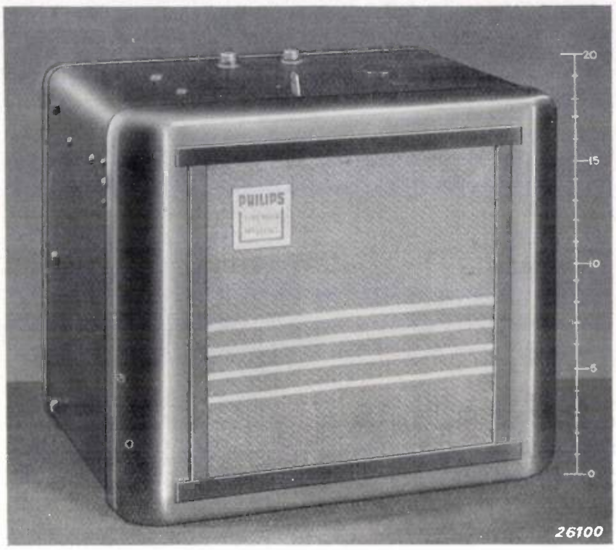

Fig. 4. A car radio set which is only 18,5 cm high.

the dashboard or on the column of the steering wheel. Since, however, there is usually no space available at these points for installing the set itself, it is necessary to introduce remote control. The parts necessary for operation are then assembled in an operating box (*fig.* 7) which can be introduced under (*fig.* 8) or, by means of another housing, in the dashboard (*fig.* 9). The right-hand knob of the

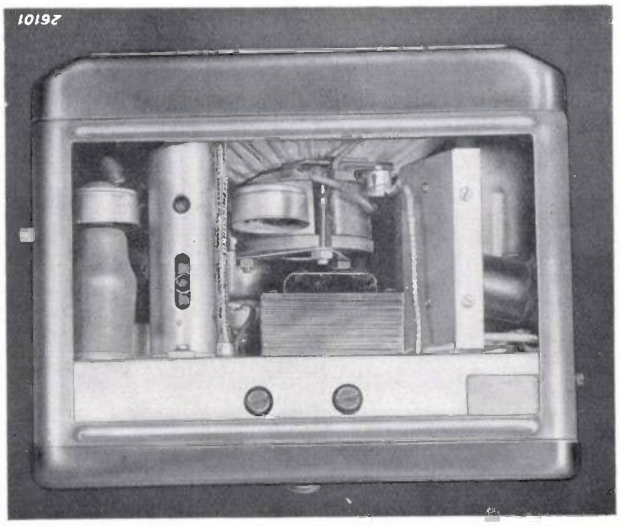

Fig. 5. View of the apparatus with a side wall cut open. Two valves have been removed from the middle in order to show that part of the apparatus which lies behind them.

box (fig. 7) is for tuning, and is connected by a Bowden cable with a worm in the set which turns the shaft of the tuning condenser. The transmission from the Bowden cable to the pointer is in the operating box itself. The dial is of a special type, since the requirement must be made that the tuning be easily visible at night as well as in the daytime, and moreover that it must not radiate light in such a way as to be disturbing. When driving in the dark an entirely lighted tuning scale is very disturbing; for that reason the following design has been chosen. The scale itself is dark with light letters (for reading in the daytime); the light letters allow light to pass through and are lighted from behind at night. Between the two wavebands there is a lighttransmitting band upon which a light coloured stripe is visible as a pointer in the daytime. This pointer is placed in the middle of a light transmitting sector of an otherwise non-transparent disc beneath the scale. Under this

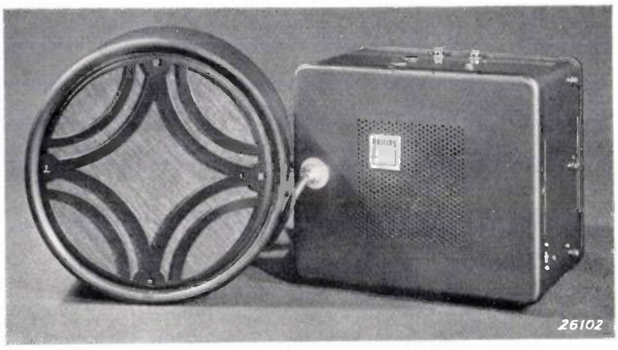

Fig. 6. Car radio set with separate loudspeaker.

disc is the lamp for illumination. In this way only a sector-shaped part of the scale is lighted, and the correct tuning point is then indicated by the shadow of the pointer stripe on a light background.

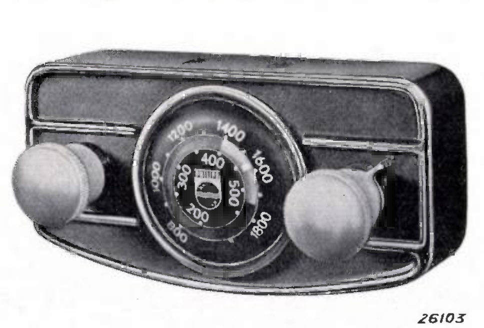

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Fig. 7. Operating box.

In addition, by using a light green lamp for the illumination, a pleasant, non-distracting but very clear reading of the scale is made possible.

On the operation box itself (fig. 7) to the right under the tuning knob there is also a switch by which the power supply to the apparatus can be switched on. The lefthand knob operates the volume control

Fig. 8. Operation box under the dashboard.

of the apparatus, also through a Bowden cable. For switching over to a different wave length range there is no extra knob. The volume regulator knob, when pulled out slightly, performs this function.

3) The apparatus must be very sturdily built, since at high speeds and especially on bad roads, it may be exposed to very heavy shocks. Because of this the parts must be extra well fastened. During assembly this feature is checked by placing the sets on a vibrating table, each corner of which falls once per second a distance of 12 mm. The four corners do not fall at the same moment, but one falls every quarter of a second. The construction is

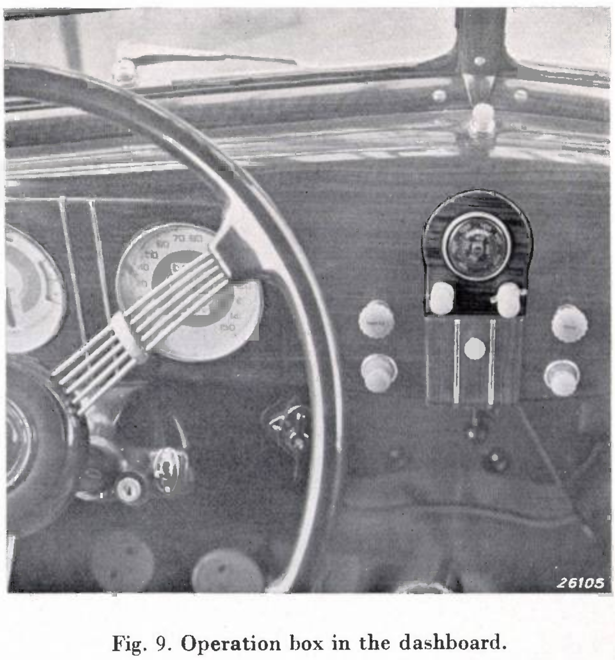

Fig. 9. Operation box in the dashboard.

such that the sets can withstand such shock tests for hours.

4) The apparatus must be easy to install. In order that this may be done quickly and without too many changes in the car, a single bolt method of installation has been chosen (*fig. 10.*) With this method it is only necessary to drill one hole in the partition behind the motor. The separate loudspeaker has the same type of fastening. In addition the cable connections for supply and aerial are made with a bayonet fastening which limits the time necessary for their assembly to a minimum.

5) The apparatus must be completely closed electrically. If the apparatus is suspended below the dashboard, it is close to connections which carry strong interference voltages. As has been mentioned in the discussion of interferences, the apparatus itself must be closed to such interferences. The housing is therefore of metal, and there is a wire gauze under the cloth screen of the loudspeaker. Since the front cover must be removable for the purpose of changing valves when necessary, there is an open crack between cover and housing. This is, however, entirely closed by spring strips. There are also such strips between the housing and the lower cover, since the latter must sometimes also be removed for inspection.

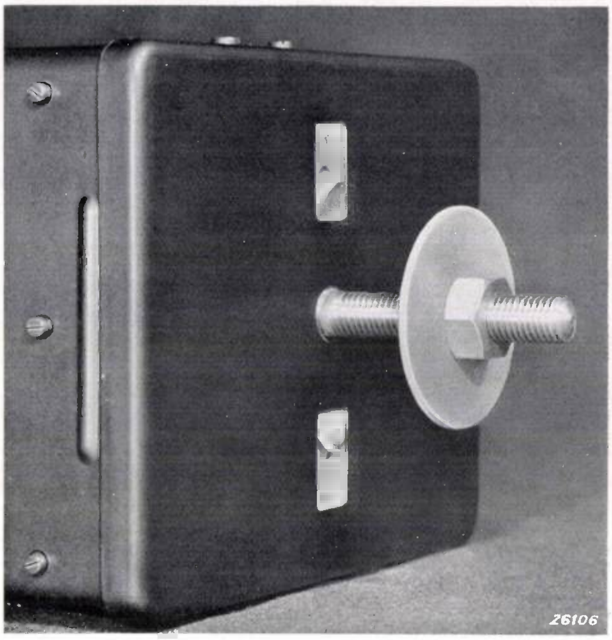

Fig. 10. The car radio set is fixed with a single bolt to the partition behind the motor.

THE SEALING OF METAL LEADS THROUGH HARD GLASS AND SILICA

by B. JONAS.

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A survey is given of the saeling of metal leads through different kinds of glass. The various conditions necessary for obtaining a tight and heat-resisting lead-in through quartz glass fused silica are discussed.

Properties of various glasses

At the beginning of the present century, after it had become possible to manufacture clear transparent fused silica on a large scale, the material quickly became indispensable in chemical and physical laboratories. Its success is due to its remarkable resistance to sudden changes of temperature and to chemical attack, and to its unusually satisfactory optical and electrical properties even at high tempeatures.

Until recently, however, silica was not used on a large scale in the construction of electric lamps or discharge tubes. There is a tendency to ascribe this fact to the relatively high price or to the technical difficulties of its use; this is, however, incorrect. The cause may rather be seen in the fact that at that time there was no perfectly good method of introducing current leading-in wires through the silica.

In order to show where the difficulties of this problem lie we shall compare it with the analogous problem in the case of ordinary glass. For this purpose the factors which are important in the process of sealing in are collected in *Table 1* to gether with some other data. The numerous technically important kinds of glass are so arranged in the table that "quartz" appears as the last member of a series, and no emphasis is laid on a contrast between "ordinary glass" and fused silica. It is obvious that this classification is somewhat arbitrary, for instance, in the establishment of the boundaries between the different types; moreover, special cases are omitted for the sake of clearness. Some of

Table 1. PHYSICAL PROPERTIES OF SILICATE GLASSES

	Group No.	Class	Sum SiO_2 and Al_2O_3 by weight %	Region where softening begins (°C): Transformation point	Region where softening begins (°C): Softening point	Coefficient of linear thermal expansion $\alpha \times 10^7$	Tensile strength kg/mm²	Resistance to compression kg/mm²	Heat conductivity cal/cm degr. sec.	Electrical conductivity T_{k100} °C	Dielectric losses at 1.5×10^6 cycles/sec. $10^4 \cdot tg\delta$	Transmissibility for ultra-violet of 1 mm of glass at 3000 Å
Ordinary soft glass	1	Lead glass (alkali lead silicate glass)	35 to 65	380 to 500	420 to 600	85 to 95	3 to 7	50 to 90	0.017 to 0.037	250 to 380	8 to 80	0% to 75 %
	2	Calcium glass (alkali calcium silicate glass)	65 to 75	470 to 530	550 to 600	80 to 110	4 to 10 (15)	70 to 100	0.025 to 0.030	130 to 300	10 to 120	
Hard glass	3	Soft boro-silicate glass (alkali boro-silicate glass)	70 to 85	470 to 570	500 to 700	30 to 60				160 to 400		
	4	Hard alumino (boro) silicate glass poor in or free from alkali	75 to 95	650 to 900	750 to 1000	8 to 38	5 to 15	100	approx. 0.03	350 to 600	approx. 10 to 30	up to 90%
		Quartz glass	100	1050	1300	5.5 to 6	>20(?)	>200	0.035	>600	2	92%

the values given in the table, for instance the tensile strength, are somewhat doubtful for various reasons, but they serve to give a general idea.

When one speaks of a "hard glass" one is not referring to the mechanical hardness, but to the fact that the glass is difficult to fuse and must be worked at a high temperature. In recent years the the term "hard glass" has also been used to indicate the fact that the coefficient of thermal expansion is small. By the term "soft hard glass" we shall in the following indicate a glass which has a small coefficient of expansion, but which becomes soft at a relatively low temperature.

By "transformation point" we mean the lowest temperature at which, after lengthy heating, it is still possible to relieve the stresses in the glass; at the softening temperature this can be done in a few minutes (for example in 15 minutes).

Instead of the conductivity at room temperature, the temperature T_{k100} is given, at which a conductivity g of 1/100 mhos/cm is reached.

Different forms of the metal-to-glass joint

All glass-to-metal joints must satisfy the requirement that the stresses, which appear during manufacture or in normal use due to the difference in coefficients of expansion of the materials used, shall not lead to a loosening of the joint or to the destruction of one of the parts. Joints obtained by cementing, mechanical methods or by electrolytic methods are not considered here, because of their low resistance to heat. It may be assumed that with a really good prepared glass-to-metal joint the adhesive strength is not lower than the tensile strength of the glass. When this is true the problem of the glass-to-metal joint is reduced mainly to the limitation of the stresses appearing in the glass.

In the solution of this problem two fundamentally different lines may be chosen:

1. by judicious choice of the two components glass and metal, the coefficients of thermal expansion are so adapted to each other that the final state is stable. (This does not mean that the coefficients of expansion exactly must be equal).
2. by taking very small thicknesses of metal, dangerous stresses are avoided in the glass. In this method no regard is paid to exceeding the limit of elasticity of the metal.

The second method is of only slight importance in the scaling in of wires, since with an appreciable difference in expansion the permissible thickness of wire is too small. For the fusing on of ring- or cap-shaped parts this method is very often used, especially when there are not many or important temperature changes at the boundary of glass and metal. In the case of fused silica such a construction would be accompanied by very great difficulties, considering the extremely high temperature during working. In general this method must be considered more as a subsidiary to the more usual process in the case of a large diameter of the metal part.

A method which can also be followed is the employment of a flux which is adapted to the glass of the tube, and has such a low softening temperature that deviations in the coefficient of expansion are permissible. The temperature interval through which the weld still cools after becoming solid then becomes shorter, and the thermal stresses also become smaller. In the cases where this method is of practical importance it comes down to a facilitation of methods 1) or 2), but it is not in principle a different solution.

Finally, when it is found impossible to adapt the cofficients of thermal expansion of glass and metal to each other, it is possible to bridge the difference by a series of kinds of glass with progressive properties. This solution can be used with fused silica (*fig. 1a*) and is very important for laboratory purposes. It is, however, not cheap, and in the case of lamps it leads to rather elaborate constructions, so that it cannot be considered for mass production.

In the following we shall confine ourselves to the primary form of a weld: the leading in of a wire or rod through a glass wall, a type of construction which can be achieved with all kinds of glass, even with pure silica. We shall now discuss the

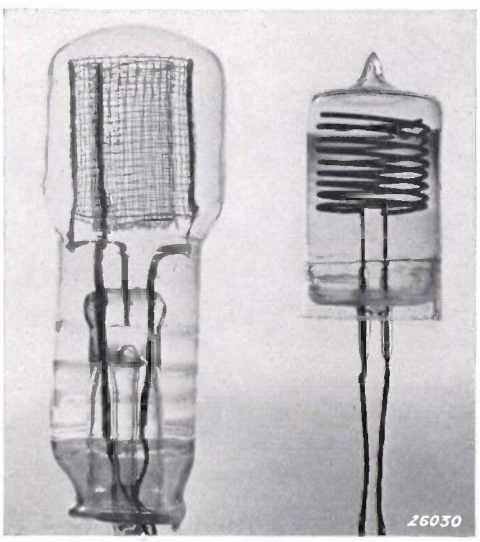

Fig. 1. Scaling of tungsten into silica bulbs with the aid of: *a*) transition by means of gradually changing kinds of glass, *b*) flux glass with a high melting point ($^3/_4$ of real siz

process with reference to Table I successively dealing with glasses of different degrees of hardness.

Sealing leads through "soft glasses" and "soft hard glasses"

In the case of the first two groups in the table I (soft glasses) limitation of the stress by adaptation of the thermal expansion gives no difficulties. For the region of coefficients of expansion from 80 to 110 $\times$ 10^{-7}, in addition to platinum and its alloys, there are several series of alloys with gradually varying properties. In addition there are composite metal wires, such as copper clad wire, which can also be adopted.

In order to prevent the leading-in wires, which unfortunately seldom consist of highly conducting metals, from becoming too hot, the limiting load is usually 10, or at the most 20 A per wire. If the load is not continuously and artificially cooled, upon exceeding this value there is danger of a break due to electrolysis of the glass (from about 200° C), since the conductivity of the glass increases rapidly with the temperature. In such cases a cap-shaped or cylindrical piece must be used, with highly conducting metals for the current leads.

For the third group in the table, the "soft hard glasses", the solution is also usually simple. With the softer kinds, whose thermal expansion is as a rule not particularly low, adaptable sealing-in alloys can again be used; with many of these glasses the metals tungsten and molybdenum can also be used. The appropriate alloys are ternary or quaternary Fe-Ni-Co alloys derived from the well-known Invar (approximately 64 Fe, 36 Ni). It is well known that Invar has a very small thermal expansion at room temperature. This is the result of a reversible thermal transformation which is complete at about 200°C. The expansion curve, which is practically flat to about 150°C, has the normal steeply mounting form above 200°C (*fig. 2*). By the addition of cobalt as the third constituent the transition from the flat to the steep part can be displaced to higher temperature, while the slope of the flat part increases. Such alloys can therefore — if necessary with the addition of other components such as Mn, Cu or Cr — be adapted for use with the softer kinds of glass of the third group. For the harder kinds of the third group the beginning of softening often lies at a much higher temperature than the bend in the expansion curve of the alloy. The disadvantage arising from this can be avoided by the introduction of a softer intermediate flux glass between the glass in question and the metal, when tungsten or molybdenum cannot be used.

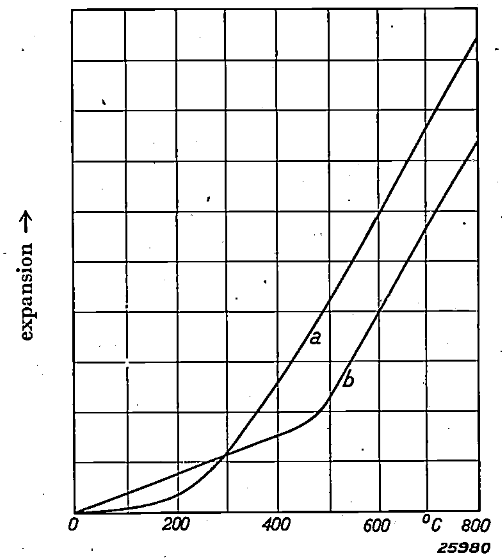

Fig. 2. Diagram of the variation of thermal expansion: *a*) Invar (36 Ni, 64 Fe), *b*) a Co-Ni-Fe alloy resembling Invar.

Summarizing, we may therefore say that for the three groups of softer glasses considered, a satisfactory solution can always be found no matter what the desired ratio of coefficients of thermal expansion may be (within certain limits). For the permissible value of this ratio, in the literature and in practice, a range is usually taken from 0.9 to 1.1. Sometimes one finds the permissible difference $\alpha_{\text{metal}} - \alpha_{\text{glass}}$ given in the literature as 20 $\times$ 10^{-7}.

Wire leads through hard glass and fused silica

In the case of the last two groups, if we wish to keep to the practical rule that the coefficients of expansion of the materials to be fused together, metal and glass, may not differ by more than 10 per cent, the problem of sealing in leads through these kinds of glass is insoluble. Of all the available metals, tungsten with 47 $\times$ 10^{-7}, has by far the lowest, and thus the most favourable, coefficient of expansion. Upon comparison with the kinds of glass now being considered, there are differences of 20 to 40 $\times$ 10^{-7}, and the ratio mentioned would therefore become 1.6 to 8 instead of 1.1. Moreover, the difference in coefficient of expansion is active over a much larger region, since the softening temperature is so much higher.

Upon closer consideration of the nature and magnitude of the stresses occurring, however, it is found that they do not exceed the limit of strength of the glass or of the glass-metal joint under circumstances which are not all too unsatisfactory.

For simple forms of the joint, as given in *fig. 3*, the stresses can be quantitatively determined by means of the double refraction of the glass under stress in polarized light. Such measurements have been carried out by Hull and Burger [1]) with joints of various dimensions and of many different kinds of glass and wire. The values measured were also compared with those calculated on the basis of simple assumptions, and were found to give a good agreement.

We shall now apply this method of calculation to the lead-in through hard glass of the fourth group which was not studied by Hull and Burger, as we believe that the method will also in this case give a good picture of the situation.

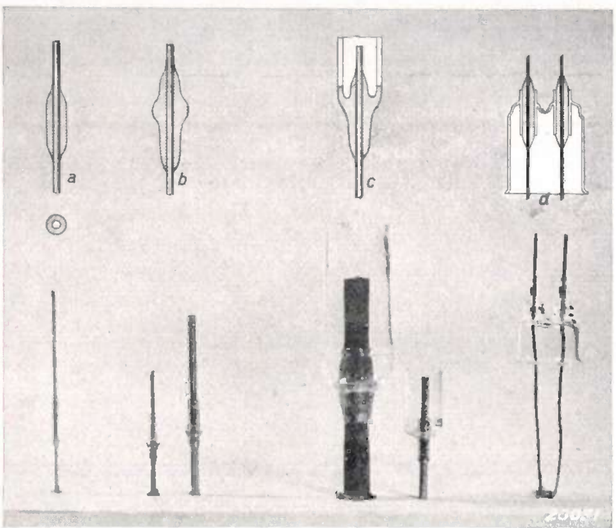

Fig. 3. Various forms of lead-in. Photographs and sectional drawings (scale 1 : 3).

Calculation of the distribution of stress

The calculations are based upon the following model: a long glass cylinder containing a concentric wire core. At the lower limit of the softening region of the glass the whole is in temperature equilibrium and without stress. We assume that upon further cooling only elastic changes in shape occur.

The metal core has the tendency to shrink more upon cooling than the hard glass covering. *Fig. 4* shows how in the glass tensile stresses occur in a radial direction and compressive stresses in tangential (and axial) directions. For a glass (see table I) the permissible loading on compression is about ten times that on stretching. In the model considered the ratio between the tensile and compressive stress is usually [2]) less than 10, so that rupture will occur only due to tensile stresses, and we shall concern ourselves in the following only with the latter (in the radial direction).

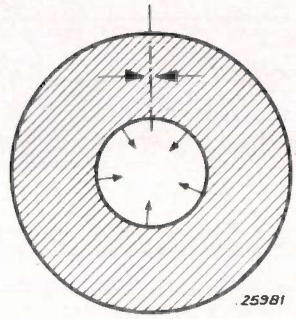

Fig. 4. Stresses occurring in glass part of lead-in, when $a_g < a_m$.

The radial stress σ_r as a function of the distance r to the axis is given by the equation:

$$\sigma_r = \frac{p}{n^2 - 1}\left(\frac{r_g{}^2}{r^2} - 1\right),$$

where r_g is the radius of the glass cylinder and n is the ratio of this radius to the radius of the metal core.

The radial stress σ_r is zero at the outer surface of the glass ($r = r_g$), and has a maximum p at the boundary glass-metal, where the ratio $r_g/r_m = n$. In passing it must be noted that in the metal core the radial stress is constant and equal to the maximum value p of the stress in the glass. In *fig. 5* the variation of σ_r/p is given as a function of the ratio r/r_g for several values of n.

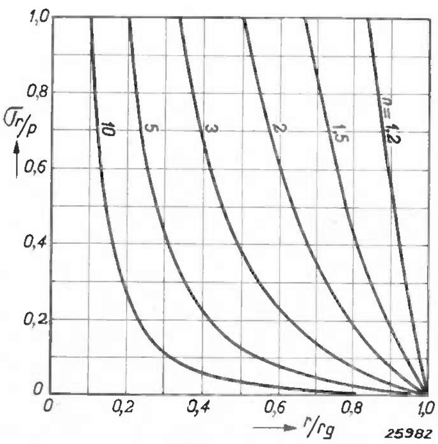

Fig. 5. Relative value of the tensile stress (σ_r/p) as a function of (r/r_g) for various values of n ($= r_g/r_m$).

[1]) A. W. Hull and E. E. Burger, Physics 5, 384 - 405, 1934.

[2]) With a very thin layer of glass the tangential (compressive) stress can become greater than 10 times the radial (tensile) stress. The value of the compressive stress then occurring is, however, according to calculation, always smaller than twice the maximum tensile stress with a very thick layer of glass. The axial stresses are never greater than the tangential ones at the boundary metal-glass. It is therefore sufficient to restrict the tensile stress.

The function which indicates the variation in the stress is the same as that which appears in the well-known case of a thick-walled tube with internal excess-pressure. The determination of the maximum stress p, however, demands a deeper study of the particular circumstances of the sealing-in problem. This has been discussed by Poritsky [3]) whose results were used by Hull and Burger after several simplifications and a correction.

From the calculation it follows that the maximum stress p is proportional to the product of the modulus of elasticity of the glass E_g and the difference δ in expansion per unit of length between glass and metal for the temperature interval under consideration. Further p depends on the ratio n ($= r_g/r_m$) and on the elastic properties of metal and glass.

$$p = \frac{E_g \cdot \delta}{\varrho_g \left(\frac{E_g}{E_m}\right) + \frac{\gamma n^2 + \varrho_m}{n^2 - 1}}$$

E_g, E_m are respectively the moduli of elasticity of glass and metal, while γ and ϱ are connected with Poissons's ratia μ of glass and metal.

$$\gamma = \frac{(1+\mu_g)\ [1 + (n^2-1)\, E_g/E_m]}{(1+\mu_g) + (n^2-1)\,(1+\mu_m)\, E_g/E_m} \approx 1$$

$$\varrho_g = 1 - 2\mu_g;\ \varrho_m = 1 - 2\mu_m$$

The order of magnitude of ϱ_g and ϱ_m is 0.5.

It may perhaps seem strange that the stress p on the inner side of the glass does not depend explicitly on the dimensions of the wire fused in or of the glass surrounding it, but only on the ratio n. Experience of course shows that it is more difficult to make a "heavy" lead-in than a "light" one. This fact is, however, a result of the increasing difficulties in the manipulation and cooling, when the dimensions of the piece of work become larger.

The expression given for p enables us to calculate the dangerous stresses in a weld with an accuracy depending upon the reliability of the elastic constants and the assumptions concerning n. As an example we have chosen one of the most unfavourable cases out of group 4, an alumino-boro-silicate glass which, as experience has shown, can easily be sealed to quartz. With $E_g = 6700$ kg/mm², $E_m = 36000$ kg/mm², $\delta = 3.0 \times 10^{-3}$, $\mu_g = 0.25$ and $\mu_m = 0.17$, we find for different

Table 2.

Maximum stress p (kg/mm²) occurring with a simple leading-in wire through an alumino-borosilicate glass (group 4) with $E_g = 6700$ kg/mm², $E_m = 3600$ kg/mm², $\delta = 3 \times 10^{-3}$, $\varrho_g = 0.5$, $\varrho_m = 0.66$.

n —	1.0	1.2	1.5	2.0	3.0	5	10	∞
p —	0	3.5	6.9	10.1	12.4	12.9	14.5	15.5

values of n the numbers given in *table 2* (cf. *fig.* 6). As was to be expected, the values found for p are very high, for a high value of n the breaking limit of this glass appears to be attained or exceeded. In practice, however, the joints under consideration hold well, even at a high temperature and upon sudden large changes in temperature. We shall investigate the reason for this unexpectedly high resistance.

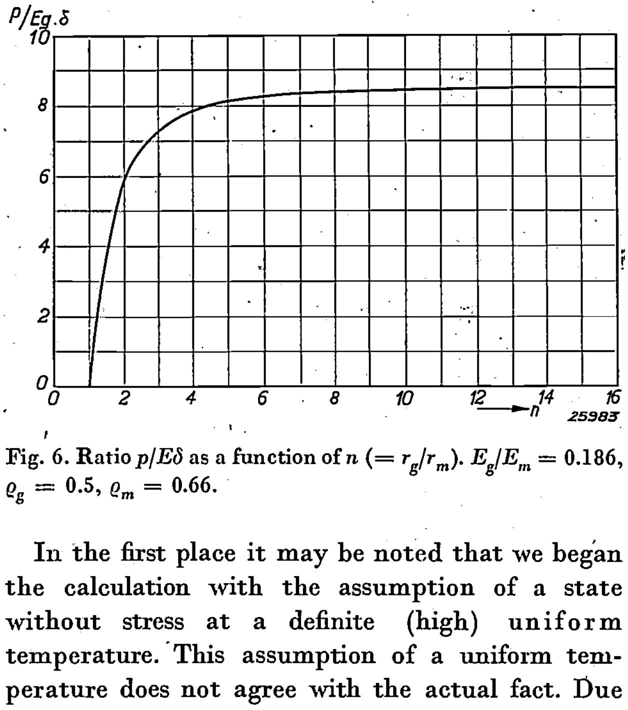

Fig. 6. Ratio $p/E\delta$ as a function of n ($= r_g/r_m$). $E_g/E_m = 0.186$, $\varrho_g = 0.5$, $\varrho_m = 0.66$.

In the first place it may be noted that we began the calculation with the assumption of a state without stress at a definite (high) uniform temperature. This assumption of a uniform temperature does not agree with the actual fact. Due to the high temperature during working and to the good heat conduction of the metal (tungsten), a temperature gradient occurs in a radial direction toward the axis. The metal then shrinks less later on, and the maximum stress p is reduced, so that the values calculated are to be considered as an upper limit which is never actually reached.

The incidental factors do not always act in a favourable sense, that is, to decrease the maximum stress. Due to uneven cooling the stresses occurring may become greater than the value calculated. This occurs when $\alpha_m > \alpha_g$, and the glass covering cools more rapidly than the metal wire. In the case which we are considering, on the other hand, a fall of

[3]) H. Poritsky, Physics 5, 406 - 411, 1934.
G. Heller has called our attention to several errors in the derivation given in the article referred to. In the formulae used here the necessary corrections have been made.

temperature toward the axis is ensured during working.

A possible doubt as to the strength of the adhesion between glass and metal calls for care. When $a_m > a_g$ the maximum tensile stress occurs at the boundary surface, and even a local loosening of the seal between metal and glass can easily lead to rupture of the glass.

The sealing of tungsten to the kinds of glass here discussed gives a weld which is one of the strongest known with respect to the adhesion of metal to glass. If, however, no details are known about the adhesion, this factor must not be assumed to be too high, and a smaller difference in coefficients of expansion must be chosen.

Under certain circumstances much higher tensile stresses may be encountered in glass than the maximum values determined in the usual way from tensile tests. In such tests a great variation in the results is found which is usually ascribed to uncontrollable injuries and faults in the glass surface. With glass rods whose surface is already under tangential and axial compression before being loaded, due to a sudden cooling, appreciably higher values of the tensile strength and resistance to bending occur, even though there are previous tensile stresses in the core.

An analogous phenomenon is known in the case of the so-called "Prince Rupert's Drops" which can be made by allowing drops of glass to fall into a cooling liquid. The surface of the drops is under a high compressive stress. From optical measurements it may be deduced that extraordinarily high tensile stresses occur in the interior of the drops. This same heat treatment is applied technically to the "hardened" safety glass. In these examples, due to the avoidance of tensile stresses at the surface, an abnormally high tensile strength occurs at the core. With seals of tungsten through glass with a lower coefficient of expansion an analogous distribution of stresses is obtained, at least when the surface of contact between metal and glass need not be considered as a glass surface. As long as glass and metal adhere, all practical experience seems to support this conclusion. Because of these favourable circumstances a much higher tensile stress for the glass may be assumed than is usual.

It has actually been found possible to make reliable leads of tungsten through almost all the kinds of glass in the fourth group, even with the glass referred to in Table 2 with excellent results. It thereby becomes possible to introduce lead-in wires into silica walls by means of a weld which is tight and resistant to heat, since some of the glasses in group 4 can be sealed to silica. With pure silica alone, it is, according to calculation, which is confirmed by experience, impossible to make a tight weld, at least of ordinary dimensions.

Fig. 3 shows the practical form of a lead-in of tungsten through fused silica with the aid of an intermediate glass with a high melting point (see also fig. 1b). It may be seen that remarkably "heavy" welds can be made. In this connection we finally call attention to the very high current load permissible. If the temperature of the wire increases by development of heat in the wire or by conduction of heat from the inside of the tube, the stresses in the glass become smaller, in contrast to the usual case of equal expansion ($a_m = a_g$).

Thanks to this decrease in stress it is possible to seal in leads for currents up to many hundred amperes. The current is limited by the oxidation of tungsten occurring at dark red heat, and to a smaller degree by the danger of electrolysis, which with the glass in question, only becomes of importance at still higher temperatures.

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FLUORESCENCE

by W. DE GROOT.

535.37

In this article an attempt is made to give a survey of the phenomena of fluorescence by beginning with the simplest case, the fluorescence of metal vapours, and passing on to phenomena of a more complex character. The fluorescence of the following substances is discussed: monatomic (Na, Hg, Ne) and molecular gases (I_2), solid substances (uranyl compounds) and liquids (fluorescein, eosin).

Introduction

The discovery of the phenomena of fluorescence and phosphorescence dates from the renaissance. It was the shoemaker-alchemist Casciarolus who discovered phosphorescence in 1602, when he heated a barite in the fire and obtained a substance which, after having been exposed to the light continued to glow with a red light in the dark. In the case of fluorescence it may be noted that Grimaldi in 1665 found that an extract of lignum nephriticum, a wood then considered to possess healing properties, is yellow by transmitted light and blue by reflected light (seen against a dark background). The blue colour had already been described by Monardes in 1575.

Fluorescence is the property of a substance of radiating energy which it has absorbed, in the form of light usually with an altered spectral composition. Phosphorescence is generally used to describe the phenomenon when the substance continues to radiate light for a long time after it has been irradiated.

The development of physics and chemistry in the nineteenth century learned to know a large number of fluorescent substances. Kayser's well-known handbook (IV, 1908) gives more than one thousand, among which many organic compounds. Phosphorescent substances were first investigated in detail and prepared in large numbers by Lenard and his students (1894). The importance of all these substances, the fluorescent as well as the phosphorescent, in the technique of lighting has become steadily greater, particularly in connection with the increasing use of gas discharge lamps. In addition to the visible, and therefore directly useful, radiation, an appreciable amount of invisible ultra-violet radiation is emitted by such lamps, particularly those which contain mercury vapour. With the aid of fluorescent and phosphorescent substances it is possible to transform a portion of this invisible radiation into visible light, and in that way to improve the lighting efficiency of the lamp. The substances mentioned are also important because of the fact that the visible light which they emit differs in spectral composition from that of the discharge lamp itself, and therefore offers the possibility of changing the colour of the light and adapting it to a given purpose.

It is our intention to discuss later some of the properties and the technical applications of the fluorescent and phosphorescent substances now in use, after we have given first an account of the physical side of the phenomena under consideration.

Fluorescence of gases (resonance)

The simplest case of fluorescence is that of substances in the gaseous state, especially monatomic gases. Let us for example consider the vapour of the metal sodium. This vapour consists mainly of sodium atoms which move about quite independently of each other.

It was previously explained (Philips techn. Rev. 1, 2 - 5, 1936) how, according to the concepts of the quantum theory, an atom may be not only in the lowest state of energy, but also in various higher energy states (excited states). *Fig. 1* gives the energy level diagram (term scheme) of the sodium atom; several of the most important energy states are given in the figure with their energy content expressed in electron volts (1 electron volt, often simply called volt, is equal to 1.6×10^{-12} erg). It may be seen from this diagram that in addition to the normal or lowest state there is an important energy level at 2.1 electron volts. Upon transi-

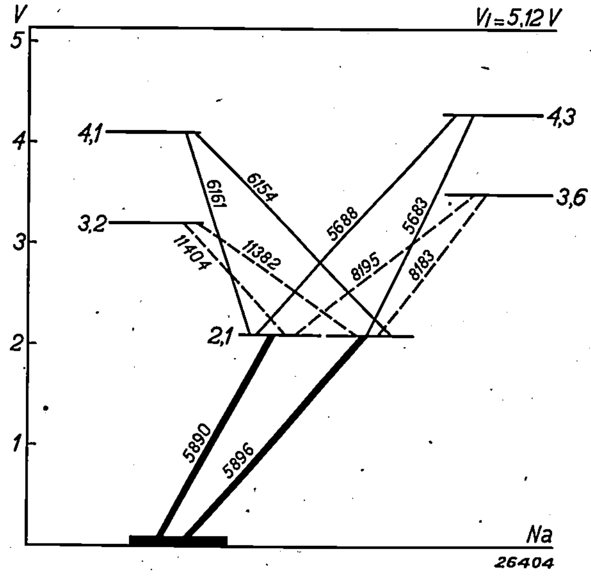

Fig. 1. Energy level diagram of a sodium atom. The lines indicate possible transitions (broken lines lie outside the visible part of the spectrum). The numbers indicate the wave length in Å of the spectrum line corresponding to the transition.

tion from this state to the lowest state the atom can emit a light quantum with an energy of:

$$2.1 \cdot 1.6 \cdot 10^{-12} = 3.96 \cdot 10^{-12} \text{ erg}.$$

The wavelength in Ångstrom units of the light emitted can be found from the formula [1])

$$\lambda_{\text{Å}} = 12390/V_{\text{volts}}.$$

In this case the wave length is 5890 Å; as a matter of fact the 2.1 volt level is double, and there are two lines with wave lengths 5890 Å and 5896 Å, the two D-lines of the sodium spectrum, which are also responsible for the yellow colour of the sodium flame and the yellow light of sodium lamps.

Let us now imagine the following experiment (*fig. 2*). *L* is a light source in which sodium atoms are made to emit the D-lines, for instance a sodium lamp. Part of the light from *L* is concentrated by a lens *l* on a bulb made of a glass resistant to sodium and containing metallic sodium. The bulb is maintained at a temperature of about 200° C by an oven *O*. We then see that the bundle of rays stands out inside the bulb as a cone of yellow light. Experiments like this were first carried out by R. W. Wood in 1905. On the basis of the then prevailing conceptions about the emission and absorption of radiation this phenomenon, which was termed resonance, was considered to be due to the sympathetic vibration of the oscillators present in the vapour atoms with the light waves sent out from *L*, in the same way that a tuning fork vibrates when a nearby tuning fork of the same pitch emits sound waves. The name resonance has been retained, although the explanation of the phenomenon is at present different. According to the quantum theory it must be assumed that the atoms in the light source pass from the 2.1 volt states to the lowest state. These atoms had reached a 2.1 volt state by taking up energy from the electrons which had obtained a certain kinetic energy in the positive column of the sodium lamp under the influence of the electric field. They radiate the energy taken up in the form of light quanta which are absorbed by the vapour atoms in the bulb *B*, so that the latter now reach the 2.1 volt state. These atoms will in turn almost immediately radiate the energy taken up again, and this radiation is the resonance light observed. If a foreign gas, nitrogen for instance, is added to the sodium vapour, the intensity of the resonance light is diminished. This may be interpreted to mean that the sodium atoms in the 2.1 volt state can give their energy to a nitrogen molecule by collision. The nitrogen molecule either moves forward (energy of translation) or it takens on a vibratory motion (energy of vibration) in which the two nitrogen atoms of the molecule vibrate with respect to each other. The sodium atoms then return to the lowest state without having radiated any light.

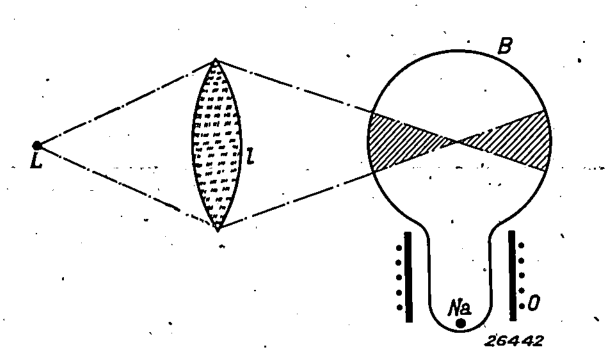

Fig. 2. Resonance radiation of sodium vapour. *L* light source, *l* lens, *O* oven, *B* bulb in which sodium (Na) is evaporated.

Resonance in mercury vapour

While for the case discussed above (sodium) it seems to be a matter of indifference whether one chooses the theoretical quantum or the older classical interpretation of the phenomenon of

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[1]) In the quoted article of G. Heller the number 12340 is given; since for the fundamental physical constants *e* and *h* some slightly changed values are adopted, through which this number is somewhat changed.

resonance, in more complicated cases, several of which we shall now discuss, the great advantage of the quantum theory becomes evident.

Let us consider the energy diagram of the mercury atom (*fig. 3*). In this case we encounter three states of energy above the lowest state with energies of 4.7, 4.9 and 5.45 electron volts. From the lowest state the atom can pass into the 4.9 volt state by the absorption of an ultraviolet quantum of a wave length of 2537 Å. By letting light of this wave length fall on mercury vapour in a quartz bulb resonance may be observed in the same way as with sodium. The observation must be photographic in this case and not visual, since the eye is not sensitive to radiation of a wave length of 2537 Å.

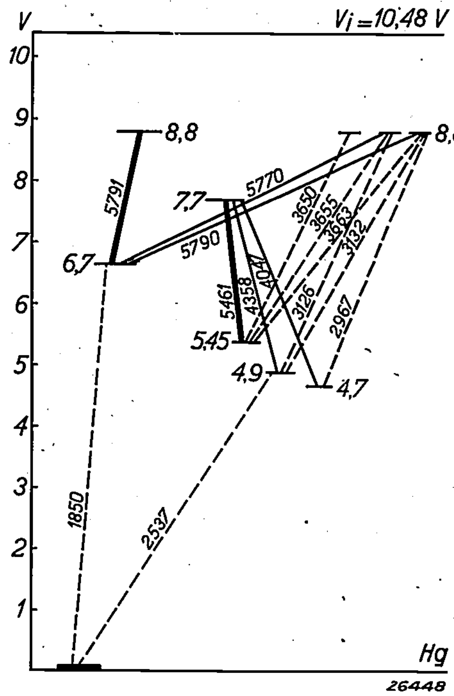

Fig. 3. Energy level diagram of a mercury atom.

If we consider the energy scheme of mercury more closely, we see that an atom in the 4.9 volt state can, by absorption of a quantum of violet light with a wave length of 4358 Å, pass over into a still higher state (7.7 electron volts), while it not only can return to the 4.9 volt state with the radiation of 4358 Å, but also to the 5.45 volt and the 4.7 volt state with the radiation of a green or a violet quantum respectively (wave lengths 5461 and 4047 Å). W o o d irradiated mercury vapour with light of the wave lengths 2537 and 4358 Å simultaneously, and saw that the vapour began to radiate the green line 5461 Å (*fig. 4*). This shows that an atom which has absorbed 2537 Å remains in the 4.9 volt state long enough to have the opportunity of absorbing a quantum of 4358 Å. This experiment would be difficult to explain with the conception of oscillators.

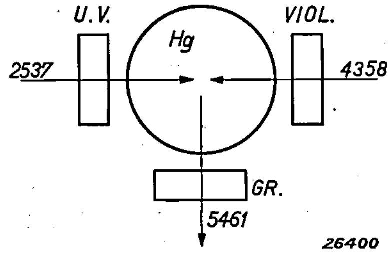

Fig. 4. Composite resonance in mercury vapour according to W o o d. *UV* ultraviolet filter, *VIOL* violet filter, *GR* green filter.

When W o o d added a neutral gas, such as nitrogen (N_2) to the mercury vapour, radiation with 2537 Å then resulted in a strong absorption of the line 4047 Å. The triplet of lines 5461, 4358 and 4047 Å appeared in great intensity (*fig. 5*). This phenomenon may be explained in the following way. The mercury atoms which have attained the 4.9 volt state by absorption of 2537 Å, can pass energy on to a nitrogen molecule by collision, as was the case with sodium. The mercury atom may among other possibilities pass over from the 4.9 volt state to the lower 4.7 volt state. The latter state differs from the states so far discussed in the fact that the atom is unable except to a small degree to return from this state to the lowest state by radiation. The line (2656 Å) corresponding to this transition, which is not drawn in fig. 3, is a so-called f o r b i d d e n line of the mercury spectrum; this means practically that its intensity is smaller by a factor of 10^6 than that of corresponding non-forbidden lines. The atom which has once reached the 4.7 volt state therefore will remain in that state much longer than in the 4.9 volt state for instance. The 4.7 volt state is therefore called m e t a s t a b l e. Under the conditions of the experiment (irradiation with 2537 Å in the presence of nitrogen molecules), there is an appreciable accumulation of metastable atoms which explains the strong absorption of 4047 Å.

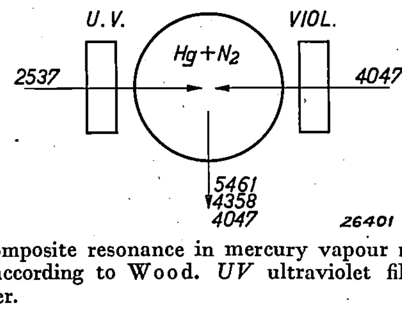

Fig. 5. Composite resonance in mercury vapour mixed with nitrogen according to W o o d. *UV* ultraviolet filter, *VIOL* violet filter.

Resonance in neon

The resonance phenomena in the gas neon are somewhat more complex. *Fig. 6* is a photograph of resonance light in neon. In the lower half of the bulb, which is filled with neon at a pressure of several millimetres, there are two electrodes between which a glow discharge is maintained. If the bulb is irradiated with a horizontal beam of light from a tube which emits red neon light, the result is a beam of red resonance light in which particularly the line 6402 Å dominates (see *fig.* 7). In the glow discharge the two resonance lines of neon (735.7 and 743.5 Å), among others, are excited. This radiation spreads out over the whole bulb and is absorbed by the other neon atoms which thereby attain states of energy of 16.84 and 16.67 volts (see fig. 7).

Fig. 6. Resonance light in neon. A horizontal beam of neon light is incident from the right (←).

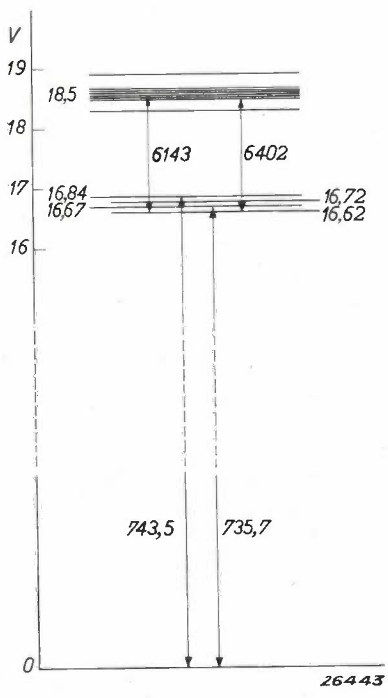

Fig. 7. Energy level diagram of a neon atom.

K. W. Meissner and H. B. Dorgelo have shown that under such circumstances various orange and red neon lines are strongly absorbed, namely lines 6402 and 6143 Å, which appear as combinations of the lower metastable level with several energy levels lying around 18.5 electron volts. Considering the absence of foreign atoms or molecules, it must be assumed that in this case the metastable atoms occur by collision between normal neon atoms and atoms which have been brought into the non-metastable 16.67 volt state, for instance, by the absorption of light of short wave lengths. In the collision the latter atoms have passed on a small amount of energy to the former in the form of energy of translation.

Considering the fact that the distance between the lower (metastable) level to the next higher one is only 0.05 electron volts, an amount which is comparable to the average kinetic energy of thermal agitation, it may be expected that neon atoms can in this case pass on energy of translation to metastable atoms and bring them into a state from which they can return to the lowest level by radiation. Dorgelo actually found that in neon gas at 200° C the concentration of metastable atoms is much lower than at room temperature under otherwise similar conditions. It is possible, as in the cases of sodium and mercury, to influence the absorption and resonance phenomena by adding foreign gases. If hydrogen gas is added to the neon in the bulb of fig. 6 through a hot palladium tube, the resonance light disappears almost completely. Considering the fact that the energy necessary to dissociate a hydrogen molecule is 4.4 electron volts, it may be assumed that in the case in question the energy of the metastable atoms is used up in dissociation of the hydrogen.

We have considered the composite resonance phenomena in neon and mercury vapour in some detail, because later on in the discussion of the phosphorescent substances we shall encounter situations which will remind us very much of the phenomena in monatomic gases. As a transition to the fluorescence phenomena in liquids and solid substances we shall first discuss the fluorescence of diatomic molecules and of simple groups of atoms in the gaseous state.

Fluorescence of iodine vapour

Iodine vapour (I_2) is a typical example of a diatomic gas. The fluorescence of this gas was also investigated by Wood and others. Since the spectrum of a molecular gas (band spectrum) is much richer in lines than the line spectra of the monatomic gases, we must expect a much more complicated scheme of energy levels in the case of a molecule. With a molecule, just as with an atom, an electron can move in orbits of higher energy, and the position of the energy levels here also is mainly determined by these orbits. Instead of one level, however, in the case of a molecule a given electron orbit is responsible for a series of approximately equidistant levels which are due to the different vibration states of the two atoms with respect to each other. Moreover, each vibration level is found to be made up of a number of levels due to the different possible values which may be assumed by the rotation energy of the molecule.

In iodine vapour at room temperature all the molecules are in the lowest state, and, moreover, in the lowest vibration level of that state, since thermal agitation is not sufficient to excite higher vibration states. The molecules, however, will rotate due to thermal agitation, so that not only the lowest, but also higher rotation states will be represented. For all these rotation states transition to a higher electron state is possible by absorption of light of a suitable wave length. The maximum of these transitions lies at 2.25 electron volts, i.e. in the green; this results in the violet colour of iodine vapour by transmitted light. Wood irradiated iodine vapour with light of the green mercury line 5461 Å. This line practically coincides with an absorption line which indicates a transition from the 33rd rotation state of the lowest vibration state of the lowest electron energy level to the 34th rotation state of the 26th vibration state of the higher electron energy level. From this state the molecule can upon radiation not only return to the lowest but also to the 1st, 2nd etc. vibration states of the lowest electron level. According to a rule of spectroscopy which only permits a jump of one in the rotation quantum number in a transition, the molecule can only pass over to the 33rd or the 35th rotation state of these vibration states. The result is that a number of double lines are observed in the fluorescence light which follow each other in the spectrum at approximately equal distances, and all of which have a wave length greater than 5461 Å. In all, 27 such doublets are observed.

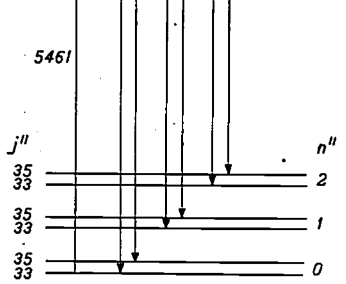

Fig. 8. Energy level diagram (partial) of an iodine molecule.
$n'n''$ vibration quantum numbers
$j'j''$ rotation quantum numbers.

In *fig. 8* we have reproduced the scheme of energy levels for iodine with the transitions in question, and in *fig. 9* the fluorescence spectrum.

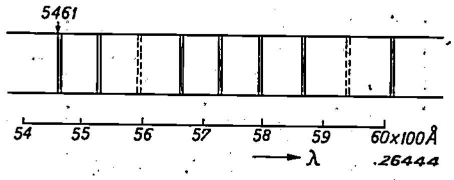

Fig. 9. Fluorescence spectrum of iodine vapour irradiated with the green mercury line 5461 Å. The marked doublets indicated by dotted lines are invisible due to selective absorption.

If a foreign gas, for instance helium, is added to the iodine vapour, an exchange of energy takes place between the excited molecule and the foreign atoms. As a result of this, excited molecules now appear in other vibration states and in rotation states other than the 34th, and the emission spectrum becomes richer in lines and resembles more nearly the normal band spectrum of iodine.

If iodine vapour is irradiated with non-monochromatic light, sunlight for example, the possibilities of absorption become more numerous, and the emission spectrum becomes richer in lines. In this case also a band spectrum is observed in fluorescence which, with respect to the wave length of maximum absorption (about 5400 Å) is displaced in the direction of longer waves. The fluorescence light of iodine vapour is indeed yellow-green in colour, while the maximum absorption lies in the green and blue-green part of the spectrum. We shall encounter this phenomenon again in the fluorescence of more complicated molecules. It is known as Stokes' law.

Use of potential curves

On the basis of the energy level diagram alone it is impossible to obtain a complete insight into the optical phenomena of molecules, since the phenomena are partially determined by the mutual separation r of the atoms in the molecule. We shall explain the use of potential curves in this connection by the example of iodine. If we consider two iodine atoms at a great distance from each other (*fig. 10*), and let them approach each other, the energy will decrease as the distance decreases until the atoms (I) have reached the distance at which they are normally bound in the molecule (2.7 Å). If we force the atoms to approach each other still more closely, they repel each other, and

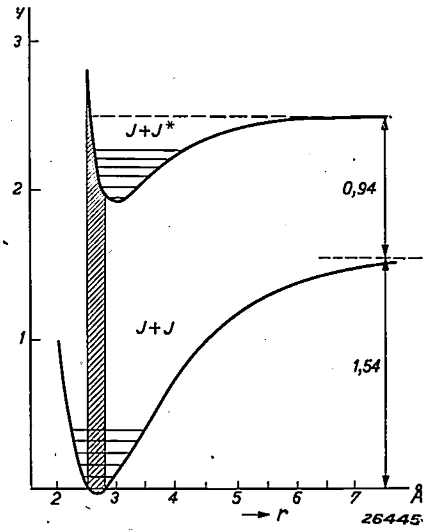

Fig. 10. Potential curves of an iodine molecule. The potential energy V in electron volts counted from the level corresponding to the lowest state as a function of the mutual separation r of the nuclei in Å. The vibration levels 0, 5, 10, 15, 20 and 25 are drawn.

the mutual energy begins to increase. This energy is often conceived as a potential energy, and the curve giving this energy as a function of the distance is called a "potential curve". The potential curve of the normal iodine molecule, I_2, has a minimum, therefore, at 2.7 Å. The depth of the minimum (1.54 electron volts) is equal to the chemical binding energy. Besides the normal state, among others a higher energy state of the atom I is known with an energy of 0.94 electron volt. If we repeat the process with an ordinary iodine atom and an excited atom at 0.94 volt, we find, as may be seen from the analysis of the band spectrum, a curve with a minimum at 3 Å; the depth of this minimum is 0.55 electron volt.

In fig. 10 several vibration levels are also indicated, for both curves the levels with the numbers 0, 5, 10, 15, 20, 25. We count the energy of the molecule from the zero vibration state of the lowest electron level and read from the figure for instance that the 10th vibration state of the normal molecule has an energy of 0.16 volt. Moreover, the amplitude of this vibration may be deduced from the figure, since it is clear from the curve that the distance apart of the centres of the two atoms fluctuates between 2.3 and 3.2 Å. There is an important principle formulated by J. Franck and E. V. Condon which says that upon absorption of a light quantum, there is preferably a transition between states which lie directly one above the other in the potential curves of the diagram, so that it may be said that upon absorption the separation of the centres of the atoms changes as little as possible. In fig. 10 the limits of the absorption spectrum are constructed approximately by drawing vertical lines through the outermost points of the zero vibration level. One then finds (see cross hatched area) for the points of intersection with the upper curve values of 1.95 - 2.8 electron volts (6500 - 4400 Å). The last value (2.8 electron volts) lies above the energy necessary to separate two atoms and to excite one atom ($1.54 + 0.94 = 2.48$ electron volts). By absorption of light with a wave length shorter than $12390/2.48 = 5000$ Å the molecule I_2 is actually dissociated into one normal and one excited atom; fluorescence is then no longer possible.

Fluorescence of uranyl compounds

The compounds of the radical UO_2, for example $UO_2(NO_3)_2$, $K_2UO_2(NO_3)_4$, form a group of solid substances to which what has been stated about gaseous compound also applies. The radical UO_2, which occurs in the compounds mentioned as a doubly charged positive ion, $(UO_2)^{++}$, forms such a closed unit that even in the solid state absorption and emission bands appear which are comparable to those of free molecules in the gaseous state. We shall illustrate this with the help of potential curves. The radical UO_2 is entirely comparable with the molecule CO_2, which is known to have a symmetrical extended structure (*fig. 11*). The three atomic nuclei lie in a straight line with the two oxygen atoms at the ends. It is known that such symmetrical molecules can vibrate symmetrically, the two oxygen atoms approaching and receding from each other by turns, and the central atom remaining at rest. The frequency of this vibration for UO_2 in the lowest state is 2.5×10^{13} periods/sec, corresponding to a reciprocal wave length of 830 cm^{-1}, and therefore the vibration quantum is about 0.11 electron volt. If an uranyl salt is irradiated with blue or violet light, a number of bands are found in the fluorescence spectrum whose differences in frequency correspond to an energy difference of 0.103 electron volt.

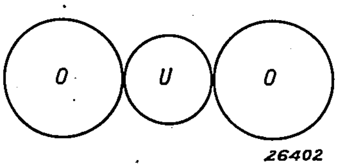

Fig. 11. Model of the symmetrical extended group of atoms (UO_2).

In order to obtain a complete picture, however, the absorption spectrum must also be considered. From this it is found that the distance between the emission bands is greater (0.103 electron volt) than that between the absorption bands (0.086 electron volt). When a uranyl compound is irradiated at a low temperature (—180° C for instance), the absorption band lying farthest toward the red coincides with the emission band lying farthest toward the violet. At room temperature a weak band is added to the series of absorption bands at the red end at a distance equal to that between the emission bands (0.103 electron volt), while at the violet end of the emission spectrum a new band appears at a distance equal to that between the absorption bands (0.086 electron volt). At a higher temperature, therefore, there are three absorption and emission bands which exactly coincide (*fig. 12*). The

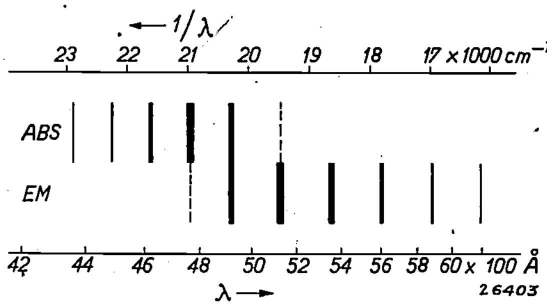

Fig. 12. Emission and absorption spectra of potassium uranyl sulphate (25° C). The lines represent diagrammatically the position and intensity of the bands. The bands indicated by broken lines disappear at very low temperatures 1000 cm^{-1} = 0.1239 electron volts).

late N. Moerman of Eindhoven gave the following interpretation of this phenomenon (see *fig. 13*). At a low temperature (—180° C) only the lowest level is occupied, absorption is therefore only possible out of this level. The separation (0.086 electron volt) of the absorption bands is given by the spacing of the vibration levels of the excited state. It is known that the radiation takes place very slowly with uranyl (time 10^{-5} sec compared with 10^{-8} sec for sodium); this time is so long that the vibration in the excited state becomes adapted to the temperature. At —180° C therefore, in the excited state also, only the zero vibration level is represented. The emission bands due to this state exhibit the spacing 0.103 electron volt, which belongs to the various vibration levels of the lowest state. When the temperature rises, not only the first vibration level of the lowest state (absorption) but also that of the excited state (emission) plays a part.

Uranium glass, which is important in the technique of discharge tubes, exhibits a fluorescence which resembles that of the uranyl salts. Here also there is a maximum of emission in the green, while the absorption extends from the blue-green to far in the ultraviolet. There is, however, only a vague band structure, which is not surprising with a glass.

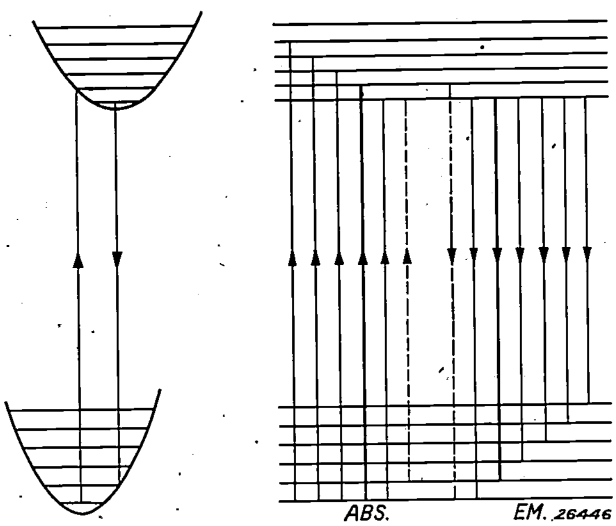

Fig. 13. Left: potential curves of the UO_2 radical with the strongest transitions in emission and absorption. Right: energy levels with transitions in the same order as the lines in fig. 12. The levels indicated by dotted lines do not occur at — 180° C, while they do occur at room temperature.

Fluorescence of organic compounds

Among organic compounds there are many which fluoresce strongly in solution. Fluorescein and eosin are well known examples. Eosin, which has a beautiful rose colour by transmitted light, is distinguished by strong absorption in the green and blue parts of the spectrum; the maximum of this absorption lies at 5250 Å. If light from this part of the spectrum is allowed to fall on a solution of eosin, a greenish fluorescence light is radiated, whose spectral distribution possesses a maximum at 5450 Å. In *fig. 14* the absorption and emission spectra of such a solution are reproduced. It may be seen that Stoke's law is fulfilled. Although it is not exactly known to which group of atoms in the complicated

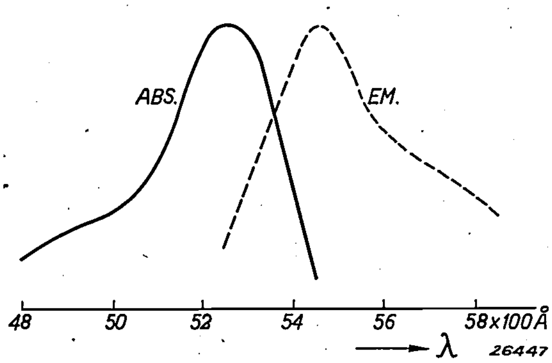

Fig. 14. Spectral distribution of the absorption and emission of eosin.

eosin molecule the fluorescence is due, and although it is impossible to indicate the correct position of the potential curves for this group of atoms, we may assume that the curves are similar to those in the cases of iodine and uranyl. The fact that the emission spectrum in this case does not consist of fine lines but of a continuous band, must be explained in the same way as in the case of uranium glass, as due to the fields of the water molecules which blur the spectrum.

The influence of concentration and of foreign substances on the efficiency of fluorescence of solutions is important. If the ratio of the radiated fluorescence light to the absorbed light is measured for a very dilute solution of fluorescein, it is found that per quantum of light used for irradiation, approximately one quantum is radiated. The efficiency is, therefore, as high as possible. If the concentration increases, the efficiency decreases. Energy must therefore be destroyed. The same thing happens when an electrolyte, KI for instance, is added to a dilute solution of fluorescein. In this case it can be proved that as a result of the irradiation free iodine is formed which darkens to a starch solution. This may be explained in the following way. The fluorescein molecule, which has attained a higher energy state by absorption of a light quantum, uses up this energy by collision with an I^{-1} (H_2O) ion in a chemical reaction giving free iodine:

$$I^{-1}\,(H_2O) \rightarrow I\,H + (OH)^{-1}.$$

We are therefore concerned here with a case analogous to that of the fluorescence of gases, where we saw that the energy taken up by the atom as a result of light absorption can be transformed wholly or partially into atomic energy of another form (translation, vibration, dissociation).

The decrease in efficiency by increase of concentration of the fluorescent substance may then be interpreted as follows. The molecules of the substance can easily take over energy from an excited molecule and transform it into vibration and translation. With normal absorption of light, without fluorescence, this occurs even in the absorbing molecule itself. The fluorescing substances must, therefore, be distinguished by certain groups of atoms which are "shielded" from their environment in such a way that they have time enough to emit again as radiated light the energy they have received by the absorption of light. In the case of phosphorescent substances, which will be discussed in a later article, we are concerned with even longer lifetimes of the excited states.

TECHNICAL APPLICATIONS OF SECONDARY EMISSIONS

by J. L. H. JONKER and M. C. TEVES. 537.533.8 : 621.385.15 : 621.385.833

Several types of valves are described in which the electron current is amplified by secondary emission: the dynatron, electron multipliers and hot cathode amplifier valves.

The phenomenon of secondary emission, which was described in detail in a previous number of this periodical[1]), was discovered about 35 years ago. Since its discovery means have been sought of controlling the phenomenon. In cases where it arises spontaneously in electron tubes and is undesired, attempts are made to avoid it or at least to nullify its effect. On the other hand attempts have been made to design tubes in which this new source of electrons should be technically useful. Some of the tubes which have found a certain degree of technical application will be treated in this article.

The dynatron

The first technical application of secondary electron emission was indicated by Hull in 1915. He designed a valve, called a dynatron, in which secondary emission was used to obtain a negative resistance. The action of this valve is as follows (*fig. 1*). Part of the electrons leaving the cathode k arrive on the positive electrode a, while part of them pass through the opening of this pierced electrode and reach the secondary emitting plate p, which is also at a positive potential but lower than a. The secondary electrons freed from the plate are captured

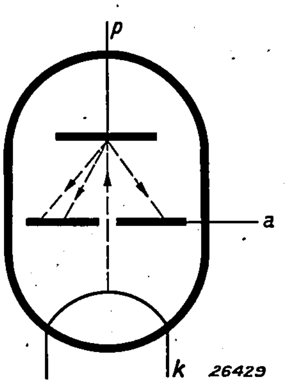

Fig. 1. Diagram of the dynatron.

by the electrode a. If I_{prim} is the primary current to the plate, δ the coefficient of secondary emission, then the total current to the plate is

$$I_p = I_{prim}\ (1 - \delta).$$

Since the coefficient δ may be considerably greater than one (for instance 5), the current to the plate will in many cases become negative.

Fig. 2 gives an illustration of the current I_p

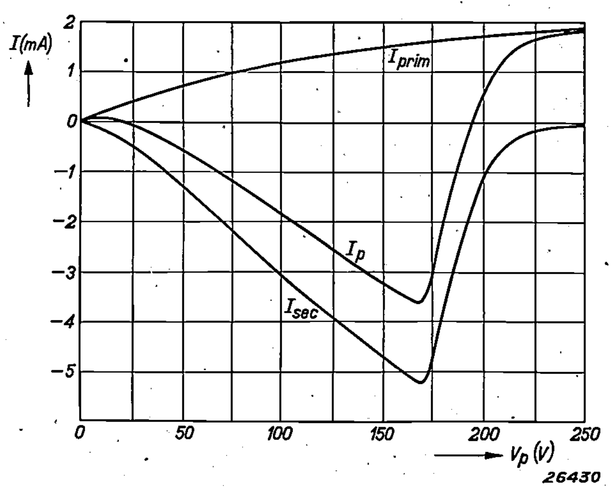

Fig. 2. Variation of primary (I_{prim}), secondary (I_{sec}) and resultant (I_p) electron currents to the plate as functions of the plate voltage V_p at constant anode voltage V_a.

as a function of the voltage V_p of the plate. The voltage of the electrode a is here kept constant at $V_a = 180$ volts. At low values of V_p only a few secondary electrons are formed per primary electron. The coefficient of secondary emission increases, however, with the voltage, and becomes greater than unity when $V_p > 20$ volts. At higher voltages than this, therefore, the plate current is negative. With increasing voltage the plate current becomes steadily more negative, due to the increase in δ. In this region, therefore, the valve has a negative resistance, i.e., $\mathrm{d}V_p/\mathrm{d}I_p < 0$. When, however, V_p becomes greater than V_a, the current of secondary electrons begins suddenly to decrease, since now the electrons must run counter to a potential difference in order to leave the plate. The plate current consequently increases strongly, and for $V_p = 250$ volts it is practically equal to the primary current.

The most important property of the dynatron is that, because of its negative resistance, it offers a simple method of exciting electrical oscillations. If there is a negative resistance in parallel with an oscillating circuit, and if this resistance is smaller in absolute value than the positive equivalent resistance of the circuit itself, the damping of the

1) H. Bruining, Philips techn. Rev. 3, 80, 1938.

combination becomes negative; that is, an oscillation excited in the circuit continues to increase in amplitude. Actually this increase is limited because the internal resistance of the dynatron is only negative in a limited voltage interval. If the oscillation takes place in the linear portion of the descending characteristic, and, therefore, at not all too great an amplitude, a practically sinusoidal oscillation can be generated.

Oscillations of very high frequencies can be excited by means of the dynatron. This shows that secondary emission is a phenomenon which takes place very rapidly. An observable lag between the incidence of the primary electron and the emission of secondary electrons does not appear under normal conditions; the difference in time is certainly shorter than the transition time of the secondary electrons from the plate to the pierced electrode, which time is usually less than 10^{-8} sec.

The most obvious advantage of the dynatron over the familiar oscillator valve is the absence of a reaction coil.

For high powers with high working voltages the dynatron is less suitable as a generator, since the coefficient δ decreases again at high voltages.

Electron multipliers

One primary electron may, as we have seen, free a number of secondary electrons. Use is made of this multiplication in the electron multiplier for amplifying a current. A very high amplification can be attained in this way by employing secondary emission several times successively. A primary electron current, emitted by an illuminated *photo-cathode*, is focussed by means of suitably chosen electric fields or by a favourable combination of electric and magnetic fields, on an electrode which has a surface with a high capacity for secondary emission. The secondary electrons from this electrode, which are already a multiple of the primary electrons, are focussed in the same way on a following electrode where multiplication again takes place. This process may in principle be repeated any number of times; the limit is practically given finally by the permissible load on the last electrode or the final anode of the system, or by possible space charges.

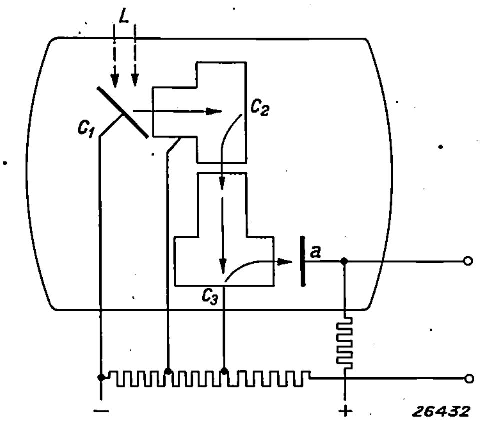

Fig. 4. *T*-electron multiplier. Explanation as in fig. 3.

There are various possible ways in which the electrons in a multiplier may be directed to the succeeding secondary emitting surface in each case. Zworykin has worked out several models, in which only electric fields are used (*figs. 3* and *4*). Only a few successive steps can be applied, however, because of difficulties in keeping the beam of electrons concentrated. The voltages of the successive electrodes are taken from a common potentiometer.

The tube constructed by Weiss uses also only electric fields. In this tube a number of wire gauzes with a high secondary emission (*fig. 5*) are connected in series with from left to right successively higher positive potentials. The electrons which strike a wire of the gauze free secondary electrons which are attracted to a following gauze, etc. The electrons which pass through a mesh are not multi-

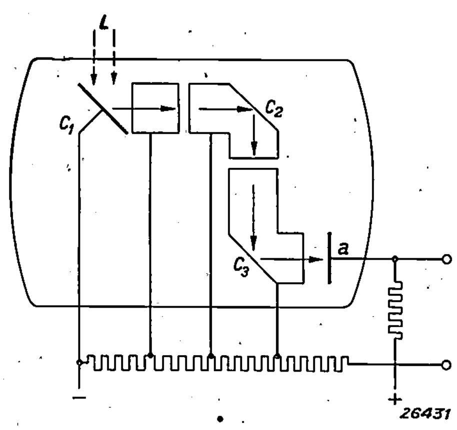

Fig. 3. *L*-electron multiplier, c_1 photo-cathode, c_2, c_3 secondary cathodes and *a* final anode.
The cylindrical tubes serve as focussing electron lenses, the arrows indicate approximately the path of the electrons.

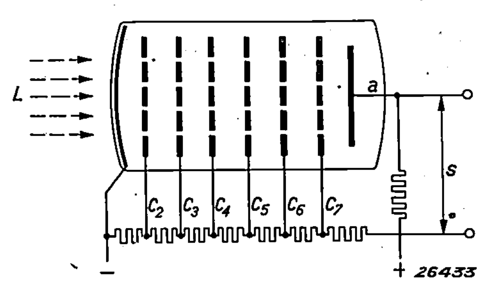

Fig. 5. Wire-gauze electron multiplier with transparent photo-cathode, c_2 ... c_7 wire gauzes which serve as secondary cathodes, *a* final anode.

plied, and the amplification is, therefore, relatively low. In addition a transparent photocathode is used here and the light is incident from one side while the electrons are emitted from the other.

The most satisfactory construction, and one which permits a large number of amplification stages, is one in which the electrons are focussed on the successive secondary cathodes by a transverse magnetic field (*fig. 6*). The primary and all the secondary cathodes (c_1, c_2 ...) are here mounted in a straight line in one plane. In a second plane, parallel to the first there is an equal number of anodes (a_1, a_2 ...). Each cathode is connected to the foregoing anode so that upon application of the voltage the field on the cathodes is sufficiently high to remove the electrons from it. The transverse magnetic field H of adequate intensity causes the accelerated electrons to describe almost semicircular paths. The electric and magnetic fields are so adjusted relative to each other that the electrons just reach the following cathode. The secondary electrons freed there, a multiple of the incident electrons, are directed upon the following cathode in the same way, and so on.

The magnetic field must be carefully adapted to the accelerating electric field between the electrodes, for instance by regulation of the excitation current of the magnet. Moreover, the magnetic field must be homogeneous over the whole length of the electrode system, and it must be accurately directed, that is perpendicularly to the longitudinal axis of the system and parallel with the cathode and anode planes (see fig. 6). When these conditions are satisfied, tubes with ten or more successive stages may be constructed. With a voltage of 200 volts between the successive cathodes and their corresponding anodes, and when caesium oxide is used as secondary emitting surface, an amplification of the current by a factor of 5 per stage is easily attained. For a multiplier of this type with 10 stages this means an amplification of 5^{10} or about 10 million.

In a multiplier there is not only a photo-cathode but also secondary emitting cathodes. If different kinds of surfaces were used for these cathodes, there would be a danger that they might have undesirable effects upon each other during preparation. It is, therefore, advantageous to use the same kind of surface for both purposes. At first sight it seems difficult to find a suitable surface, since quite different requirements are made of a surface which is to have a high secondary emission from those made of a surface which must have a high photosensitivity. There is, however, such a surface, and it is used in the multipliers here described. A layer of silver is covered with caesium oxide and adsorbed caesium. The photo-sensitivity of this cathode is due to the adsorbed caesium atoms which can easily be ionized by light. The capacity for emitting secondary electrons is due to the caesium oxide; the adsorbed caesium atoms are, therefore, not essential for the secondary emission. It has been found that this cathode has a higher secondary without a covering of caesium atoms than with such a covering. With the covering of caesium atoms the photo-sensitivity is at a maximum.

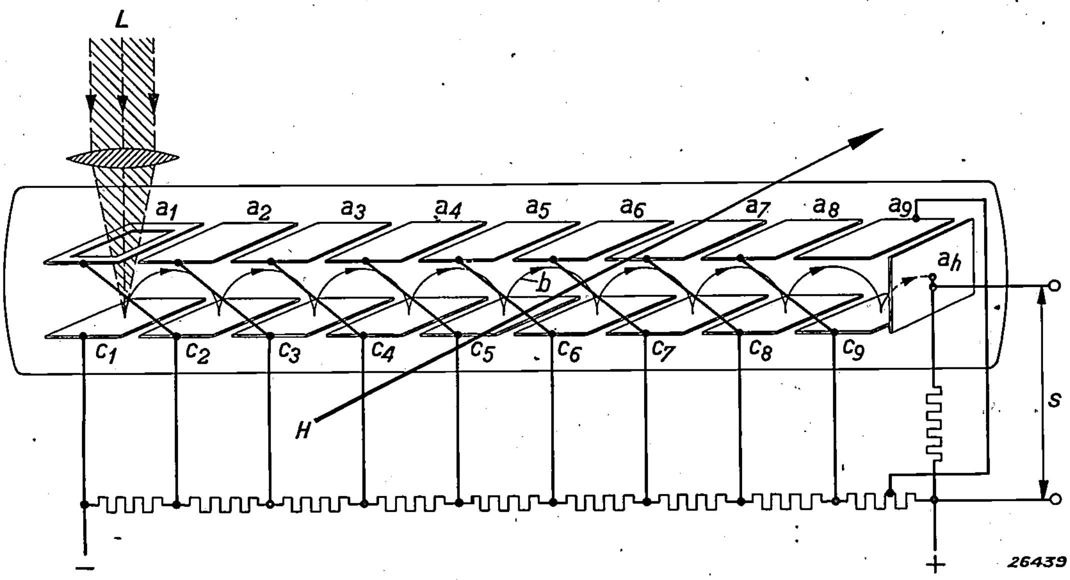

Fig. 6. Electron multiplier with magnetic focussing.
c_1 photo-cathode, c_2, c_3 ... secondary cathodes, a_1, a_2 ... auxiliary anodes and a_n final anode
→ direction of the transverse magnetic field H.

The great advantage of electron multipliers is that they can amplify very rapid current variations without deformation, as long as the variations do not occur in a shorter time than 10^{-8} sec. The resistance and self-induction of the leads and the capacity between the electrodes have practically no influence on the multiplication factor, even at high frequencies. The current at the final anode exhibits a time difference with respect to the photo-current which is independent of frequency, but this gives practically no difficulties.

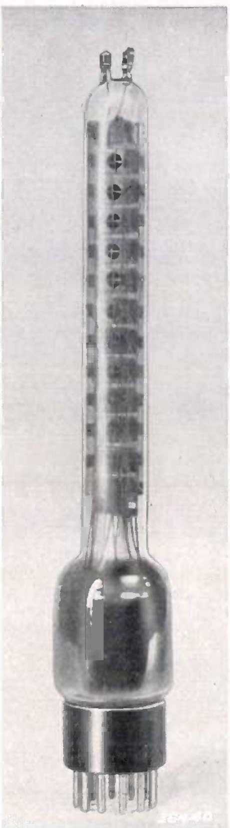

Fig. 7. Electron multiplier type 4698, constructed according to the scheme of fig. 6 with eleven stages.

The electron multiplier is, therefore, an ideal amplifier for photocurrents of all frequencies, and even for direct current. However, because of the complications inherent in the use of such a tube the application is restricted to the region of very small amounts of light, as in television with mechanical scanning [2]), astronomical measurements, spectrophotometers, etc. where we are concerned with photoelectric currents of the order of magnitude of from 10^{-14} to 10^{-16} A. In such cases with normal amplification by means of amplifier valves the noise due to the first coupling resistance (thermal motion of the electrons) and to the relatively high cathode current in the first amplifier valve (shot effect) is predominant [2, 3]). These noises are of course present even without a photo-current. In the multiplier practically the only interference is the inevitable shot effect of the photo-current itself, and this is proportional to the square root of the photo-current.

The use of the multiplier offers no advantage when the photo-current becomes greater than about 10^{-8} A (for constant illumination) or about 10^{-7} A (for flickering light, of audio-frequency for instance). At currents higher than these values it is simpler to use ordinary amplifiers.

Fig. 7 is a photograph of an electron multiplier with eleven stages of amplification by secondary emission (type 4698).

Secondary emission in amplifier valves with hot cathodes

Just as with the current in a photocell, attempts have been made to amplify the electron current in an amplifier valve by means of secondary emission.

The realisation of this concept, shown diagrammatically in *fig. 8*, was prevented until recently by practical difficulties. If a hot cathode and a surface which in itself has a high and satisfactorily constant secondary emission are both included in one bulb, the properties of the secondary emitting surface may be adversely affected by the substances evaporated from the cathode (barium, barium oxide) and condensed on the secondary emitting surface. This alteration in the surface will as a rule mean a decrease of the secondary emission, and will have an unfavourable influence on the amplification of the valve.

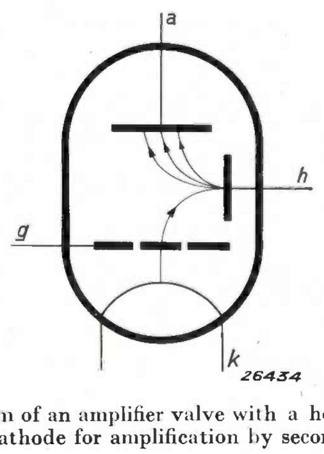

Fig. 8. Diagram of an amplifier valve with a hot cathode and an auxiliary cathode for amplification by secondary emission.

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[2]) H. Rinia and C. Dorsman, Television system with Nipkow disc. Philips techn. Rev. 2, 72, 1937.

[3]) M. Ziegler, Noise in amplifiers contributed by the valves, Philips techn. Rev. 2, 329, 1937.

In the electron multipliers described above it was possible in the way there described to overcome a similar difficulty by the use of a material which is suitable not only for photoelectric but also for secondary emission. The idea of covering the hot cathode and the plate with the same kind of surface would occur here also, but a material suitable for both purposes has not been found. Another solution had to be sought.

This solution was based on the fact that molecules evaporating from a surface in a high vacuum move in practically straight lines, so that they will not strike the secondary emitter if the latter is "out of sight" of the hot cathode. The primary electrons emitted from the cathode must then be conducted to the secondary emitting plate by a roundabout way. This can be done by means of electric and magnetic fields in a way analogous to that discussed under the various systems of electron multipliers.

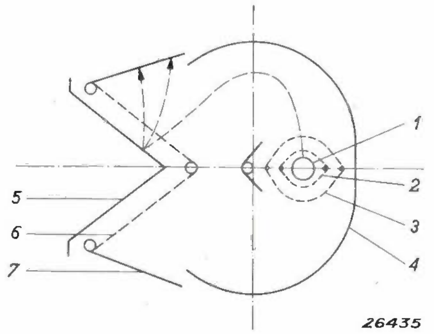

Fig. 9. Cross section of amplifier valve type 4696. *1* cathode, *2* control grid, *3* screen grid, *4* screen at cathode potential, *5* auxiliary cathode, *6* anode grid and *7* anode.

For amplifier valves the use of an external magnetic field is not so simple, and the adjustment of that field is very critical. Electrostatic deflection is in this case greatly to be preferred to magnetic deflection. Electrostatic deflection is possible for example by repulsion by an electrode at a low potential or by attraction by an electrode at a high potential. This electrode may, when its shape is suitably chosen, also bring about a concentration of the electron beam.

The interior of an amplifier valve with one stage of secondary emission, type 4696, is shown in the two accompanying figures. *Fig. 9* shows a cross section of the system of electrodes; *1* is the cathode, *2* the control grid and *3* the screen grid. The electrons pass through the screen grid *3* and are as it were reflected by the curved screen *4* and concentrated on the auxiliary cathode *5*. The surface of the latter has a high secondary emission and is at such a potential (150 volts) that a good factor, $\delta = 5$, for secondary emission is obtained. The secondary electrons are attracted away from the cathode by the grid *6*, which is at anode potential (250 volts), and move toward the anode *7*. *Fig. 10* is a photograph of the construction with

Fig. 10. Photograph of the interior of valve type 4696.

the screen cut away to show the interior. To the right may be seen the cathode with the surrounding grids, to the left the auxilary cathode, the anode grid and the anode plate. Finally *fig. 11* shows the curvature of the equipotential planes in a valve constructed as described above. This diagram was recorded with an enlarged model in an electrolytic trough [4]). Between screen grid and auxiliary cathode the electrons pass through two concentrated potential fields, while a kind of reflection takes place at

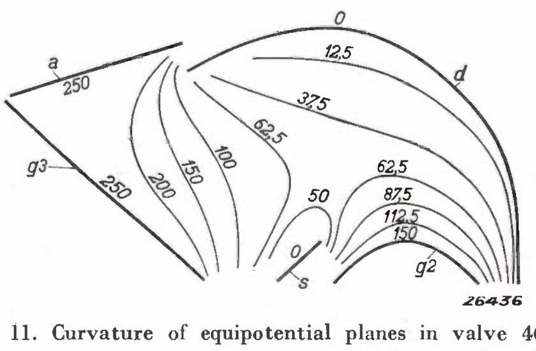

Fig. 11. Curvature of equipotential planes in valve 4696.

[4]) In this method an enlarged model of the construction to be investigated is placed in a container filled with an electrolyte. When the electrodes are given the required potentials current flows between them, and the potential can be measured at each point with the aid of a probe electrode.

the region of low potential caused by the curved screens.

Since in this amplifier valve one is concerned with currents of a few milliamperes, considerable space charges may appear, which would not only deform the field in the neighbourhood of the screen, but also at the surface of the auxiliary cathode, and thereby prevent part of the secondary electrons from reaching the anode. This is avoided by the introduction of the grid *6* at anode potential directly in front of the auxiliary cathode *5*. If this measure is not taken the anode current becomes quite closely dependent on the anode voltage, in other words, the internal differential resistance of the valve falls, which is undesirable for amplification purposes.

The great advantage of this secondary emission valve lies in the fact that, not only the anode current, but also the slope $\mathrm{d}I_a/\mathrm{d}V_g$ is increased by about a factor of 5 as a result of amplification by secondary emission. Particularly when the external anode impedance may only be very low, as in the amplification of wide frequency bands (television), this steepness of slope is very welcome.

If two valves are compared, one with and one without amplification by secondary emission, it is found that with the same anode current the slope of a valve with secondary emission is considerably greater. For example in *fig. 12* let the variation of

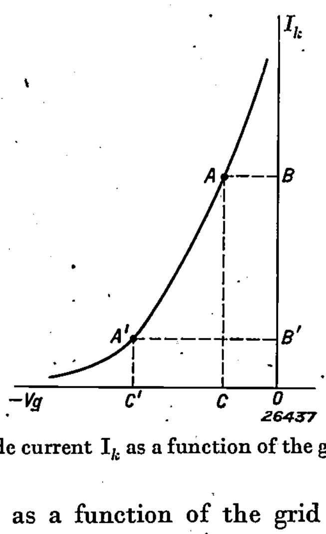

Fig. 12. Cathode current I_k as a function of the grid voltage V_g.

I_k be given as a function of the grid voltage V_g. At a given grid voltage OC without amplification by secondary emission the cathode current I_k as well as the anode current I_a is equal to OB. If an auxiliary cathode for secondary emission is now introduced, then, when I_a is kept constant, the current I_k will become smaller, since $I_a = \delta \cdot I_k$. δ is here the ratio of the electron current emitted from to that incident on the auxiliary cathode. A factor $\delta = 5$ can be attained with the difference of potential of 150 volts between cathode and auxiliary cathode, so that the necessary current V_k falls to about $^1/_5$ of its former value (OB' in fig. 12). At the point A' on the characteristic $I_k = f(V_g)$, the slope $\mathrm{d}I_k/\mathrm{d}V_g$ is, however, greater than $^1/_5$ of that at point A; the slope of the valve with secondary emission, $\mathrm{d}I_a/\mathrm{d}V_g = \delta \cdot \mathrm{d}I_k/\mathrm{d}V_g$, is therefore greater. The following calculation will make this clearer.

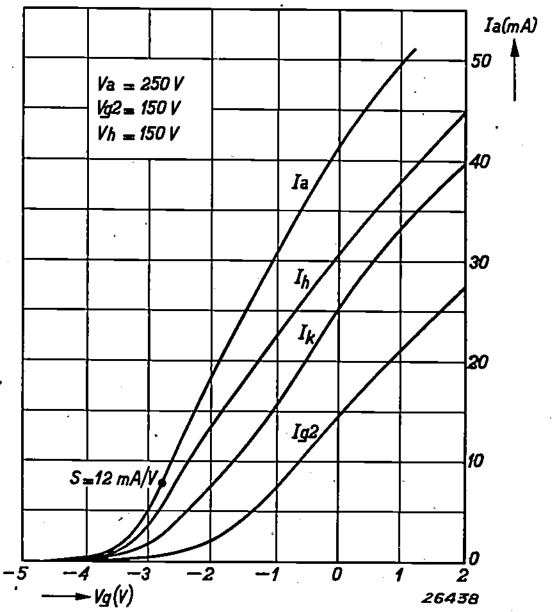

Fig. 13. Currents to the various electrodes of valve type 4696 as functions of the grid voltage V_g. I_a is the anode current, I_k is the cathode current, I_h is the current to the auxiliary cathode, I_{g_2} is the current to the screen grid.

For the cathode current I_k the following formula holds:

$$I_k = A\ (V_g + b)^K,$$

A and b are determined by the construction and the anode potential V_a. Without secondary emission the slope is:

$$\frac{\mathrm{d}I_k}{\mathrm{d}V_g} = A\,K\,(V_g + b)^{K-1} = A^{\frac{1}{K}}\,K I_k^{\frac{K-1}{K}}.$$

With secondary emission:

$$\frac{\mathrm{d}I_a}{\mathrm{d}V_g} = \delta^{\frac{1}{K}}\,A^{\frac{1}{K}}\,K\,I_a^{\frac{K-1}{K}}$$

For *the same* anode current, therefore, the slope is greater in the second case by a factor $\delta^{1/K}$. Since K is 1.6 for not too small currents, one finds for $\delta = 5$ that the slope has become $5^{0.6} = 2.6$ times greater.

Finally *fig. 13* gives the currents to the various electrodes of a valve of the above-described construction as functions of the voltage of the control grid. With this valve one stage of amplification by secondary emission has been employed and a slope of from 12 to 14 mA/V has been reached at an anode current of 8 mA, with a cathode current of only 2.5 mA, of which about 2 mA go to the auxiliary cathode *5*, and 0.5 mA to the screen grid (fig. 9). In principle is is possible to employ more stages, when the construction of course becomes more complicated and the required voltage V_a higher. For these reasons only one stage has been used in this amplifier valve.

AUDITORIUM ACOUSTICS AND INTELLIGIBILITY

by R. VERMEULEN.

534.843.5

The phenomena which influence the intelligibility of speech in an auditorium are discussed. Starting from knowledge gained by experience about intelligibility as a function of the intensity of the signal and of the noise, it is shown that in a large room with much reverberation the spoken word is only intelligible at certain places. Details are given of the contribution to intelligibility by the direct and the reflected sound. From the assumption that only those waves which reach the listener within 1/15 of a second after the direct sound contribute effectively to the intelligibility, are derived the principles for the design of an auditorium with good acoustics.

An architect who is designing a theatre auditorium will not fail to provide an unhindered view of the stage from every seat. There is, however, some doubt whether he always makes as good a provision for the intelligibility of the spoken word, which is certainly equally important in the enjoyment of a play. It is not sufficient simply to provide for the proper reverberation time in an auditorium, the significance of which was explained in a previous article [1]), because the intelligibility, especially of the spoken word, is determined also by other factors. Although the reverberation time is the same at all points in the hall as required by Sabine's law, nevertheless it may be more difficult to follow speech at one point than at another, more favourable position. It is mainly due to the great importance of intelligibility of speech for the development of telephone communication, that many experiments have been carried out in the last few decades on the dependence of intelligibility on all kinds of disturbing and deforming effects [2]).

Factors of influence on intelligibility

It will only be possible here to give several results. In the first place intelligibility is to a high degree dependent on intensity. In order to express this fact in figures it must be possible to measure intelligibility. This can be done by having a large number of sentences, words or sylables spoken and finding out what percentage is understood correctly. If syllables with no relation or meaning are used (nonsense syllables), the percentage correctly understood, which we shall call "intelligibility", will of course be much smaller than when a connected sentence is spoken, since in the latter case the words which are less well heard will be automatically supplied. The relation between this "comprehensibility" and "intelligibility" is shown in *fig. 1*.

If the intelligibility is determined at different sound intensities of the speech, without the listener being disturbed by reverberation and foreign sounds, the ratio given in *fig. 2* at 0 is obtained. The other

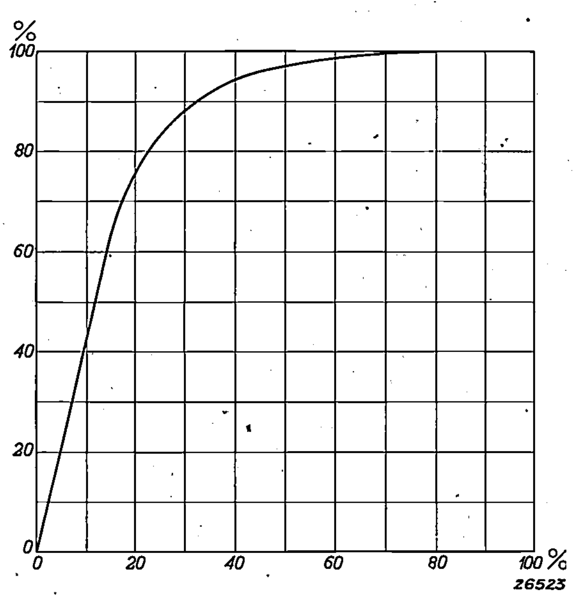

Fig. 1. Abscissa: intelligibility of nonsense syllables, ordinate: percentage of simple questions answered correctly. This curve shows that an intelligibility of 50% is enough to make it possible to follow an argument.

1) A. Th. van Urk, Auditorium acoustics and Reverberation, Philips techn. Rev. 3, 65, 1938.

2) Good summaries may be found in the following books: Harvey Fletcher, Speech and Hearing: F. Trendelenburg. Klänge und Geräusche.

secondary emission has been employed and a slope of from 12 to 14 mA/V has been reached at an anode current of 8 mA, with a cathode current of only 2.5 mA, of which about 2 mA go to the auxiliary cathode *5*, and 0.5 mA to the screen grid (fig. 9). In principle is is possible to employ more stages, when the construction of course becomes more complicated and the required voltage V_a higher. For these reasons only one stage has been used in this amplifier valve.

AUDITORIUM ACOUSTICS AND INTELLIGIBILITY

by R. VERMEULEN.

534.843.5

The phenomena which influence the intelligibility of speech in an auditorium are discussed. Starting from knowledge gained by experience about intelligibility as a function of the intensity of the signal and of the noise, it is shown that in a large room with much reverberation the spoken word is only intelligible at certain places. Details are given of the contribution to intelligibility by the direct and the reflected sound. From the assumption that only those waves which reach the listener within 1/15 of a second after the direct sound contribute effectively to the intelligibility, are derived the principles for the design of an auditorium with good acoustics.

An architect who is designing a theatre auditorium will not fail to provide an unhindered view of the stage from every seat. There is, however, some doubt whether he always makes as good a provision for the intelligibility of the spoken word, which is certainly equally important in the enjoyment of a play. It is not sufficient simply to provide for the proper reverberation time in an auditorium, the significance of which was explained in a previous article [1]), because the intelligibility, especially of the spoken word, is determined also by other factors. Although the reverberation time is the same at all points in the hall as required by Sabine's law, nevertheless it may be more difficult to follow speech at one point than at another, more favourable position. It is mainly due to the great importance of intelligibility of speech for the development of telephone communication, that many experiments have been carried out in the last few decades on the dependence of intelligibility on all kinds of disturbing and deforming effects [2]).

Factors of influence on intelligibility

It will only be possible here to give several results. In the first place intelligibility is to a high degree dependent on intensity. In order to express this fact in figures it must be possible to measure intelligibility. This can be done by having a large number of sentences, words or sylables spoken and finding out what percentage is understood correctly. If syllables with no relation or meaning are used (nonsense syllables), the percentage correctly understood, which we shall call "intelligibility", will of course be much smaller than when a connected sentence is spoken, since in the latter case the words which are less well heard will be automatically supplied. The relation between this "comprehensibility" and "intelligibility" is shown in *fig. 1*.

If the intelligibility is determined at different sound intensities of the speech, without the listener being disturbed by reverberation and foreign sounds, the ratio given in *fig. 2* at 0 is obtained. The other

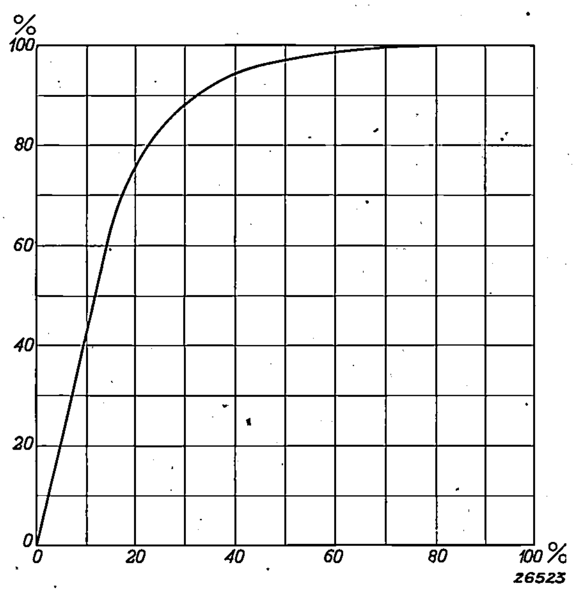

Fig. 1. Abscissa: intelligibility of nonsense syllables, ordinate: percentage of simple questions answered correctly. This curve shows that an intelligibility of 50% is enough to make it possible to follow an argument.

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curves in this figure refer to cases where the listener is disturbed by noise of different intensity. While the nature of the noise is not immaterial, the character of the curve remains practically unaltered. Considering that the curves are practically parallel, the conclusion may be drawn that the influence of noise corresponds to a decrease in the intensity level of the speech, or in the other words, to an increase of the threshold of the ear. It may be seen from the figure that intelligibility at equal intensities of noise and speech is about 50 per cent, which may be considered sufficient for following an argument, although an intelligibility of 70 per cent must be required for good acoustics (see fig. 1).

A second factor which may affect intelligibility, and which is especially important when amplifier installations with microphones and loudspeakers are to be used for transmission or amplification, is the change in the distribution of sound energy over the various frequency ranges. Although the fundamental tones of the voice lie in the neighbourhood of 125 cycles/sec (male) to 250 cycles/sec (female), the range from 1000 to 2000 cycles/sec is the most important from the standpoint of intelligibility. This follows from experiments whose results are given in *fig. 3*. In this figure intelligibility is plotted as a function of the limiting frequency of a filter which only passes those tones of the voice which lie below (L = low), or above (H = high) that limiting frequency. The inscription under the figure gives further details. It may be concluded from the figure that an intelligibility of 60 per cent can be obtained even though all the tones below 1300 cycles/sec are missing from the speech. That the quality of the spoken word is thereby very much altered, is obvious.

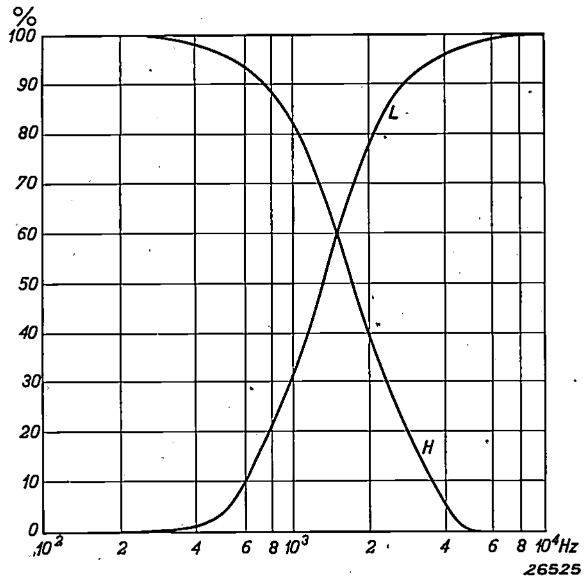


Fig. 3. The percentage of nonsense syllables understood correctly by the listener plotted as a function of the limiting frequency of a filter included in the circuit of the sound producer. Curve L was determined with a filter which passes the tones below the limiting frequency so that the influence of the lack of higher frequencies is shown here. In the absence of frequencies above 2000 cycles/sec, 75% of the syllables were still correctly understood. Curve H is determined with a filter passing the high frequencies: from this curve it may be deduced that when all tones below 1000 cycles/sec are lacking the intelligibility is still above 80%.

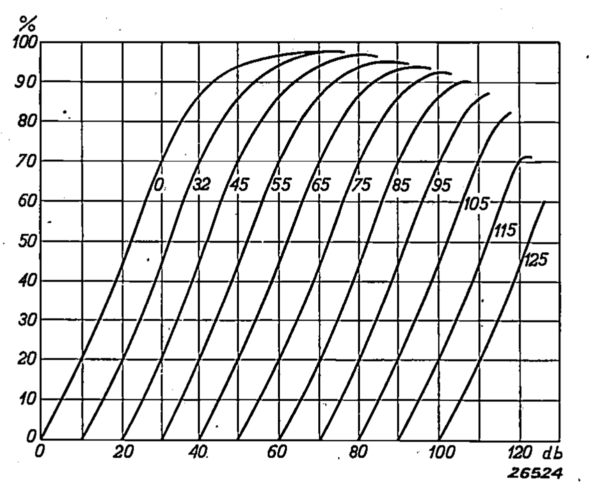


Fig. 2. The percentage of nonsense syllables plotted as a function of the level of the spoken word above that of the threshold value and that of the noise. The successive curves were recorded in each case with an increase in noise such that the threshold value increased by 10 decibels. It may be seen that the curves are practically parallel and that the threshold value lies constantly 25 decibels below the level of the noise when the latter is not greater than 40 decibels.

Reverberation as a disturbing factor

Noise which decreases the intelligibility of the spoken word does not consist only of that caused by the audience, and such other sounds as the hum of fans and street noises, etc. In addition, the sound of previously spoken words which remains in the auditorium as reverberation disturbs the listener as he tries to understand the new words.

Even if it is possible, by increasing the strength of the voice, and, to a greater degree, by the use of sound amplifying apparatus, to raise the level of the spoken word sufficiently above that of the intruding noise, these methods fail when we are concerned with the reverberation of that spoken word itself: the intensity of the disturbing reverberation will increase in proportion to the direct sound. The result is that a speaker in a hall with bad reverberation can only make himself understood over a limited area of surface which cannot be increased by the use of loudspeakers.

Without going into too much detail a rough calculation will demonstrate this fact. When a number of sources of sound with a total power W are set up in a hall whose walls have an absorption equivalent to an open window of area A, the average energy density of the reverberation is:

$$E_m = \frac{4W}{cA}, \quad \ldots \ldots \ldots \quad (1)$$

where c is the speed of propagation of sound (van Urk, loc.cit.). The density of energy of the direct sound from the source indicated by the subscript i, at a distance r_i from the source,

$$E_d = \frac{W_i}{4\pi r_i^2 c} \quad \ldots \ldots \ldots \quad (2)$$

From these two formulae it follows that only over a limited distance will the intensity of the direct sound be greater than that of the disturbing reverberation as was found necessary for reasonably good intelligibility (50 per cent). This maximum distance is given by

$$r_i^2 = \frac{W_i}{W} \cdot \frac{A}{16\pi} \quad \ldots \ldots \ldots \quad (3)$$

If, to choose the most favourable situation, it is assumed that each source of sound is open to listeners on all sides, the area where the sound is intelligible is then:

$$\sigma = \pi \Sigma r_i^2 = \frac{\Sigma W_i}{W} \frac{A}{16} = \frac{A}{16} \quad \ldots \quad (4)$$

Equation (4) may be used to describe the situation in large rooms with much reverberation where only a small part of the space is occupied by an audience. It would for instance be possible to calculate by means of this equation, over how large a part of an indoor swimming pool the results of a competition could be announced at the same moment.

If, however, the audience occupies a large part of the area, the intelligibility area can be increased by making use of the fact that the absorption coefficient of the audience is nearly 100 per cent. This obviously completely nullifies the basis of Sabine's reverberation theory. The sound which reaches the audience directly will not be reflected and will not contribute to reverberation. The rest of the sound will give rise to reverberation whose average energy density is given by equation (1). If p is the part of the sound which reaches the audience directly, the total power of the sources of sound which contribute to reverberation is $W(1-p)$, and equation (1) must now be replaced by

$$E_m = \frac{4W}{cA}(1-p) \quad \ldots \ldots \ldots \quad (1a)$$

The area which can be reached with intelligibility thereby becomes greater:

$$\sigma = \frac{A}{16(1-p)} \quad \ldots \ldots \ldots \quad (4a)$$

This result is important chiefly in the case when loudspeakers are used as sources of sound. In order to make p as great as possible loudspeakers with directional effect may be used, and they may be directed toward the audience. In this way, even under unfavourable circumstances, an adequate intelligibility can be reached over the whole area occupied.

Influence of the growth of sound

The above considerations are based on the assumption that only the direct sound contributes to intelligibility, and that all the reflected sound must be considered as noise. This assumption will approach the actual situation in very large closed spaces; it does not, however, give a correct picture of the acoustic relations in a smaller room.

The truth of this will be seen when the intelligibility in a room is compared with that which can be attained out of doors. The direct sound is equally intense in both cases, but indoors reverberation is present as a disturbing factor. One might therefore expect that the voice of a speaker in a hall would only be intelligible over a shorter distance than out of doors.

Experience, however, shows that the opposite is true. Although at the rear of a hall one direct sound often supplies only a very small part of the total intensity, a speaker can usually be understood much farther away in a closed room than out of doors. It is therefore apparent that sound reflected by the walls and ceiling is, at least partially, intelligible, and is not experienced as noise. The size of this part will depend upon the way in which the sound grows at the beginning of each new sound.

It will be a surprise to many to learn that when a constant source of sound is turned on the sound intensity grows in a way analogous to the way in which it dies out after the source has been interrupted. This is surprising, because while the echoing of a voice in a large church is a common experience, one can not recall having observed a gradual growth. A concise, although not entirely correct representation of this difference in the observation of reverberation and growth, is obtained by assuming that the sound impression is proportional to the logarithm of the intensity, which comes down to this, that the ear judges equal percentages of change in intensity as equal differences.

In *fig. 4b* the growth and decay of sound in a room is shown diagrammatically, with a linear scale of intensity for the variation of power radiated according to fig. 4*a*. The intensity I_g during growth

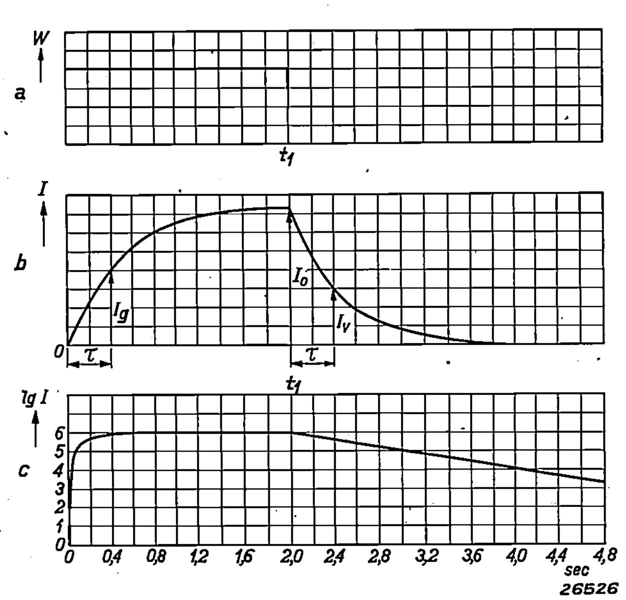


Fig. 4. Diagrammatic representation of the growth and decay of sound when a source of constant strength is switched on at $t = 0$ and stopped at $t = t_1$. a) Radiated power of the source as a function of the time. b) Intensity in the room as a function of time. The growth is the complement of the reverberation (the decay), i.e. $I_g + I_v = I_0$. c) Logarithm of the intensity as a function of time. The last curve represents approximately the subjective impression, and shows that growth seems to be of shorter duration than reverberation.

is here described as a function of the time by

$$I_g = I_0\,(1 - e^{-at}),$$

while the decay takes place according to the formula

$$I_v = I_0\ e^{-a(t-t_1)}$$

(see van Urk, loc.cit.). The growth and reverberation are complementary, i.e., the intensities I_g and I_v measured at the same time τ after switching on and off respectively satisfy the relation $I_g + I_v = I_0$, which is understandable when one considers the interruption of the source as the addition of a second equally intense source with a negative sign.

If the same variation is now drawn on a logarithmic scale of intensity (fig. 4*c*), the difference between growth and reverberation is clearly seen: growth takes place suddenly and is completed in a fairly short time, reverberation proceeds slowly and uniformly. Long after the intensity (fig. 4*b*) could no longer be drawn, the reverberation is still above the threshold value.

It is not easy to represent in numbers the influence of growth on intelligibility because it is quite different for different sounds. By speaking slowly the vowels can be so much drawn out that they always have the opportunity of growing to sufficient intensity. Since during that time the reverberation of the previous sound has completely disappeared, the speaker will always be able to reproduce the vowels without disturbance and with sufficient intensity. This is the basis of the preaching tone usual in churches.

The consonants, and especially b, p, d, t, g, k, etc. cannot be drawn out by speaking slowly. They cannot, therefore, on the one hand reach their full development, and, on the other hand, they suffer interference from the previous sounds. For these letters, therefore, a gradual growth of sound is permissible only to a much smaller degree than for the vowels.

In order to find out to what degree reflections from the walls may produce an increase in intelligibility even for such short sounds, an approximation of growth by an exponential curve is not sufficient. Actually the intensity varies by jumps which are connected with the arrival of the waves reflected by the various walls, and this variation will differ essentially for example at the back of an auditorium from that in the neighbourhood of the speaker.

In order to make this clear, in *fig.* 5 the propagation of a sound wave is followed step by step in an imaginary room. The 15 diagrams represent successive instantaneous states at intervals of $^1/_{150}$ sec. The waves start from a source of sound A, and it is obvious that the front B of the "direct" wave is the first sound to reach every listener. The intensity of this wave decreases with the square of the distance from the source. Actually the direct sound will be even more weakened in many cases, since it passes over an audience which absorbs sound strongly. In such a case it may be assumed that the sound pressure, which is the determining factor in observation, is given by the current of sound energy which is incident upon one unit of area of the absorbing surface, and not by the intensity of the sound wave which would exist at that point in the absence of an absorbing surface. The sound pressure will therefore usually be very small, because the direct sound from a speaker or an orchestra nearly always travels in a practically horizontal direction, and a given area therefore receives energy from the source only within a very small solid angle. Only with loudspeakers, or by placing the speaker on a high projecting rostrum, is it possible to direct the sound toward the audience at a steep angle. Without these aids very little of the direct sound reaches the back of the hall.

In fig. 5 it will be seen, however, that it is at the back of the hall, where the direct wave is so much weakened, that a large number of reflected waves with slight differences in time reach the audience.

The question now is: to what degree can these waves help to increase the intelligibility?

Useful and detrimental sound

Since the ear, like the eye, possesses the charac-

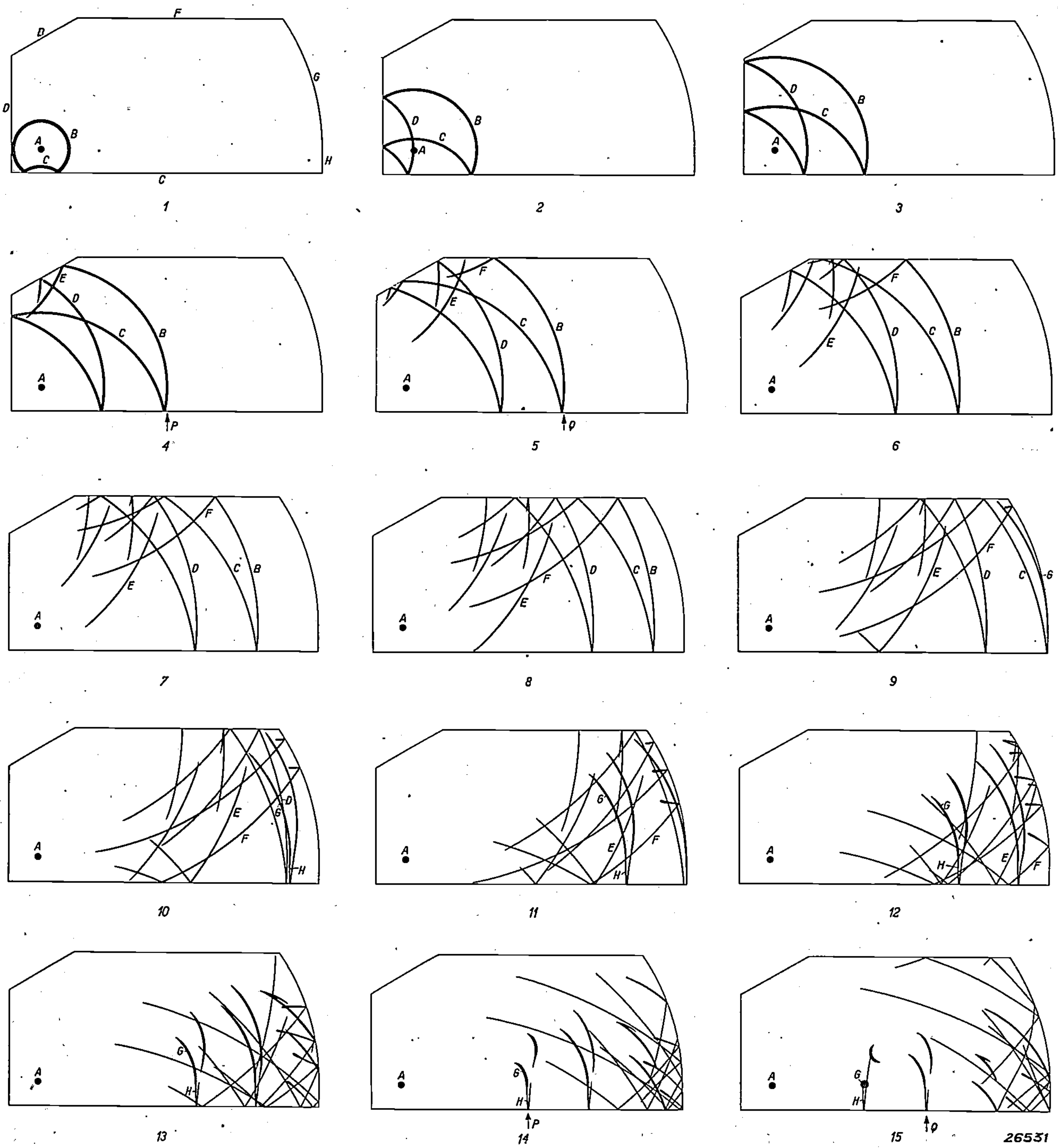


Fig. 5. Let the above figure represent the cross section of an auditorium. A source of sound is placed at A, and it begins to radiate at the moment $t = 0$. In the successive pictures the wave front is drawn at intervals of $^1/_{150}$ sec. The walls are considered to be completely reflecting, while the strength of the wave is represented approximately by the thickness of the lines. The waves are indicated by the same letters as the walls by which they are reflected. The most important of these are:

- *B:* the "direct" wave, i.e. the wave which has not been reflected by any wall;
- *C:* the reflection by the floor;
- *D:* the reflection by the wall behind the source;
- *E:* the reflection by the sloping roof above the source;
- *F:* the reflection by the ceiling;
- *G:* the reflection by the curved rear walls;
- *H:* the reflection by the vertical part of the rear wall.

Because of the complexity of the combinations of repeated reflections, it is impossible to continue this method of indication consistently.

teristic of blending stimuli received during a certain period of time to a single impression, the waves which are incident one after another can reinforce each other if their succession is rapid enough. The time within which the ear is unable to perceive successive reflections as separate from each other is of course not sharply limited, but it is of the order of $^1/_{15}$ sec. The waves in fig. 5, which reach the listener within $^1/_{15}$ sec after the first sound impression, therefore act together to increase the intelligibility: they are "useful" sound. That which comes later may be "detrimental" because of the fact that it is observed as a new impression and, as such, may be disturbing. It is now possible to find out in a simple way where "detrimental" sound first occurs in fig. 5. In the first 10 states only useful sound can reach the listener, since these 10 states occur within $^1/_{15}$ sec. In the next few states also the audience receives no "detrimental" sound, because at those places reached by reflected sound the direct wave B is incident only several hundredths of a second after the source has begun to work. From a comparison of the 4th with the 14th state and of the 5th with the 15th, it will be seen that at the points of the arrows P and Q respectively, the reflection of the rear wall is incident just $^1/_{15}$ sec after the direct sound. The wave H, which up to this moment as useful sound could make a contribution to the intelligibility, now begins to be detrimental. To an even greater degree is this the case with the wave G reflected by the curved rear wall and concentrated in a focus. This gives an excessively reinforced sound wave which arrives 0.08 sec later than the direct wave. It is distinguished from the other reflections which make up the reverberation by its greater intensity, and is therefore observed separately as an echo.

The analysis as carried out in fig. 5 is too time-consuming for practical application and, moreover, it extends only over $^1/_{10}$ of a second. It must usually suffice to find out how much energy the different parts of the audience receive from the direct beam and from the principal reflected waves which differ in length of path covered by not more than 20 metres (difference in time about $^1/_{15}$ sec). This can be done quite simply by drawing a beam of sound rays starting from a speaker at various suitable places, and following these rays through their first and sometimes second reflection, until they reach the listeners. It must hereby be kept in mind that the aim is to throw the sound as much as possible on the audience and, in addition, to distribute the available energy as uniformly as possible to prevent one listener from obtaining too much at the expense of another.

In this investigation those rays of sound are also discovered which differ in length of path by more than 20 metres from the direct sound, and which therefore, if they are sufficiently intense, can give rise to an echo.

Influence of the shape of the hall

In connection with the foregoing we shall discuss several points about the influence exerted by the shape of the hall and its walls on intelligibility. This discussion cannot be a complete one, but only an explanation of the ways in which the useful sound can be increased and the detrimental sound decreased.

Useful sound

As we have already shown, except with a raised source of sound (a loudspeaker for example), the direct sound radiation will only be of importance for the first rows of seats. The rest of the audience will have to rely upon sound reflected by the walls. The ceiling, considered as one of the walls, will fulfil the most important function in this reflection, and a plane surface will in ordinary cases be found favourable. The front part of the ceiling, however, throws the sound into the front of the hall where it is not needed, and it is therefore an advantage to make this part of the ceiling sloping as shown in fig. 5 at E. The rear wall also, especially under the balcony, becomes effective if it leans forward at a suitably chosen angle. In some cases there will be a conflict between the position of a wall desirable for other reasons and that necessary for the correct reflection of sound. This is true for instance when the ceiling slopes upward toward the rear, whereby too much sound is reflected to the extreme rear of the hall. A step-shaped design, as indicated in *fig. 6* may then often solve the difficulty. Care must

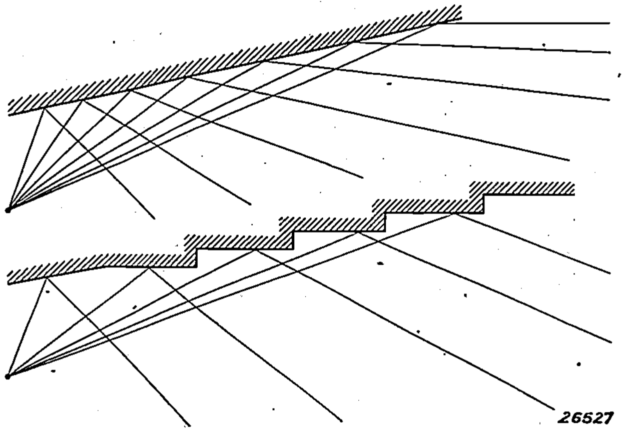


Fig. 6. If it is necessary that the ceiling of an auditorium should be higher at the rear it can to advantage be constructed in the form of steps.

be taken not to exaggerate by carrying the principle of the reflecting action of the walls too far. It will in general be satisfactory if the audience receives reflected sounds from several directions. This can, for example, be achieved by making the reflection somewhat diffuse in the transverse cross section. This is not only desirable for the more uniform distribution of the sound and for the greater independence of this distribution on the position of the source, but also in order to avoid the speaker being disturbed by noise from the public by a reversal of the ray diagram.

As a check on our calculations in the case of complicated shapes of halls, optical models can be made of the diagrams constructed. In these models,

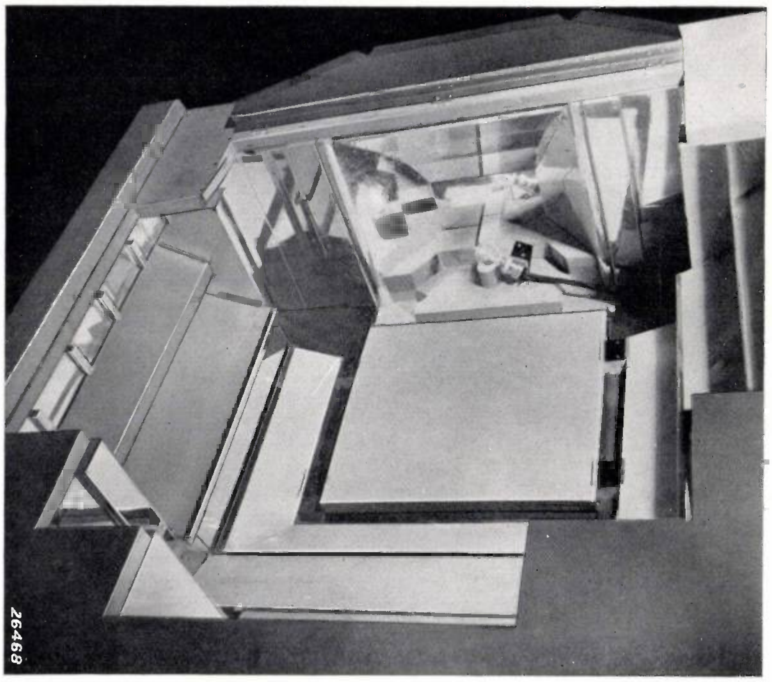
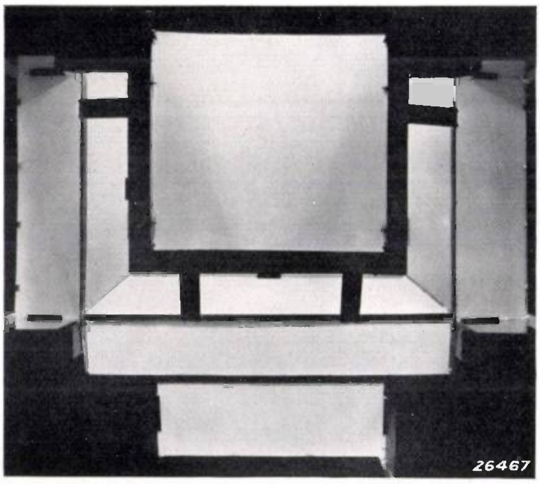


Fig. 7. Optical model of the large assembly hall in the League of Nations Palace in Geneva. The place of the speakers is represented by an electric lamp. At the intended positions of the audience, plates of frosted glass are introduced. The intensity of illumination at these places is a measure of the sound intensity available for the audience.

a



b

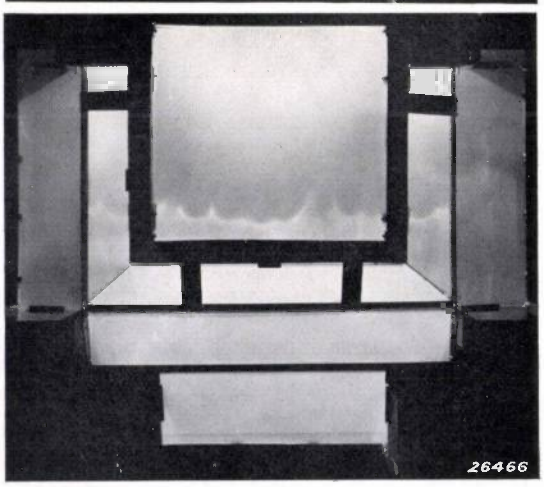


Fig. 8a). Distribution of light on the floor and rows of seats around the floor.
b). Distribution of light when the quasi-parabolic reflector behind the speaker rostrum is inactive. Not only at the rear but also along the sides of the hall is the illumination and therefore the sound intensity, less.

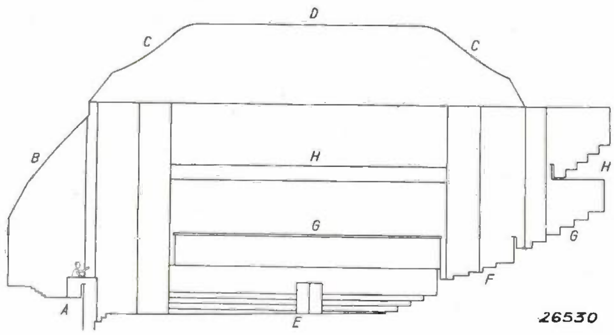


Fig. 9. Cross section of the large assembly hall in the League of Nations Palace. The speakers rostrum is at A and is surrounded by a marble sound reflector B. The ceiling is partially strongly absorbent (C), and partially occupied (B) by a glass lighting element. The audience is situated in the centre portion of the hall E and in the various balconies F, G and H.

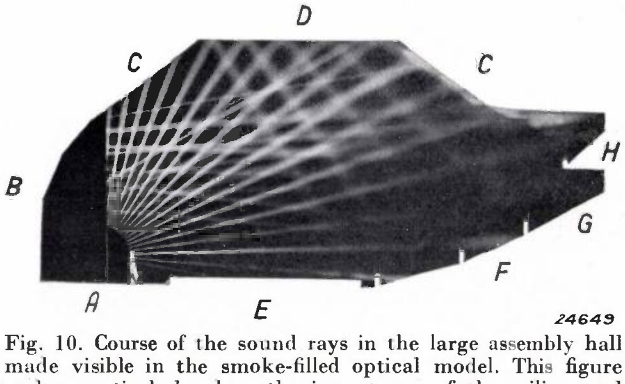


Fig. 10. Course of the sound rays in the large assembly hall made visible in the smoke-filled optical model. This figure makes particularly clear the importance of the ceiling and of the reflector for the distribution of sound in the direction of the horizontal axis of the room. The letters have the same significance as in fig. 9.

an example of which was described previously in this periodical [3]) the source of sound is replaced by a small lamp and the coefficient of optical reflection of the walls of the model are made to correspond somewhat to the acoustic coefficient of reflection in the actual structure. The intensity of illumination of every surface is then a measure of the sound intensity at that place.

In the attempt to detect the cause of undesired deviation the rays may be made visible by filling the model with a mist. *Fig. 7* shows such a model of the large assembly room of the League of Nations Palace in Geneva, the cross section of which is given in *fig. 9*. Figures *8a*, *b* and *10* illustrate several investigations carried out on this model. The inscriptions under the figures furnish an explanation.

Detrimental sound

Echo occurs especially in the case of concave curved surfaces, an example of which was given in fig. 5 at *G*. Such surfaces must therefore be avoided as much as possible, or, when they cannot be avoided experiments must be carried out to find out whether that shape is really permissible.
Such a form is permissible when the radius of curvature is sufficiently great: it may be assumed as a rule in the transverse cross section that a radius equal to twice the height is desirable. In that case the wave is projected on the audience as a fairly parallel beam.

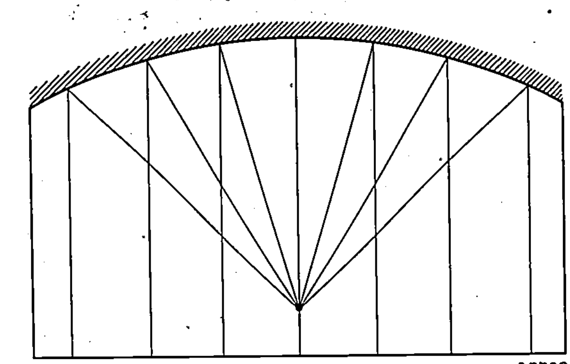


Fig. 11. By means of a ceiling with a radius of curvature of twice its height above the source of sound, the sound is reflected upon the audience in the form of a nearly parallel beam.

Cupolas must always be regarded with suspicion, as well as elliptical forms and the like. Rounded angles with a sufficiently small angle of curvature are often succesfully used. In such cases the focus must be far above the audience which serves to give a strong spreading of the sound. This method can be used to advantage in the angle formed by the ceiling and the rear wall which has a strong tendency to throw the sound back to the speaker as a disturbing echo.

If the reverberation is too strong and absorbing materials must be applied, the best places are the side walls which usually do not play an important part in the distribution of the useful sound.

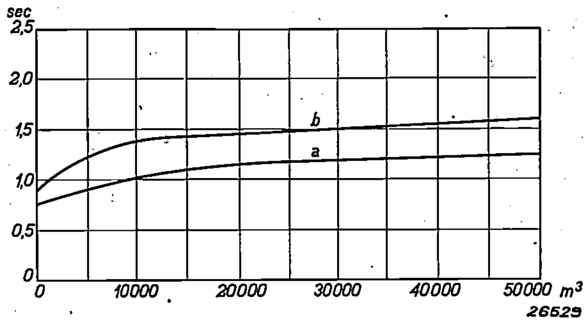


Fig. 12. Reverberation time for the spoken word as a function of the volume of the room, according to Knudsen. Upper curve: recommended values for ordinary theatre auditoria, lower curve: reverberation times at which intelligibility is at a maximum.

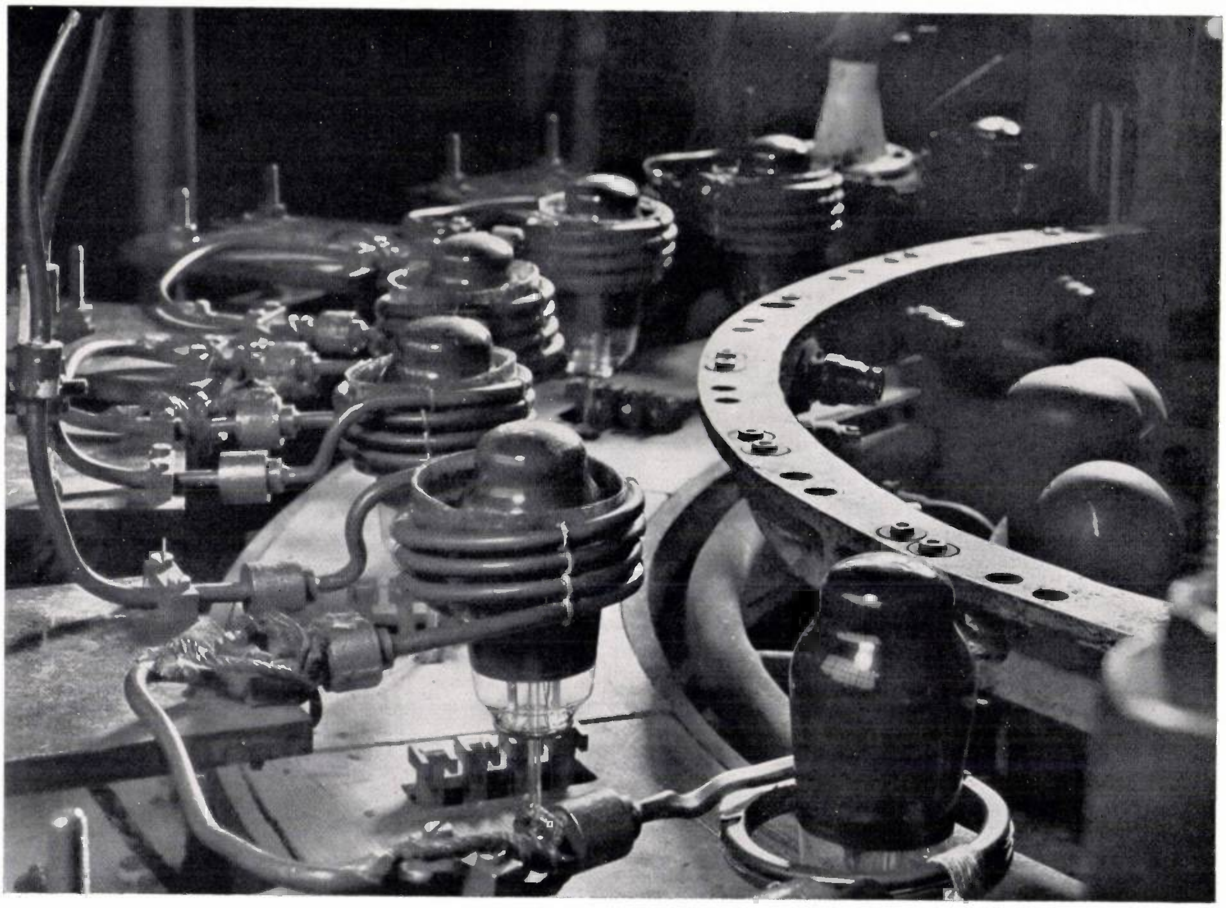
In general it is not wise to reduce the reverberation more than necessary; in the first place because the spoken word becomes dull and less pleasing, and in the second place because there is no sharp difference between walls which reflect useful and those which reflect detrimental sound, so that a decrease in reverberation is in fact always accompanied by a decrease in the intelligible sound. By means of measurements on a large number of halls with unusually good acoustics and on theoretical considerations, various investigators have found optimum values of the reverberation time. These values increase with the volume of the room (see *fig. 12*), which means simply that in large halls a longer reverberation must be allowed in order not to suppress the desired reflections (from the ceiling for example).

In general it may be said that the acoustics of a hall become pleasanter when the wall surfaces which do not contribute to the useful reflection are so constructed that they scatter the sound which is not absorbed. For this purpose a coarse rough surface is necessary since unequalities of the surface which are small with respect to the wave length of sound are ineffective.

In many cases it may be doubted whether modern architecture, with its preference for simple lines and smooth surfaces, is always as efficient from the standpoint of acoustics as it pretends to be.

[3]) R. Vermeulen and J. de Boer, Optical model experiments for studying the acoustics of theatres, Philips techn. Rev. 1, 46, 1936.

PUMPING PLANT FOR RADIO VALVES



The illustration shows the method employed for evacuating radio valves. Unlike the evacuation of incandescent lamps great care must be taken here to outgas the parts mounted in the valve such as cathode, molybdenum grids and nickel anode. The rotating pumping plant has a number of points of connection which move through a series of positions on the circumference of a circle. Each connection point has its own pump which accompanies the valve from one position to the next. The small mercury vapour pumps have a common backing pump which is disconnected during the passage from one position to the other. The connection between the valve and the pump is formed by a thin glass tube. The inner parts are outgassed by means of high frequency fields induced by currents through spiral tubes lying around the valves at certain positions. These spiral tubes are connected in series and are supplied with current from an installation not visible in the drawing. At other positions the outgassing is continued by electron bombardment if necessary. The gas is liberated from the cathodes by electrical heating. In the photograph the electrical contacts for connecting grids, plate and cathode can be seen. The black coating on part of the bulb is for the purpose of eliminating secondary emission.

APPLICATIONS OF CATHODE RAY TUBES II

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In a previous article several examples were given of measurements with the cathode ray oscillograph on various electric mains. We shall here discuss several typical oscillograms of current or voltage in apparatus connected to the mains.

Oscillograms of gas discharge lamps

Our first examples are the voltage oscillograms recorded with two types of gas discharge lamps, namely the sodium lamp SO 650 and the super high pressure mercury lamp HP 300.

A discussion of the functioning and characteristics of these would be out of place in this article. Several peculiarities may, however, be pointed out, since they are particularly striking in the oscillograms.

In *fig. 1* four oscillograms are recorded of the voltage on the lamp when it is supplied from 50 cycle A.C. mains. In the upper half next to each other are those of the sodium lamp, first just after being switched on, and then after it has become warm in normal use. Those of the mercury lamp are given below, first cold and then warm.

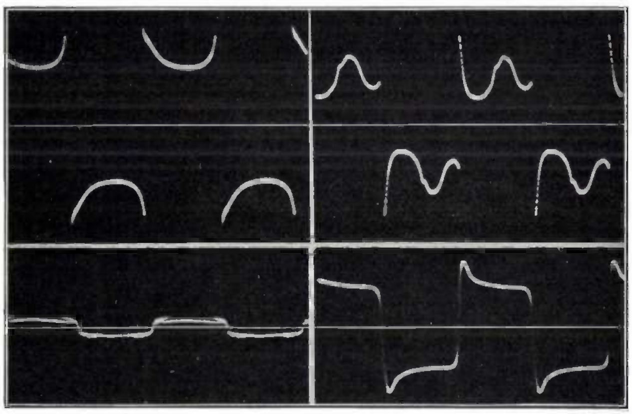


Fig. 1. Voltage of a sodium lamp (above) and of a mercury lamp (below). Left: immediately after being switched on, right: in the working state.

In the first oscillogram the discharge takes place in an atmosphere of neon, no sodium has yet been vaporized. It may be seen that the voltage over a large part of every half period is fairly constant. The fact that, with the exposure time chosen in this case, the photograph of the oscillogram seems to be interrupted between the positive and negative half periods is due to the extremely rapid alternation of the voltage of the lamp after the interruption of the discharge.

After the lamp has become warm the oscillogram takes on a different character. The voltage at which the discharge begins remains about the same, but the working voltage is lower since the discharge now takes place in an atmosphere of sodium. In the middle of the half period, however, there are not enough sodium atoms for ionization and the voltage curve moves toward the level of the neon discharge, since neon atoms must now be ionized. After a half period the current alternates.

A phenomenon which is very well illustrated in this oscillogram is the appearance of "striae". Immediately after the inset of the discharge (in the descending branches) it may be seen that the line is dotted. This indicates that an alternating voltage of much higher frequency is superposed on the curve for the voltage under discussion. The appearance of this fluctuation must be explained in the following way: along the path of the discharge there are points where there is a quite sudden voltage jump. These discontinuities which are called "striae", whose mechanism is not yet entirely clear, usually run from one electrode to the other, and whenever such a point arrives at an electrode the total lamp voltage changes by the amount of the voltage jump in question.

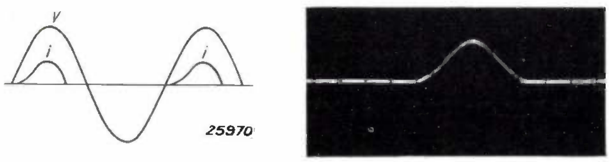
In the oscillogram of the mercury lamp which has just been ignited (lower left) one sees again a discharge under relatively low pressure, since there is as yet only a small amount of mercury vaporized. (The fact that a change was taking place during the time of the exposure may be seen from the appearance of a double line).

When the lamp has become warm and the mercury is vaporized, both the breakdown voltage and the practically constant working voltage are much higher. In the high pressure mercury lamp the amount of metal vapour available is always sufficient to carry the discharge, so that the rare gas also present (ignition gas) no longer plays a part. In contrast to the second oscillogram, the working voltage in this case remains fairly constant.

Oscillographic investigation of a disturbance

A case which was investigated many years ago by means of the cathode ray oscillograph, was that of the occurrence of a rattling noise superposed on the output of an audio amplifier. The first oscillogram of the loudspeaker circuit immediately showed that the disturbance occurred fifty times per second, so that suspicion immediately fell on the supply arrangement of the amplifier. The fact that such disturbances may occur when gas-filled rectifiers are used was known, but with the high vacuum rectifier used here such a disturbance was not immedately to be expected, since the characteristic of a high vacuum rectifier, while not linear, nevertheless exhibits no great discontinuities.

The current through the rectifier is connected to the alternating voltage, as is shown in fig. 2a



a *b*

Fig. 2. Variation of the current through a rectifier valve.

for single phase rectification. The passage of the current lasts less than a half period, due to the fact that the counter voltage of a battery or a charged condenser must first be overcome.

In *fig. 2*b the oscillogram is given of the current through the high-vacuum rectifier valve. To the left may be seen the passage of current during a part of the half period, to the right the horizontal part which indicates that the current cannot flow in the other direction.

While during the growth of the current on the extreme left the transition from the horizontal line takes place quite uniformly, there is a fairly sharp angle at the beginning of the blocking period. That is to say, to the left before this point there is a change in the current, while immediately afterwards $\mathrm{d}i/\mathrm{d}t$ becomes equal to zero. The current now passes through the secondary winding of the transformer with its leakage inductance L_s (see *fig. 3*). As long as there is a definite value of $\mathrm{d}i/\mathrm{d}t$, it causes a counter e.m.f. in the leakage inductance $v = - L_s\, \mathrm{d}i/\mathrm{d}t$, while this voltage disappears as soon as $\mathrm{d}i/\mathrm{d}t = 0$. This voltage component must of course be considered to be superposed on the normal alternating voltage of the secondary winding.

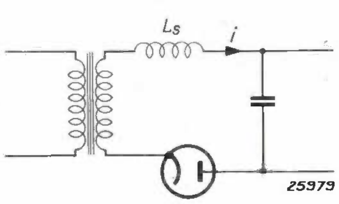


Fig. 3. Circuit diagram of a rectifier valve.

In *fig. 4* is given the oscillogram of the voltage

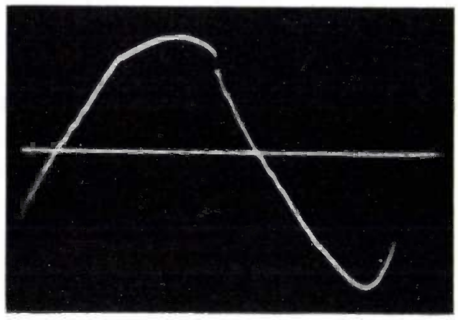


Fig. 4. Variation of the voltage in the secondary of the transformer of fig. 3. The moment at which the current through the rectifier valve is interrupted is distinguished by a voltage jump.

on the secondary winding, and it may be seen that there is actually a sudden jump just after the maximum of the sine curve. This is, therefore, the moment when $L_s\, \mathrm{d}i/\mathrm{d}t$ suddenly becomes equal to zero. The fact that the phenomenon is very rapid, much more rapid than was expected, appears from the fact it is manifested as a break in the oscillogram.

The fact that such a voltage jump at the terminals of the secondary of the transformer can influence the grid of an amplifier valve by capacity coupling is obvious, and the source of the disturbance is thus definitely discovered.

One of the methods of removing such a disturbance is by bridging the secondary of the transformer with a condenser. It is theoretically impossible for such abrupt voltage jumps to occur across a condenser. The effect of a condenser of 0.1 μF is shown in *fig. 5*. The presence of a jump in voltage may still be seen clearly. But from the fact that the curve is continuous, it may be inferred that the

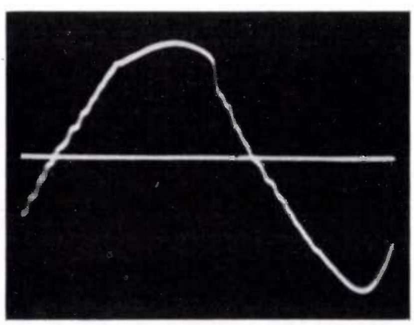


Fig. 5. Variation of the voltage in the secondary of the transformer of fig. 1 when a condenser is connected across the terminals of the secondary. The discontinuity is less sharp.

phenomenon takes place much more slowly. This has been found in practice to be sufficient to prevent a transmission of the effect to the sensitive points in the amplifier.

That without this preventive measure the voltage jump contains very high frequencies is shown by the fact that under certain conditions such interferences also occur in radio receivers, where they are even able to penetrate by way of the aerial and aerial circuits tuned to radio frequencies. This effect is shown in *fig. 6* where the transformer

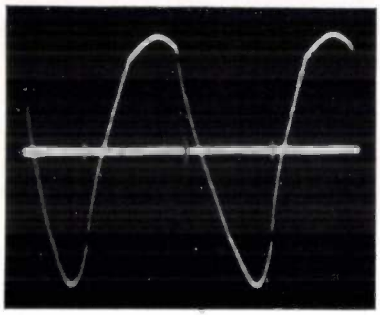


Fig. 6. Intermediate frequency voltage of a radio set (heavy horizontal line; the individual oscillations are not separated from each other), and voltage on the transformer of the supply arrangement. The discontinuities in the transformer voltages (breaks in the curve) lead to vislent, strongly damped, intermediate frequency oscillations.

voltage and the intermediate frequency voltage, measured on the intermediate frequency amplifier of a radio set which was weakly coupled with the rectifier, are reproduced on the same oscillogram. There are now two breaks in the oscillogram of the transformer voltage which may be ascribed to the above-mentioned phenomenon in the two-phase rectifier. The horizontal straight line indicates the time axis, and therefore zero voltage. Just under the breaks in the transformer oscillogram may be seen the rapidly damped intermediate frequency oscillations, one somewhat stronger than the other.

The recording of several curves on one oscillogram

The last case mentioned above was an example of the recording of two curves on one oscillogram. Such records are desirable in other cases also, in order to be able to ascertain the coincidence or the phase shift of characteristic points. There are different methods of producing this result, two of which we shall describe.

The method by which fig. 6 was obtained is the simplest one. The time axis voltage is synchronized with the voltage of the mains with which the apparatus to be investigated is connected. The portable cathode ray oscillograph, type 3952, is especially constructed for this purpose. With this external synchronization a photograph is first made of the transformer voltage and then, on the same plate, one of the intermediate frequency voltage. Since both phenomena are fixed with respect to the mains voltage, the correct mutual relation is obtained. In this way it is possible to superpose any number of oscillograms over each other.

Although this method is very practical for the photography of oscillograms, it is useless for visual observation unless it is sufficient in each case to indicate by points on the screen the characteristic points of the curves. A picture like that of fig. 6 directly on the screen of the cathode ray tube would be preferable.

In order to achieve this, use must be made of one of the circuits which are sometimes termed "electron switches". The principle is as follows. The two voltages are applied in very rapid alternation (very rapid compared to the frequency whose oscillogram is being examined) to the pair of vertical plates of the oscillograph. A double oscillogram is then built up as shown in *fig. 7*. Actually, however, the frequency of commutation is taken so high (15 000 cycles/sec, for example) that an apparently continuous line is obtained.

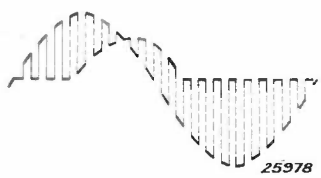


Fig. 7. Movement of the electron beam when two oscillograms are traced simultaneously.

The pinciple of the electron switch is given in *fig. 8* from which details have been omitted. The valves L_1 and L_2 are amplifier valves with a common anode resistance R, which amplify the voltages to be measured, v_1 and v_2, respectively. However, both valves never work at the same time. The screen grids are not connected to a constant voltage, but to a "square-topped" alternating voltage. The screen grid of L_2 is negative when that of L_1 is positive and vice versa. The valve with positive screen grid gives a certain amplification, that with negative screen grid passes no signal at that moment. In this way reproductions of v_1 and v_2 are obtained alternately on the resistance R_1 coupled with the oscillograph, as indicated in fig. 7.

The "square top" alternating voltage for control of the screen grids is obtained from a multivibrator. This consists of two valves L_3 and L_4, which are mutually coupled by the condensers C_3 and C_4. We shall explain briefly the functioning of the circuit.

If one assumes that, due to some cause or other, the grid of L_3 becomes more negative, the current through L_3 and thus the voltage drop on R_3 decreases, and the anode therefore becomes more positive. The latter is, however, coupled to the grid of L_4 which then also becomes more positive. On the other hand the anode of L_4 becomes more negative. This anode is again coupled with the grid of L_3, our starting point. This grid, therefore, becomes more negative, and it is clear that the first assumption introduces an unstable condition in which the first grid becomes more and more negative.

curve is the mains voltage. Figs. 9*d* and *e* give the same data after the lamp has become warm.

In such combined oscillograms the phase relation is naturally the most striking characteristic. In fig. 9*a* the phase shift between mains voltage and current consumed is nearly 90°. While the lamp is warming up this shift may be seen on the oscillograph to decrease gradually, while at the same time the current also decreases slowly (fig. 9*d*). This illustrates an important characteristic of gas discharge lamps. It is well known that account must be taken of the current intensity in choosing the material of the leads, etc., but the power actually consumed is much less than would follow from the product of current and voltage. A low "power factor" must be taken into account.

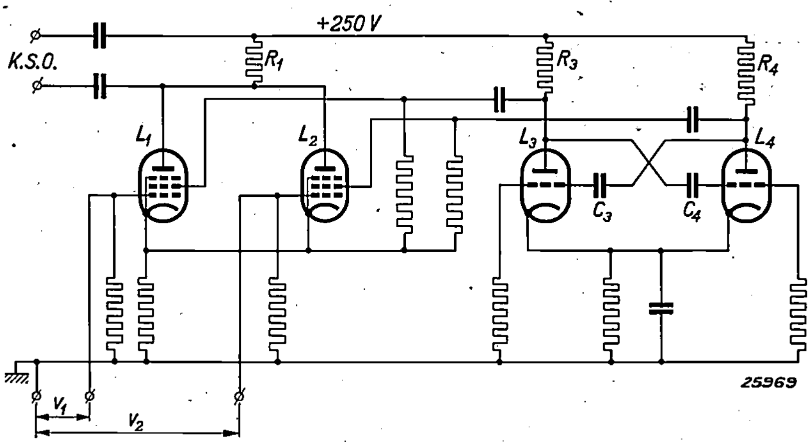


Fig. 8. Diagram of the circuit by means of which the voltages v_1 and v_2 are applied alternately to the deflection plates of the cathode ray oscillograph.

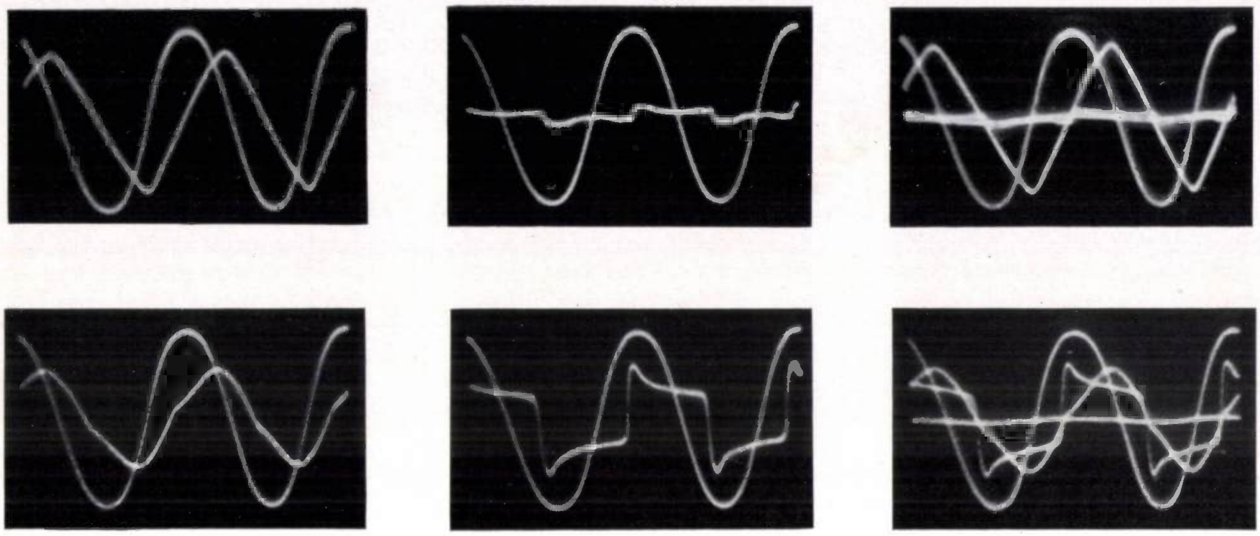
The anode of L_3 rapidly becomes positive, that of L_4 negative. The two amplitudes are, however, limited by the available anode voltage of the supply apparatus, here indicated as 250 volts. As soon as this limit is reached there is a quiescent interval, during which period the anodes keep the voltages they have attained and the grids begin to take on their original voltages again.

As soon, however, as the anode voltage of L_3 begins to fall again, another unstable condition is introduced which is just the reverse of the one described above. In this way the two anode voltages vary in opposite phase regularly back and forth between two limits, and thereby block alternately the two valves L_1 and L_2. The oscillograms of *fig. 9* were recorded with the aid of such a circuit. These curves refer to a high pressure mercury lamp. Figs. 9*a* and *b* give the lamp current and lamp voltage, immediately after switching on the lamp when it was still in the "cold" state. The pure sine

Oscillogram *b* again shows the fairly low working voltage of the still cold lamp (all voltages are reproduced on the same scale) and also the large phase shift. During warming up the lamp voltage increases to that shown in fig. 9*e*. It may be seen that the ignition voltage remains lower than the mains voltage.

While these combined oscillograms are extremely useful for direct observation, it will perhaps sometimes be desired to combine even more curves for the sake of illustration. For example it might be desired to demonstrate that the lamp current and voltage of figures *a* and *b*, and *d* and *e*, respectively, are in phase. This can be done since there is no objection to applying the method first discussed to this case, namely that of making two exposures on one plate. In fig. 9*c* an exposure was first made of the current curve with a zero axis, and then on the same plate and with the same synchronization an exposure like 9*b* with the result that all the curves are combined in one picture. The same method was used to obtain fig. 9 *f*, and in this case particularly it may be seen very clearly that lamp current and voltage are in phase but that both are shifted in phase with respect to the mains voltage.

a *b* *c*



d *e* *f*

Fig. 9. Various combined oscillograms. Upper row: cold mercury lamp; lower row mercury lamp in working condition. The sine curve in each case represents the mains voltage.
a and *d* current combined with mains voltage;
b and *e* lamp voltage combined with mains voltage;
c ans *f* current, mains voltage and lamp voltage combined.

REVIEW OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS GLOEILAMPENFABRIEKEN

1263: J. H. de Boer and H. H. Kraak: Aktivierte Adsorption von Sauerstoff an Molybdänschichten (Rec. Trav. chim. Pays-Bas **56**, 1103 - 1110, Nov. 1937).

When thin metal films come into contact with oxygen recrystallisation occurs and an oxide is formed on the surface, as may be seen from a change in the electrical conductivity. The second effect can be measured separately in the case of molybdenum. Only a very small portion of the surface atoms hold the oxygen by chemical bonds at — 185° C. Van der Waals' adsorption, which causes no change in resistance, also takes place at this temperature, as was demonstrated by the increase of oxidation upon warming the film in a vacuum. As was expected, the activation energy of the oxidation increases in the order: caesium, molybdenum and carbon.

1264: Balth. van der Pol and H. Bremmer: The diffraction of electromagnetic waves from an electrical point source round a finitely conducting sphere, with applications to radiotelegraphy and the theory of the rainbow. (Phil. Mag. **24**, 141 - 176 July 1937, and 24, 825 - 864, Suppl. Nov. 1937).

The diffraction about a sphere of electromagnetic waves from a point source outside the sphere is investigated in different ways. In the first place Watson's method is described, which consists in the transformation of the infinite series of spherical functions for the Hertzian vector into another series which converges much more rapidly for wave lengths which are small with respect to the dimensions of the sphere. A large number of numerical results are given for the case where the point source represents a radio transmitter which may lie either on the surface of the conducting sphere or at a small distance from it, and where the conductivity and the dielectric constant of the sphere may have different values. This method is compared with that in which there is a development in characteristic functions, which latter shows analogy with methods used in the treatment of the vibrating string and in wave mechanics.

In the further treatment the original harmonic series for the Hertzian vector is divided into an infinite sum of other harmonic series. Each of these series may be considered as the description of a train of waves which are reflected one or more times at the inner surface of the sphere. For every high frequency these waves pass over into rays. These series may be approximated by means of a saddle point development. This method gives good approximations for the field of a radio transmitter in the region above its optical horizon. If the point source is at infinity, we obtain a strict interpretation of the well-known theory of Airy for the intensity of the light of the rainbow.

1265: Balth. van der Pol: Propagation of short waves over a spherical finitely conducting earth (Ref. u. Mitt. Int. Kongr. Kurzwellen, Wien, 1937, pp. 34 - 35).

In this lecture the International Congress on Short Waves (Vienna 1937) the same subject was discussed as in **1264.**

1266: K. Posthumus: Kurzwellenröhren (Ref. u. Mitt. Int. Kongr. Kurzwellen, Wien 1937 pp. 78 - 88).

In this lecture a survey was given of the different kinds of valves for the excitation, amplification and modulation of high frequency oscillations. From the method of functioning of the triode generator it is deduced that it can only generate ultra short waves of several metres wavelength with a reasonable efficiency when care is taken that the transition times of the electrons in the valve are sufficiently small. Valves are then discussed whose action depends upon the occurrence of Barkhausen oscillations, and magnetrons with a single anode and with several anodes. With this latter type of magnetron it is possible with air cooling to emit an energy of 60 W at a wave length of 60 cm with a efficiency of 50 per cent. The modulation of these magnetrons is obtained by allowing them to radiate for longer or shorter periods at full power with intervals of non-radiation between, the intermittency being of supersonic frequency.

Finally means are indicated for providing that, upon amplification, the frequency of the amplified oscillation will be determined exclusively by the frequency of the modulating transmitter.

1267: W. G. Burgers, J. D. Fast and F. M. Jacobs: Zug- und Rekristallisationstextur von Zirkondraht (Z. Metallk. **29**, 410 - 412, Dec. 1937).

By hammering and drawing a zirconium rod in an iron jacket without any heating, a zirconium wire is obtained which shows unusually plainly a drawing texture. This texture is distinguished from that of a normal wire by the fact that not only the basic plane (0001) but also a diagonal axis of the second sort (1010) is parallel to the axis of the wire (with a certain spreading). Upon recrystallization a texture appears which deviates from the drawing texture. With the same position of the basic plane in this case a diagonal axis of the first sort (1120) lies parallel to the axis of the wire. Upon heating above the transition temperature and cooling to room temperature this texture remains.

1268: J. L. Snoek: Scattering of X-Rays by conduction electrons in metals (Ned. T. Natuurk. 4, 236 - 244, 1937).

From the spectral composition of the random scattering of X-rays, i.e. from the form of the Compton line, it follows that the velocity distribution of the conduction electrons in a metal is not that given by Maxwell, but that given by Fermi. On a proposal by Debye, Scharwächter investigated in 1937 the variation of the intensity of the Compton small angle scattering of beryllium. The variation found experimentally is found to agree reasonably well with the variation of intensity calculated theoretically for two "free electrons" and not with that for one or three.

1269: J. H. de Boer and J. F. H. Custers: Absorption by Van der Waals' forces and surface structure. (Physica 4, 1017 - 1024, Nov. 1937).

In many cases gases and vapours are adsorbed by means of non-polar Van der Waals' forces. From the adsorption energies observed it follows that a porous structure of the surface is necessary to cause adsorption by means of Van der Waals' forces. Upon the absorption of light by molecules which are adsorbed on layers of salts deposited by sublimation in a vacuum the Van der Waals' adsorption can be distinguished optically from adsorption in which electrostatic forces also take part. In the case of iodine the Van der Waals' adsorption gives a brown coloration and with caesium a blue coloration. Layers of alkali and alkaline earth halides deposited by sublimation in a vacuum are very porous and become very brown in iodine and blue in caesium. Layers of SiO_2, Al_2O_3, ZrO_2 and AgC_1 prepared in the same way show a compact structure and are not coloured; powders of these substances on the other hand absorb strongly and become deeply coloured. Upon sublimation in an atmosphere of argon, SiO_2 is also obtained in the form of a porous layer which adsorbs iodine well.

The layers of SiO_2 deposited in a vacuum, however, are also found to adsorb caesium or iodine. From the absorption of light by caesium which is adsorbed on layers of SiO_2 deposited in a vacuum it follows that the caesium is only adsorbed as individual atoms, as is also the case in the very first stages of the adsorption on CaF_2. On the other hand no Van der Waals' adsorption could be shown. The behaviour of caesium on Al_2O_3 and of iodine on SiO_2 agrees with this. A porous surface is necessary for the attainment of adsorption exclusively by Van der Waals' forces.

1270: A. Th. van Urk: On the cohesive forces of liquids with simple molecules and the so-called law of Stefan. (Physica 4, 1025-1033, Nov. 1937).

Stefan's law is still encountered in handbooks of physics. This law is that "the work necessary to bring a molecule from the depths of a liquid to the surface is equal to one half the heat of vaporization". This, however, is not even valid by approximation, as is proved in this article both experimentally and theoretically. The correct ratio is 1/4 or even less; for hydrogen only it is 0.4.

1271: J. H. de Boer and G. Heller: Die Anisotropy der Van der Waalschen Kräfte. (Physica 4, 1045 - 1057, Nov. 1937).

The orientation of molecules under the influence of the Van der Waals' combining forces has the following causes:

1) the additive character of the Van der Waals' forces which tries to arrange the molecules in such a way that each atom has as many "neighbours" as possible and
2) the anisotropy of the tendency to polarisation, which favours a situation such that the axes of greatest polarizability lie in the line joining the molecules.

These two effects usually oppose each other. In general 1) dominates, so that long molecules lie side by side and ring shaped molecules lie with their planes parallel. With diatomic molecules, however, 2) might possibly be the stronger; and the molecules then take up positions with their axes end to end. Both effects also appear in the attraction of a molecule by a solid wall. Only at sufficiently great distance from the wall can the anisotropy be the more important, but it seems improbable that at such great distance the orientation forces would be important.

1272: H. C. Hamaker: The London-Van der Waals attraction between spherical particles (Physica 4, 1058 - 1072, Nov. 1937).

It is obvious that the adhesive forces between small particles may be ascribed in a large degree to the combining forces of London and Van der Waals. In order to discover something about their order of magnitude the London-van der Waals forces are calculated between two spheres. These forces depend only upon the ratio between the sphere diameters and their mutual separation. A table of numerical data is compiled. For the case in which the mutual separation of the particles is small approximations to the derived formulae are indicated. Finally the change in the results is studied when particles are immersed in a liquid. It is found that in this case also the combining London and Van der Waals forces always produce an attraction.

1273: J. D. Fast: Zirconium (Foote Prints Chem. **10**, 1 - 24, Dec. 1937).

In this article several particulars are given about the preparation of ductile zirconium by thermal dissociation of ZrI_4. Some discussion is devoted to the use of zirconium in transmitter valves, where it is employed for three different purposes. One of these applications is based on the unusually high solubility of nitrogen, oxygen and hydrogen in solid zirconium. The article gives in addition a survey of the work done in the last seven or eight years by various investigators on the subject of metallic zirconium and zirconium oxide with special emphasis on systems consisting of ZrO and other oxides which are fusible with difficulty, and on alloys of zirconium with other metals. Various observations which have not yet been published elsewhere are included in this survey.

1274: F. M. Penning and J. H. A. Moubis: Eine Neutronröhre ohne Pumpvorrichtung (Physica 4, 1190 - 1199, Dec. 1937).

A discharge tube is described in which neutrons are produced by bombardment of a plate of zirconium, loaded with heavy hydrogen, with ions of heavy hydrogen. With the aid of a magnetic field a glow discharge can be maintained with a potential difference of only a few kilovolts at a pressure of about 10^{-3} mm. This acts as a source of ions, so that it is not necessary to have a pump connected between the source of ions and that part of the tube where the ions are accelerated. In a certain case the number of neutrons produced by the tube was equal to that of a mixture of radium and beryllium of about 10 milli-curies. This number can be made considerably greater by water-cooling of the plate which is bombarded by ions.

1275: F. A. Heyn: The radioactivity of nickel, copper and zinc (Physica 4, 1224 - 1228, Dec. 1937).

The radioactivity of nickel, copper and zinc excited by neutrons of varying energies was investigated. Several new nuclear reactions were found.

1276: W. de Groot: The relation between penetrating power and velocity of various elementary particles (Ned. T. Natuurk. 4, 268 - 275, Dec. 1937).

In this lecture before the Netherlands Physical Society, on the basis of experimental data of Mano, Briggs and others the relation was discussed which exists according to a formula of Bethe between the ranges of various heavy charged particles (α-particle, proton and deuton). For different particles the range in relation to the velocity can be deduced from the same graphic representation with the use of different scale values. For the energy region from 10 to 100 kilo-electron volts this curve is fairly valid even for electrons. It is pointed out that in the literature on nuclear reactions the atomic stopping power is often erroneously assumed to be proportional to the number of charges of the nucleus. Finally several measurements on Lenard rays are discussed.

1277: J. H. de Boer: Interpretation of molecular phenomena by means of potential curves II. Light absorption by adsorbed atoms and molecules (Ned. T. Natuurk. 4, 276 - 283, Dec. 1937).

A discussion is given of the absorption of light by molecules and atoms which are adsorbed on a solid wall. Selective photoeffect and thermionic emission of electrons are treated for crystals upon which alkali metal atoms are adsorbed.

1278: J. H. de Boer: Interpretation of molecular phenomena by means of potential curves III. Photoelectrical conductivity, semi-conductors (Ned. T. Natuurk. 4, 284 - 290, Dec. 1937).

The properties are discussed of crystals which contain an excess of electropositive or electronegative atoms. The theory of semi-conductors is treated.

1281: N. F. Moerman: Die Kristallstruktur des Azetaldehydammoniaks $CH_3 \cdot COH \cdot NH3$ (Z. Kristallogr. A **98,** 447 - 455, Febr. 1938).

From investigation of the crystal structure it is found that aldehyde ammonia crystallizes in the rhombohedral system according to the space group. $D^5{}_{3d}$. In an elementary cell there are six $CH_3 \cdot COH \cdot NH_3$ molecules present, but from a determination of the positions of the atoms in space it is found that these molecules in no way form the elementary units of the lattice. A complicated polymerization takes place, in which an important part is played by secondary valence forces, analogous to those which hold together the water molecules in an ice crystal.

1282: H. Bruining and J. H. de Boer: Secondary electron emission, part I. Secondary electron emission of metals. (Physica **5,** 17 - 30, Jan. 1938).

For the contents of this article the reader is referred to the article contributed by H. Bruining in Philips techn. Rev. **3,** 80, Mar. 1937.

1283: J. Zernike: The Beilby layer; structure and occurrence of polished surfaces (Chem. Weekbl. **35,** 28 - 33, Jan. 1938).

A survey is given of investigations on the optically smooth surface which can be obtained by the polishing of metal or glass. It may be concluded that there exists a special so-called Beilby state of the surface layer, which is supposed to be responsible for the disappearance of the phosphorescence of certain materials upon being finely ground and for the photosensitivity of silver iodide in the daguerretype process. It is further assumed that absorption forces are active in the polishing process.

1284: H. C. Hamaker: London-van der Waals forces in colloidal systems (Rec. trav. chim. Pays Bas **57,** 61 - 72, Jan. 1938).

The intensity and the sphere of action of London-van der Waals forces between spherical particles in colloidal systems are discussed in this article, cf. **1272.** The sphere of action is introduced as the distance at which the energy of reciprocal action is equal to the kinetic energy of the Brownian movement, and this is found to be 0.06 to 0.2 times the diameter of the smallest particle; the dimensions of the largest particle is of practically no importance. The adhesion forces observed under various conditions are found to be of the same order of magnitude as was expected theoretically. Our conception of the sphere of action depends upon the dimensions of the particles considered; different definitions of the sphere of action are compared with each other. It is found that a precise comparison of the observations of adhesive forces determined by different methods is only possible on the basis of a theoretical assumption about the nature of the active forces.

1285*: W. Uyterhoeven: Elektrische Gasentladungslampen; 364 pages, 1938. (Julius Springer, Berlin).

In this book the theoretical principles as well as the working characteristics of modern gas discharge lamps are treated as a whole in the simplest possible way for technicians who may have to do with these lamps in practice. After the more theoretical chapters on atomic theory and gas discharge, there follows a chapter on the characteristics of the radiation for illumination technique. In the fourth and fifth chapter the practical forms of construction are discussed for low and high pressure lamps respectively.

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*) An adequate number of reprints for the purpose of distribution is not available of those publications marked with an asterisk. Reprints of other publications may be obtained on application to the Natuurkundig Laboratorium, N.V. Philips' Gloeilampenfabrieken, Eindhoven (Holland), Kastanjelaan.

Published in April 1938:

Philips Transmitting News **5,** No. 1:

P. J. H. A. Nordlohne: The experimental short wave broadcasting station P.C.J.

Practical experience with the ultra-rapid action safety device for rectifier equipment, incorporating mercury-vapour valves.

Tj. Douma: Internal inductance of coils and its influence on the temperature coefficient of the coil.

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ELECTRICAL PHENOMENA IN THE POSITIVE COLUMN AT LOW PRESSURE

by W. UYTERHOEVEN. 537.525

After a brief discussion of the ignition of a self-supporting discharge and of the transition from glow to arc discharge the stationary state of the positive column is treated in detail. In addition a discussion is given of the transport of charge by electrons and ions with the accompanying production of light and ionization to replace the loss of charged particles due to the current to the wall. The current density is not uniform over the cross section, but is a maximum at the axis. The influence of this current distribution on excitation and ionization is examined for sodium and mercury low pressure discharges.

Introduction

An electric current in a conductor consists of the transport of particles, electrons and ions, which carry one or more elementary charges [1]). In the case of metals one or more electrons are freed from the atoms and can move freely in the metal (conduction electrons). The remaining positive ions are bound to a given spot, apart from possible oscillations around their equilibrium positions, and form the lattice. If a potential difference is applied to the ends of a wire, the numerous conduction electrons (the number per cm^3, n_e, is 6×10^{22} for silver), because of their negative charge, move in the direction of the positive pole, the anode. In the stationary state just as many electrons enter the wire per second at the negative pole, the cathode, as leave it at the anode. Since the positive ions of the lattice are bound to their places they do not contribute to the current.

In the case of an electrolyte both positive and negative particles are mobile. Under the influence of an electric field the positive ions move toward the cathode and the negative particles, which in this case are not free electrons but negative ions, move toward the anode. Also in this case the charged particles are already present under ordinary conditions when no electric field is applied; the field serves only to set the charges in motion in a definite direction.

An insulator or non-conductor contains no freely moving charges in the normal state. If such charges are introduced in some way or other the non-conductor becomes a conductor: for example an evacuated space into which electrons are introduced by means of a hot cathode (radio valve). The same method could be used in a gas which is an insulator under normal conditions. There is, however, another method of making a gas conducting. In this method the charges in the gas are produced by ionization under the influence of the electric field applied. This is the phenomenon of breakdown which we shall discuss briefly.

Ignition of a self-supporting discharge

If to the terminals of an incandescent lamp a voltage is applied which is equal to one half of the normal working voltage, a current flows whose strength is of the order of magnitude of two thirds of the normal working current. The charges which produce the current were of course already present in the filament, their velocity of movement, however, is less than with the normal working voltage so that the number of charges which pass through every cross section of the filament per sec, *i.e.* the electric current, is smaller.

If in the same way half the working voltage is applied to a gas discharge lamp, apparently

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1) The elementary charge, which is equal to the charge on an electron is 4.80×10^{-10} e.s.u. $= 1.60 \times 10^{-19}$ coulomb. A single charged positive ion is an atom which has lost one electron, a doubly charged positive ion has lost two. A single charged negative ion is an atom with one extra electron, etc.

nothing happens, *i.e.* with an ordinary ammeter no appreciable current is measured. As a matter of fact, however, a very small current does flow, as may be shown with a galvanometer. As a result of penetrating radiation (radioactive and cosmic) a few ions per cm^3 are always being formed. The radiations referred to are able to free an electron from an atom and thus produce a positive ion at the same time. In the atmosphere such an electron is quickly captured by the oxygen, which can easily bind negative electrons to give a negative ion which sooner or later joins with a positive ion (recombination), and the free charges disappear. In a pure rare gas, however, in which no negative ions occur, the electrons remain free for a much longer time. If such a natural ionization occurs in a rare gas between two plates at different potentials, the electrons will move toward the anode and the positive ions toward the cathode, in other words a current flows through the gas. *Fig. 1* shows the variation of the current between the plates with increasing potential difference. In making these measurements the ionization was produced by irradiation with X-rays, whereby larger currents can be obtained than in ionization by cosmic or radioactive radiation.

The curve of fig. 1 consists of three parts. At very small voltages, V, the current rises and at about 20 volts reaches a saturation value: all the charges formed in the gas arrive at the electrodes, and there is therefore no recombination in the gas or at the wall. If the voltage is further increased the current at first remains constant, but at about 400 volts it begins to increase slowly and then more rapidly. This increase in current is due to the fact that the ions moving under the influence of the electric field themselves cause further ionization of the gas.

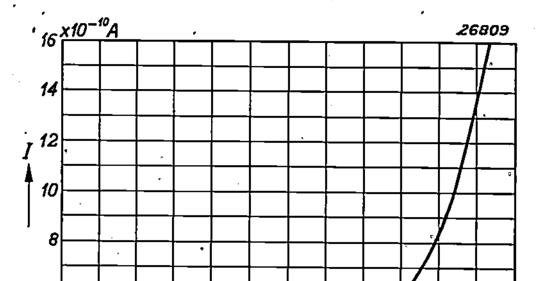

Fig. 1. Variation of the current as a function of the voltage between two large parallel plates (diameter 80 mm, separation 24 mm) with auxiliary ionization in the gas by X-rays. (Argon at a pressure of 27.6 mm).

The energy of the electrons moving toward the anode in the electric field increases. As long as their kinetic energy is less than the excitation energy of the rare gas atoms, they can only give off a very small fraction of their energy of the order of 0.01 per cent in collisions with the latter atoms. Between successive collisions, therefore, the electrons obtain more and more energy, until they are able to excite the atoms or ionize them. An electron which is freed in the neighbourhood of the cathode will then produce a number of electrons on its way to the anode, each of which in turn, after acceleration by the field, can also ionize atoms (*fig. 2*). This

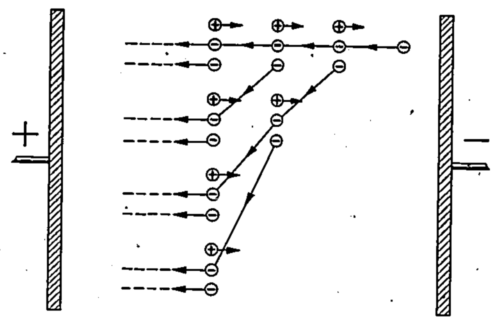

Fig. 2. Diagrammatic representation of an electron avalanche.

phenomenon has been given the appropriate name of electron "avalanche". It occurs more readily the higher the field strength.

If the irradiation with X-rays is interrupted, the current may become very small again: in such a case we are concerned with a "non-selfsupporting discharge" which cannot exist without auxiliary ionization. If, however, the applied voltage is high enough, the interruption of the auxiliary ionization no longer has any effect and we are concerned with a "self-supporting" discharge. The fact that such a discharge may continue is due to the action of the positive ions which free electrons upon striking the cathode, on an average of about one electron per 20 ions. If the field strength and consequently the ionization is so great that for every electron which leaves the cathode 20 ions arrive on it, then one electron can again be freed. This electron in turn provides for 20 ions on the cathode, in other words the process is continuous and we have a "self-supporting discharge". It is clear that this phenomenon will occur at lower field strengths (smaller number of ionizations) the more easily the electrons are freed from the cathode. A tube whose cathode is covered with a barium-barium oxide layer, which gives off electrons readily, will therefore ignite more easily than one with an iron cathode.

Transition from glow to arc discharge

If, as is assumed in the above example, we are concerned with cold cathodes, the series resistance may be so chosen that the self-supporting discharge occurs in the form of a so-called "normal glow discharge". This discharge is distinguished by the fact that it does not cover the whole cathode, and that a voltage jump (100 - 200 volts) occurs at this electrode, the so-called "normal-cathode drop". *Fig. 3* shows schematically a characteristic, *i.e.*

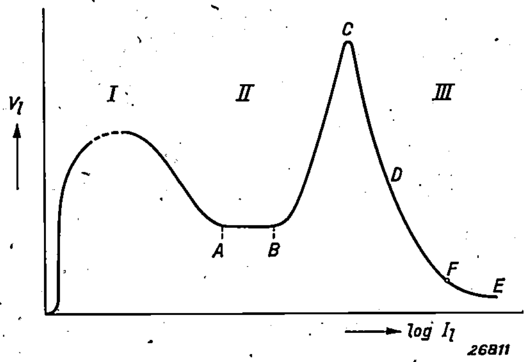

Fig. 3. Diagrammatic representation of the voltage-current characteristic for the discharge between flat plates. The three regions correspond *I* to breakdown, *II* to normal glow discharge and *III* to arc discharge.

the voltage at the terminals of the tube V_l as a function of the current I_l in which the path of the discharge is taken so short that the voltage drop in the cathode region is practically equal to V_l.

In the region of the normal glow discharge *AB* in fig. 3, the current density on the covered part of the cathode is constant for a given cathode material and gas filling. If the current I_l increases, the part of the cathode covered by the glow becomes larger, while the cathode drop V_k remains constant. This continues until the cathode is completely covered (point *B*). If it is desired to increase I_l still further, then the cathode drop increases together with the energy density; one is then concerned with a so-called "abnormal cathode drop".

Nowadays only few gas discharge lamps have a cold cathode (practically only the neon advertising tubes and glow lamps). Most discharge lamps have a hot cathode which is raised to and maintained at the proper temperature by the discharge. Upon ignition a glow discharge first occurs which must of itself pass over into an arc discharge. Such an "arc discharge" is distinguished by a low cathode drop, 10 - 20 volts, which is easily obtained with the help of a hot cathode whose surface is covered with a layer which readily emits electrons (barium-barium oxide).

Finally, on the basis of the diagram in fig. 3, we shall discuss briefly the transition from glow to arc discharge. If the series resistance is so chosen that the discharge is not stabilized in the range of the normal glow discharge *AB*, the current will continue to increase, and when the cathode drop becomes abnormal the voltage V_k will also increase. The consumption of energy in the cathode region, $I_l V_k$, a large part of which is given off to the cathode, may become so great that the latter begins to glow and emit thermions (*c* in fig. 3). The number of ions necessary to free electrons from the cathode then becomes considerably smaller, and the cathode drop falls to 10 - 20 volts (part *DE* of the characteristic in fig. 3). By appropriate choice of the series apparatus the discharge can again be stabilized at any arbitrary point of the characteristic, for instance at *F*.

This transition from glow to arc discharge is, like the original ignition, facilitated by using electrodes covered with a layer having a high electron emission. The temperature at which a given electron emission is reached, is then lower. Moreover, the electrodes are made so small, that they may be raised to the necessary temperature rapidly with a relatively small amount of power and maintained at that temperature.

Stationary state of the positive column. Part played by ionization

In the discharge whose characteristic is given in fig. 3 the length of path of the discharge was assumed to be short, so that only those phenomena taking place in the neighbourhood of the cathode played an important part. When, however, a much longer path of the discharge is taken, as for example, in a neon tube for advertising purposes, the phenomena at the electrodes become relatively less important. The long portion of the discharge joining the region near the cathode with that near the anode, the so-called positive column, then determines the character of the discharge.

In the first place one may ask how the conduction takes place in this part of the discharge. For the sake of simplicity we assume that the discharge takes place in a rare gas at low pressure (<5 cm Hg) and with direct current. A large number of electrons come from the cathode region and move in the direction of the anode under the influence of the electric field. If there were only electrons in the gas they would give rise to a considerable space charge and a steep drop in potential. Several hundred volts per centimeter would then be necessary to cause any appreciable current to pass through the positive column, whilst actually the field strength is usually several volts per centimeter. The negative space charge of the electrons is, however, neutralized by the positive ions present. The number of electrons in every cubic centimeter, the concentration n_e, is practically equal to the number of positive ions n_p [2]); the space charge in the column is practically zero. The values of the concentrations n_e and n_p found in such a low pressure column are approximately $n_e = n_p = 10^{10}$ per cm^3. At a pressure of for instance of 10 mm and a temperature of 300° C, the concentration of the neutral atoms $n = 1.7 \times 10^{17}$ per cm^3, so that one in ten million is ionized.

Although the concentrations of the electrons and of the ions are practically equal, the current may be almost entirely ascribed to the transport of charges by the electrons. This is the result of the much greater mobility of the electrons compared with that of the ions under otherwise similar conditions of field strength and gas pressure. The mobility, *i.e.* the velocity of displacement in the field direction at unit field strength is proportional to the mean free path (the average distance between two successive collisions) and inversely proportional to the mass. In the case of the electron the mean free path is longer and the mass is considerably smaller than that of a positive ion. Thus for example the mobility of the electrons K_e in neon at a pressure of 1 mm and at 0°C at 1 volt/cm equals 1.4×10^6 cm/sec, and for the ions $K_p = 7.5 \times 10^3$ cm/sec. The current density in the tube is proportional to the concentration and the mobility, according to the formula $j_e = n_e K_e \cdot E$ (E = field strength). If the current density of the electrons is compared with that of the ions, since $K_e \gg K_p$, $j_e \gg j_p$ also. In spite of the fact the ions contribute almost nothing to the electric current, they are nevertheless indispensable because they neutralize the negative space charge of the electrons.

Conduction in an ionized gas is therefore somewhat analogous to that in a metal in which electrons also move about among positive ions. There is, however, an important difference, since in a metal the ions of the lattice are bound to their places while in the positive column they can move. In the positive column they move toward the wall and are there lost by recombination with electrons of the negative wall charge, so that neutral atoms are again formed. The ions and electrons which disappear on the wall by recombination must be replaced by others, and this replacement is made by the electrons in the bulk of the gas. The energy necessary for this is taken from the electric field, which at the same time provides for their displacement in the direction of the anode.

[2]) We are supposing that the discharge takes place in a rare gas in which we assume that no negative ions occur.

Energy distribution of the electrons. Light emission.

As is well known, an electron, if it is to produce an ion in a collision with a neutral, non-excited atom, must have a kinetic energy at least equal to the ionization energy of that atom. The energy of an electron $^1/_2\, m_e v_e^2$ is usually given by indicating the potential difference (in volts) which it must pass through in order to obtain the given energy $^1/_2\, m_e v_e^2 = e V_e$. For a 5-volt electron therefore, the energy $e V_e = 4.8 \times 10^{-10}$ (e.s.u.) $\times$ 5/300 (e.s.u.) $= 8 \times 10^{-12}$ erg $= 8 \times 10^{-19}$ watt sec, whilst $v_e = 1.33 \times 10^8$ cm/sec. Not all electrons have sufficient energy for the ionization of a neutral atom; there is a Maxwellian distribution of velocities in the positive column. *Fig. 4* shows this graphically. In the figure the fraction of the electrons which have a velocity between v and $v + dv$ (cross-hatched area) is plotted as a function of v. It may be seen that there are only a few electrons which have either a very high or a very low velocity, most of them have a velocity in the neighbourhood of the commonest value v_w at the maximum of the curve. In fig. 4 are further indicated the average and quadratic velocities, $\bar{v}$, and $\sqrt{\overline{v^2}} = v_f$ respectively. Also are indicated the velocities v_a and v_i of an electron having an energy equal to the excitation ($eV_a = {}^1/_2\, m_e v_a^2$) and ionization ($eV_i = {}^1/_2\, m_e v_i^2$) energies respectively of the gas atoms. An excitation

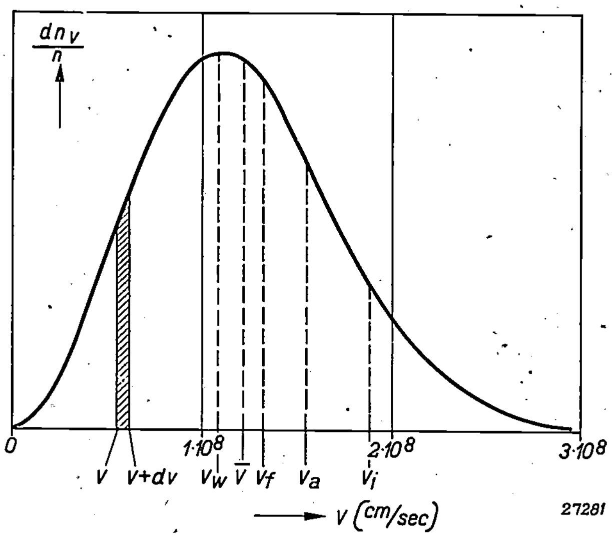

Fig. 4. Maxwellian distribution of the velocities of the electrons. The ordinates give the fraction dn_v/n of the electrons with a velocity between v and $v + dv$. v_w, velocity most commonly occurring, $\bar{v}$, average velocity, and v_f, quadratic average velocity. (The average energy in this example corresponds to a voltage of 5 volts, $^1/_2\, m_e v_f^2 = e\,\overline{V}_e$, where $\overline{V}_e = 5$ volts).

potential $V_a = 7$ volts and an ionization potential $V_i = 10$ volts is hereby assumed, while the average energy of the electrons is set at $\overline{V}_e = 5$ volts ($^1/_2\, m_e \overline{v_e^2} = e\overline{V}_e$). If in the positive column there are electrons which have sufficient energy for ionization, excitation will also occur since the excitation voltage V_a is less than the ionization voltage V_i. The light emission, a result of the excitation of atoms, although it is important in the problem of illumination, is therefore only a by-product of the discharge mechanism. If care is taken that the loss of ions from the path of the discharge is kept small by a large diameter and a suitably chosen gas pressure, a "dark" column, *i.e.* a column through which a high current flows but which gives almost no light can be obtained. Since in this case practically no ionization occurs to replace the ions which have disappeared, the average speed of the electrons remains low, and the excitation of the gas slight. Such discharges have a low field strength, since the field only serves for the displacement of the electrons, in which process only very small energy losses occur due to elastic collisions of the electrons with atoms.

Mechanism of the column discharge. Current to the wall

The following picture of the positive column may serve to summarize the foregoing discussion. A cloud of electrons whose velocities are given by a Maxwellian distribution is displaced in the direction of the anode under the influence of the electric field. The velocities of the individual electrons may have any arbitrary direction, since after covering a free path where the velocity increases in the direction of the field, a collision with an atom takes place so that the velocity continually changes its direction. Due to the action of the electric field, however, there is a resultant velocity in the direction of the anode.

While the cloud is being displaced the electrons gain an amount of energy which, in the stationary state, is just equal to the loss of energy due to elastic collisions, excitation and ionization. Over the whole length of the column the average energy remains constant.

Compared to the electrons the ions may be considered as almost stationary. They do, however, describe curved paths, the form of which is due to the fact that on the way to the wall they are also displaced somewhat in the direction of the cathode. We shall now consider in more detail the lateral movement of the ions which causes the loss of positive charges.

If a foreign gas, for instance chlorine which is easily recognizable by its odour, is introduced at a given place into a body of another gas, for instance the air of a room, it will be found that after a short time the chlorine will have spread throughout the whole mass. This is the result of diffusion which tends to abolish differences of concentration. Diffusion proceeds more rapidly the greater the free path and the velocity of the particles introduced. If an equal number of electrons and ions are introduced locally into a rare gas atmosphere, the electrons will diffuse away more quickly. The ions remaining behind will, however, due to the electrical attraction, have a retarding action on the diffusion of the electrons, and conversely the rapidly moving electrons will accelerate the diffusion of the ions.

A similar phenomenon takes place in the positive column: the electrons diffuse more rapidly away from the middle of the path of the discharge and carry the slower ions with them. They then encounter the wall as an obstacle and charge this negatively, so that the wall attracts ions and repels electrons. This negative charging of the wall proceeds until, on an average, the ions are so much accelerated and the electrons so much retarded that per unit of time just as many positive as negative charges arrive at the wall. A stationary state is then reached in which the charges arriving in equal numbers disappear by recombination on the insulating wall. The phenomenon here described bears the name of ambipolar diffusion current.

The variation of the potential along a radius perpendicular to the axis of the tube is given in *fig. 5a*. Starting from the axis the potential decreases

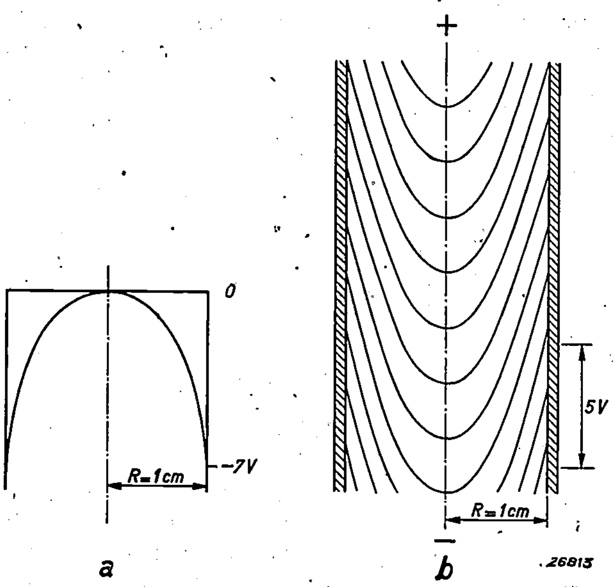

Fig. 5. *a* Potential variation along the radius, *b* Curvature of the equipotential planes.

at first slowly and then more rapidly in the neighbourhood of the wall [3]). The ions are therefore accelerated and the electrons retarded in their movement toward the wall. Only the speediest of the electrons are able to reach the wall, the slower electrons are reflected and return to the axis of the discharge path.

Fig. 5b gives diagrammatically the curvature of the equipotential planes.

For the sake of simplicity it is here assumed that no striae appear in the column as is often the case in the rare gases. The stationary striae are distinguished by the alternately light and dark parts of the discharge. The field strength, the energy and the concentration of the electrons then show periodic fluctuations. Moving striae may also appear which may be observed with the aid of a rotating mirror.

Distribution of current in the discharge cross section

Since the potential decreases from the axis to the wall, the slower electrons will be unable to run counter to the retarding field and the concentration n_e as well as n_p will decrease in the direction of the wall. The variation of the concentration as a function of the distance to the axis can be derived in the following way.

We consider in a piece of the discharge tube 1 cm in length a small volume bounded by two concentric cylinders with radii r and $r + \mathrm{d}r$ which differ only slightly and by two planes perpendicular to the axis (*fig. 6*). In the stationary state just as many charges must enter this volume as leave it, since it is assumed on reasonable grounds that no loss of charge takes place in the gas by recombination. Moreover, the current is the same at all cross sections of the tube so that just as many charges enter at one plane along the axis as leave it by the other. This holds for the electrons as well as for the ions. It is therefore enough to consider the transport of charge through the cylindrical shell.

In the direction perpendicular to the axis this transport of charge takes place by the above-described ambipolar diffusion current, whereby just as many electrons as ions pass through a cylindrical shell per second, since the resultant current must be zero because of the insulating wall.

[3]) $n_e = n_p$ holds strictly only at points on the axis; the field is there in the direction of the axis and has no radial components. At other points n_p is somewhat larger than n_e, the difference is, however, so slight that it may usually be neglected. Per centimeter length of tube the sum of negative wall charge and positive space charge is exactly zero.

The number of particles which passes through 1 cm² of the surface of the cylinder is proportional to the drop in concentration along the radius

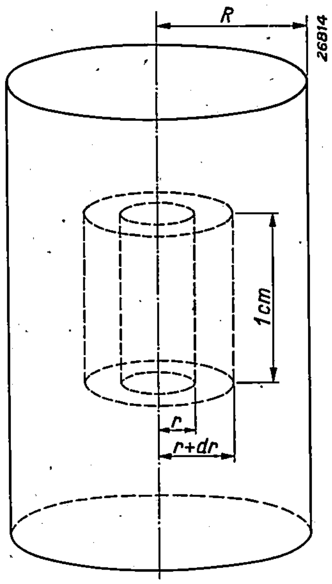

Fig. 6. Volume element between concentric cylinders with radii r and $r + \mathrm{d}r$.

$\mathrm{d}n_e/\mathrm{d}r$, and to the so-called diffusion constant D. The latter depends upon the velocity $\overline{v}$ and the mean free path x; the greater these values, the more rapid the diffusion. For the surface of the inner cylinder therefore we find the number of electrons entering:

$$-D\,2\,\pi\,r\cdot 1\cdot\left(\frac{\mathrm{d}n_e}{\mathrm{d}r}\right)_r \ ^{4}).$$

In the same way for the number of electrons passing out through the outer cylinder we find:

$$-D\,2\,\pi\,(r+\mathrm{d}r)\cdot\left(\frac{\mathrm{d}n_e}{\mathrm{d}r}\right)_{r+\mathrm{d}r}.$$

The difference between these two expressions gives the number which finally passes out of the volume considered by diffusion in excess of those which entered. This number must be equal to the number formed anew by ionization. The number of ionizations is proportional to the electron concentration n_e and to the number of collisions between electrons and atoms; in addition it increases with the average energy of the electrons. If we let z equal the average number of new electrons formed per electron per second, then the number of ionizations in the cylindrical shell per second is $2\,\pi\,r\,\mathrm{d}r\cdot z\cdot n_e$. This must now be equal to the above determined loss

[4]) In spite of its negative sign, this expression is actually positive, since $\frac{\mathrm{d}n_e}{\mathrm{d}r} < 0$ (n_e decreases with increasing r).

by diffusion. From this the differential equation below can be derived, which contains as dependent variable the concentration n_e, as independent

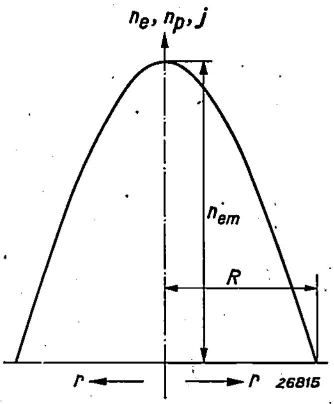

Fig. 7. Variation of the electron concentration n_e, the ion concentration n_p and the current density j as functions of the distance from the axis.

variable the distance r to the axis and in addition only known quantities:

$$\frac{d^2 n_e}{dr^2} + \frac{1}{r}\frac{d\,n_e}{dr} + \frac{z}{D}\,n_e = 0\,.$$

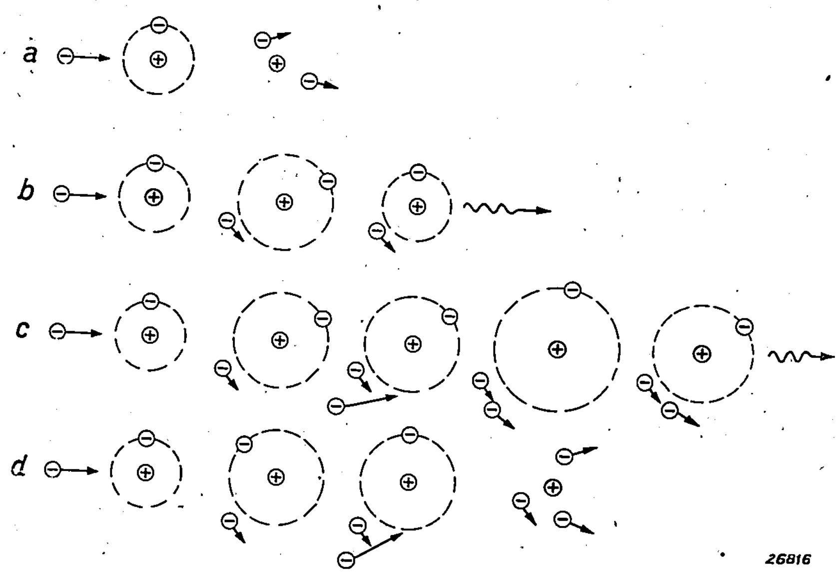

Fig. 8. Schematic representation of excitation and ionization by electron impacts: *a*) ionization upon a single impact, *b*) excitation upon a single impact with consequent emission of light, *c*) cumulative excitation by two successive impacts with light emission, *d*) cumulative ionization by two successive impacts.

If the new variable, $x = r\sqrt{z/D}$, is introduced into this equation, a differential equation is obtained a solution of which, finite for $x = 0$ (at the axis), is given by the so-called Bessel function of the zero order:

$$n_e = n_{em}\,J_0\,(x).$$

n_{em} is the concentration of electrons at the axis. The function $J_0\,(x)$ is zero for $x = 2.405$, and since it is assumed that $n_e = 0$ at the wall, for $r = R$, we may write $2.405 = R\sqrt{z/D}$, from which it follows that $\sqrt{z/D} = 2.405/R$. For the concentration n_e as a function of r we find:

$$n_e = n_{em}\,J_0\,(r\,\sqrt{z/D}) = n_{em}\,J_0\left(2.405\,.\,\frac{r}{R}\right).$$

The variation of n_e as a function of r is given in *fig.* 7. Since the field strength in the direction of the axis is the same over the whole diameter, and the current density is proportional to the concentration n_e, the curve of fig. 7 also represents the variation of the current density as a function of the radius.

Influence of the current distribution on ionization and excitation

Although the result of the above derivation is very well confirmed by measurements, it can nevertheless only be correct as a first approximation, since all the so-called cumulative effects are neglected. An example of such an effect is ionization in steps. When an atom is raised to an excited state by collision with an electron it will generally give off the energy taken up in the form of radiation. However, for a very short time it remains in the excited state, usually for about 10^{-8} sec. If the current density j of the discharge is high, and therefore also the electron concentration n_e, the excited atom will have an appreciable chance of undergoing collision with another electron. The result of this second collision may be that the atom is raised to a higher excited state or is ionized (*fig. 8*). The above assumption that the number of ionizations is proportional to the concentration n_e is then no

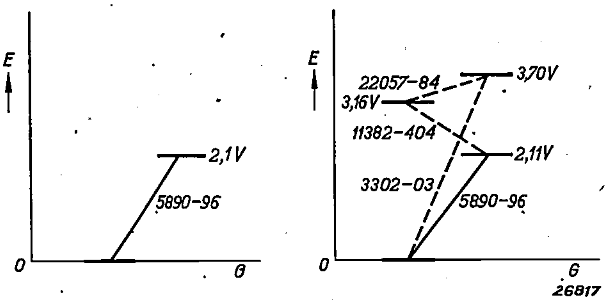

Fig. 9. Simplified diagram of the sodium spectrum. If only the 2.1 volt level is excited only the yellow lines with λ = 5890 and 5896 Å are emitted. When the 3.7 volt level is excited lines outside the visible spectrum are emitted also G = ground, or zero state.

longer entirely correct. It is clear that a current distribution like that given in fig. 7, where the current density is not the same over the whole cross section but has a maximum at the axis, favours the appearance of cumulative effects.

The current distribution found is decidedly favourable to ionization. If ionization takes place in two steps, the energy of the electrons need not be at least V_i volts, but an energy of V_a is sufficient. The diffusion of the electrons to the wall, as well as that of the ions, becomes less so that the loss of ions and the number of ionizations necessary becomes smaller. Apart from this it is already an advantage that the ions are formed mainly in the neighbourhood of the axis. The average distance which the ions must cover before they reach the wall then becomes greater, as well as the time during which they remain in the discharge. We have seen that the part played by the ions consisted in the neutralization of the space

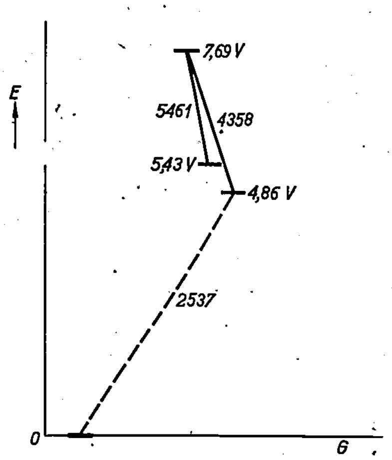

Fig. 10. Simplified scheme of the mercury spectrum. Excitation of the blue line with λ = 4358 Å and of the green line with λ = 5461 Å.

charge of the electrons ($n_e = n_p$); if every ion has a longer average life, fewer new ones need be produced per second to keep n_p at the proper value, in other words, the energy required for ionization may be smaller.

Finally we must not forget that in the case of gas discharge lamps the aim is to attain a maximum emission of light with a minimum consumption of energy. The energy used for ionization must be considered an unavoidable loss.

Whether the current distribution of fig. 7 is favourable or unfavourable to the emission of light depends upon the nature of the states which must be excited. If it is a question of the resonance line (the yellow sodium line for instance) which

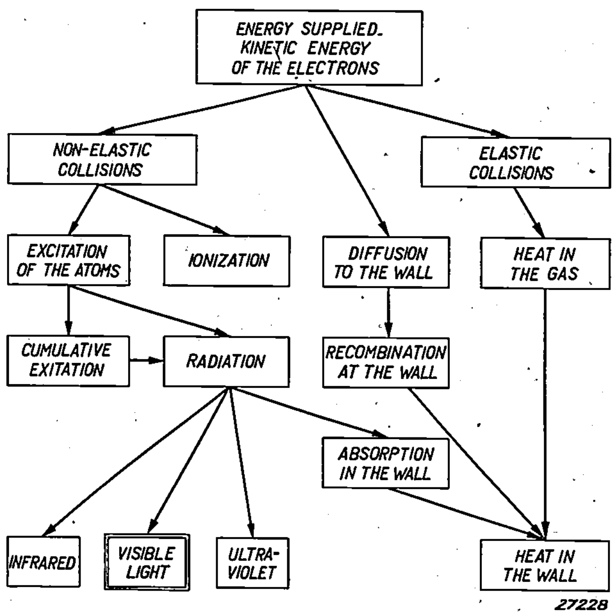

Fig. 11. Survey of the elementary processes in the positive column.

is emitted when the atom returns to the normal state from the excited state closest to it, concentration of current around the axis is in principle undesirable. Such concentration promotes excitation to higher levels, whereby lines other than the desired resonance line appear (*fig. 9*). In the case of sodium for example, lines in the invisible parts of the spectrum are obtained: infrared and ultraviolet, or lines which are less effective than the yellow line taking into consideration the sensitivity curve of the eye, with as result a decrease of efficiency for illumination.

If, on the other hand, it is a question of socalled higher lines which occur upon transitions between two excited levels, the promotion of cumulative effects is desirable. *Fig. 10* makes this clear for the green (5461 Å) and blue (4351 Å) mercury lines. For the production of the blue line, for example, it is more advantageous to bring the atom from the 4.86 volt state, to which it has fallen after emission of the blue line, immediately back again to the 7.69 volt state, than to wait until the atom has returned to the zero state and emitted the ultraviolet line 2537 Å. In the first case only 7.69 - 4.86 = 2.83 volts are needed and this energy is radiated entirely in the visible region; in the second case, however, 7.69 volts are needed of which only 2.83 volts are transformed into visible light.

These cumulative effects are very much favoured in the case of mercury by the fact that 4.86 volts is a "resonance level" and 5.43 volts a so-called "metastable level". Since line 2537 is a resonance line, it is readily absorbed by the atoms in the zero state. After about 10^{-7} sec emission again takes place, and the energy radiated can then be absorbed by another atom in the zero state. This process is repeated very many times before the resonance radiation leaves the tube, so that the concentration of atoms in the 4.86 volt state increases sharply and with it the chance of cumulative excitation from this level to a higher one. For the metastable 5.43 volt level also the chance of cumulative excitation is large. If the atom has been brought into the 5.43 volt level by collision with an electron, it cannot easily return directly to the zero state by emission of radiation, it usually does so by means of collision with other atoms or electrons. The life of such a metastable atom, about 10^{-4} sec., is much longer than that of an atom in an ordinary excited state, so that the chance of cumulative excitation is thereby increased. In the low pressure mercury discharge used in the blue illuminated advertising signs, concentration of the current therefore is an advantage. In the high pressure mercury lamps this phenomenon is even more stimulated by the compression of the discharge to a narrow zone near the axis.

Fig. 11 gives a schematic survey of the various processes here discussed which occur in the positive column.

THERMOJUNCTIONS

by J. W. L. KÖHLER. 621.317.7.082.62

The construction and action of thermojunctions are discussed in this article. A detailed account is given of the factors which affect the sensitivity and the characteristic of such junctions. The choice of the meter movement and the classification of thermojunctions are explained. Finally a survey is given of Philips thermojunctions.

Introduction

When an alternating current of very low frequency is sent through a suspended coil galvanometer which has its zero point in the middle of the scale, the meter indicates the current at every moment; the pointer moves back and forth with the frequency of the alternating current. If the frequency is raised, then after a moment of resonance the deflection becomes smaller and the pointer finally remains practically still due to the fact that the moving system of the meter cannot follow the high frequency. A suspended coil galvanometer is therefore not suitable for measuring alternating currents. Such currents can only be measured with an instrument whose deflections are always in one direction, no matter what the direction of the current, and whose pointer can therefore adjust itself to a definite average deflection. This condition is fulfilled by dynamometers, soft-iron meters, rectifying meters and thermal instruments. In this article we shall discuss the last type of instruments.

In thermal instruments use is made of the heat development in a resistance, when a current passes through it. It is obvious therefore that the direction of the current can have no influence on the deflection of the meter. The heat developed can be used in various ways:

a) The thermal expansion of the resistance filament can be indicated by a pointer in some way or other.
b) The heat developed can be used to heat a quantity of gas whose increase in volume is indicated by a drop of liquid in a capillary.
c) The change in resistance of a filament forming part of a Wheatstone bridge may be measured.
d) The heating of the filament can be measured with a thermocouple connected to a direct current meter.

We shall treat the last method in detail.

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We shall treat the last method in detail.

Construction and action of thermo-electric ammeters

Klemenčič was the first to construct a measuring instrument on this principle. The instrument consists of two wires of different materials which form a thermo-electric couple. The two wires are wound together, knotted together or welded together at the middle (see *fig. 1*). The current to be measured

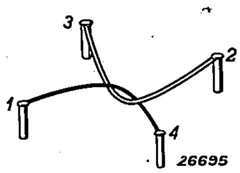

Fig. 1. Klemenčič's thermocouple. The wire of one metal is fastened to terminals *1* and *4*, that of the other metal to terminals *2* and *3*. The wires are joined in the middle electrically in some way. The current to be measured passes through along the terminals *1* and *2*, the galvanometer is connected between *3* and *4*.

is supplied to the terminals *1* and *2*; the wires become warm at the middle due to the heat development; this causes a thermal electromotive force between the terminals *3* and *4*. This voltage can be measured by connecting a sensitive meter to the terminals.

There are various objections to this construction which will appear later. *Fig. 2* gives a sketch of

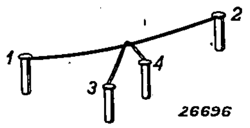

Fig. 2. Modern thermocouple. Between terminals *1* and *2* a filament is stretched, and between *3* and *4* there is a thermocouple whose junction is fastened to the filament.

a modern thermo-electric ammeter from which it may be seen that the principle is the same as that of the original. A filament is stretched between *1* and *2*. Midway between the terminals a thermocouple is fastened to the filament with its junction on or close to the filament. The ends of the thermocouple are fastened to the terminals *3* and *4*, between which the direct current ammeter is connected. The whole is often placed in an evacuated bulb. When this is done there is no loss of heat by convection, and the sensitivity of the instrument is increased. In certain types the thermocouple is electrically connected with the filament, in others it is insulated from the filament. This construction is not so simple as the original one, but it has certain advantages. One of these is immediately obvious. The thermo-electric properties of the filament material need not be considered, so that the choice becomes much greater.

We shall now discuss the action of the thermo-electric ammeter. If a current i is sent through the filament with the resistance R_g, an amount of heat equal to i^2R_gt will be developed in the filament. The temperature of the filament will thereby be raised; the heat is lost by conduction over the terminals, and by conversion into electrical energy through the thermocouple and by radiation. If the amount of heat which is given up is proportional to the increase in temperature, the latter will be proportional to the heat developed, *i.e.* to i^2. The same is true of the temperature of the junction of the thermocouple. If the thermal electromotive force is proportional to the temperature difference, it will also be proportional to i^2, and the same is true of the current in the circuit comprising thermocouple and galvanometer, which current is proportional to the thermal electromotive force. The deflection of the galvanometer is therefore proportional to the square of the current in the filament.

The ratio between the current through the galvanometer and the current through the filament is called the characteristic of the thermocouple. Here the characteristic is quadratic. Therefore, with a meter having a quadratic scale division, after calibration of one point on the scale, the primary current may be read off directly. A meter with a linear scale can also be used, the reading is then proportional to i^2, *i.e.* to the energy which can be developed by the primary current in a resistance; this method of measurement has many practical applications.

Mathematical analysis

We shall give a brief account of the mathematical analysis of the action of the thermo-ammeter. We shall first consider the filament alone without the thermocouple. The following differential equation holds for the temperature increase in a straight wire stretched between two points and heated by a current:

$$\frac{\mathrm{d}^2\tau}{\mathrm{d}x^2} - \frac{\sigma u}{\lambda q}\tau + \frac{Ai^2\varrho}{\lambda q^2} = 0 \quad . \quad . \quad . \quad . \quad (1)$$

Where:

x the coordinate of length of the wire from the middle in centimetres,

τ the temperature difference with respect to the surroundings at the point x in degrees centigrade,

u the circumference of the wire in centimetres,

q the area of the cross section of the wire in square centimetres,

i the electric current in amperes,

A the electrical equivalent of heat, 0.2388 calories per watt second,

ϱ the specific resistance of the wire in ohm centimetres,

σ the heat radiation in calories per centimetre second,

λ the specific conductivity for heat, in cal/cm sec. degr. C.

Equation (1) can be derived as follows. Consider a length dx of the wire; at the ends of this element the temperature gradients are $\left(\frac{d\tau}{dx}\right)_1$ and $\left(\frac{d\tau}{dx}\right)_2$. By heat conduction therefore the following amount of energy is supplied to the element:

$$\lambda q\left[\left(\frac{d\tau}{dx}\right)_1-\left(\frac{d\tau}{dx}\right)_2\right]=\lambda q\frac{d^2\tau}{dx^2}dx\,.$$

An amount of heat $\frac{Ai^2\varrho}{q}\,dx$ is developed in the element by the current i. If the element is situated in a vacuum it loses energy to the surroundings only by radiation; the amount of this radiation is $u\,\sigma\,\tau\,dx$, where it is assumed that the radiation is proportional to the temperature difference, which is permissible when the increases in temperature are small (see below for more detailed explanation). In the stationary state the energy supplied must equal the energy given off, and therefore

$$\lambda q\frac{d^2\tau}{dx^2}dx+\frac{Ai^2\varrho}{q}dx=u\sigma\tau\,dx,$$

from which (1) follows.
In this derivation it is therefore assumed that ϱ, λ and σ are independent of the temperature; we shall see later what changes are necessary when this condition is no longer satisfied.

If the limiting condition is introduced that the increase of temperature at the terminals is zero ($x=\pm l/2$), then the equation has the following solution:

$$\tau=\frac{Ai^2\varrho}{\sigma u q}\left\{1-\frac{\cosh x\sqrt{\frac{\sigma u}{\lambda q}}}{\cosh\frac{l}{2}\sqrt{\frac{\sigma u}{\lambda q}}}\right\}\quad\cdots\quad(2)$$

($\cosh\alpha$ is the hyperbolic cosine of α, *i.e.* the abreviated notation of $^1/_2(e^\alpha+e^{-\alpha})$; the expression hyperbolic tangent which occurs later stands for $\left.\frac{e^\alpha-e^{-\alpha}}{e^\alpha+e^{-\alpha}}\right)$. In fig. 3a this increase of temperature is drawn for different points of the filament.

At the middle the increase in temperature (t) is:

$$t=\frac{Ai^2\varrho}{\sigma u q}\left\{1-\frac{1}{\cosh\frac{l}{2}\sqrt{\frac{\sigma u}{\lambda q}}}\right\}=\frac{Ai^2\varrho}{\sigma u q}\left\{1-\frac{1}{\cosh K}\right\}(3)$$

where $K=\frac{l}{2}\sqrt{\frac{\sigma u}{\lambda q}}=l\sqrt{\frac{\sigma}{\lambda d}}$ for round wires with the diameter d. The increase of temperature is therefore actually proportional to i^2. When K is small (this is usually the case in practical applications as we shall see later), we may write:

$$\cosh K=1+K^2/2,$$

and (3) becomes:

$$t=\frac{A\,i^2\varrho\,l^2}{8\,\lambda\,q^2}=\frac{A\,i^2}{8}\,\frac{\varrho\,l}{q}\,\frac{l}{\lambda q}=\frac{Ai^2}{8}\,R_g\,R_w\quad(3a)$$

In this expression R_g is the electrical and R_w the heat resistance of the filament. In (3a) the radiation constant δ no longer occurs due to the approximation. This fact will be found later to be very important.

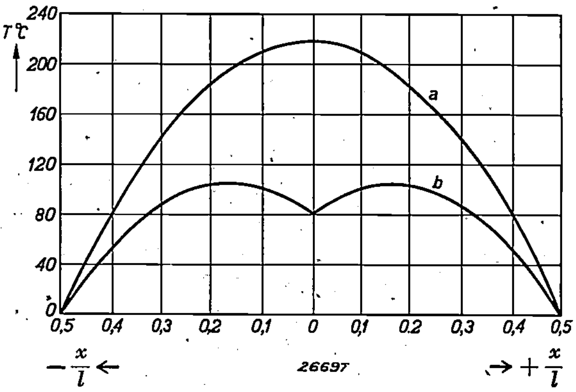

Fig. 3) *a*) Temperature of a stretched wire through which current is flowing.
b) Temperature of the same wire when a thermocouple is fastened at its middle point.

We shall now pass on to the case where a thermocouple is fastened to the middle of the filament. This can be described in good approximation by calculating that a quantity of heat M is removed at the middle of the filament per second and per degree increase of temperature. The differential equation then becomes more complicated; we shall only give the result for the increase in temperature t_M at the middle of the wire:

$$t_M=\frac{Ai^2\varrho}{\sigma u q}\,\frac{1-\frac{1}{\cosh K}}{1+\frac{M\,\mathrm{tgh}\,K}{2\sqrt{\sigma\lambda u q}}}=\frac{t}{1+\frac{M\,\mathrm{tgh}\,K}{2\sqrt{\sigma\lambda u q}}}\quad\cdot\cdot\,(4).$$

The increase of temperature is therefore greater, the smaller M. The factor $\frac{\tanh K}{\sqrt{\sigma\lambda u q}}$ takes account of the fact that the heat lost flows to the middle of the filament from parts lying near the middle. For small values of K the following holds:

$$t_M = \frac{\dfrac{A\,i^2\,\varrho\,l^2}{8\,\lambda\,q^2}}{1+\dfrac{l}{4\lambda q}M} \quad . \quad . \quad . \quad . \quad (4a)$$

In this case also the increase in temperature is proportional to i^2. The influence of the heat conduction is less than in (3a). This is clear enough. In the first case a great increase in temperature could be attained by a high heat resistance of the filament; the supply of heat to the middle of the filament has now become an important factor. In *fig. 3b* we give an idea of the variation of temperature along the filament; the middle no longer has the highest temperature.

Finally the value of M must be expressed in terms of the dimensions and constants of the thermocouple. An analogous differential equation is valid here, but with different limiting conditions. The solution is the following:

$$M_1 = \frac{\sqrt{\sigma_1\,\lambda_1\,u_1\,q_1}}{\mathrm{tgh}\,K_1}, \quad . \quad . \quad . \quad . \quad . \quad (5)$$

where the subscript 1 indicates that the values are for one of the wires of the thermocouple. The same formula with the subscript 2 holds for the other wire; the length of the wires is $\frac{l_1}{2}$ and $\frac{l_2}{2}$; $K_1 = \frac{l_1}{2}\sqrt{\frac{\sigma_1 u_1}{\lambda_1\,q_1}}$. The loss of heat through the whole thermocouple is:

$$M = M_1 + M_2.$$

At small values of K_1 (5) becomes:

$$M_1 = \frac{2\,\lambda_1 q_1}{l_1} \quad . \quad . \quad . \quad . \quad . \quad . \quad . \quad . \quad (5a)$$

If we introduce this value of M into (4) the equation becomes:

$$t_M = \frac{\dfrac{A\,i^2\,\varrho\,l^2}{8\,\lambda\,q^2}}{1+\dfrac{l}{2\,\lambda q}\left\{\dfrac{\lambda_1\,q_1}{l_1}+\dfrac{\lambda_2\,q_2}{l_2}\right\}} \quad . \quad . \quad (4b)$$

The temperature of the middle of the filament is therefore now expressed in the dimensions and constants of the wires of the thermo-couple.

Let us first consider the numerator of (4b). Leaving out the constants of the material we may write:

$$t_M \sim \frac{i^2\,l^2}{q^2} \sim i^2\,R_g^2.$$

The dimensions of the filament occur only in the form l/q; filaments, therefore, with different dimensions but with the same resistance undergo the same increase in temperature with a given current. The voltage drop caused by the filament, $i\,R_g = V$, is not dependent on the dimensions of the filament. If a definite rise in temperature must be attained with different currents, the resistance of the filament must be inversely proportional to the current. The energy necessary for a given temperature increase, V^2/R_g, is inversely proportional to the resistance, and therefore proportional to the current; thermo-ammeters for small currents, therefore, possess a greater energy sensitivity than those for heavy currents.

The foregoing remains true when the second term in the denominator is small with respect to one; this is the case when there is only a slight loss of heat through the thermo-couple, or when $l/2\,\lambda\,q$ is small. The latter is always the case with thermo-ammeters for larger currents ($i\,l/q$ must always be constant). If the second term may not be neglected, which is the case with thermo-ammeters for low currents, the sensitivity to energy is smaller. A small heat loss through the thermocouple is therefore favourable in such a case. Since λ_1 and λ_2 are usually given (the choice of material is determined by the thermo-electric properties), this can be achieved by making q_1/l_1 and q_2/l_2 small. The resistance of the thermo-couple $R_e = \frac{\varrho_1 l_1}{2\,q_1}+\frac{\varrho_2 l_2}{2\,q_2}$, however, becomes large as a consequence of this measure; a limit is thereby set to the reduction of M (see later). The dimensions of the thermocouple occur only in the form q/l, so that wires with the same electrical resistance cause the same heat loss.

In the above it is always assumed that the radiation may be neglected; this assumption is the basis of the approximation employed here.

With a given thermoelectric combination the above considerations are also valid for the thermal electromotive force E_{th}. The sensitivity of the thermo-electrometer is of course proportional to the thermoelectric force per degree. The current through the galvanometer i_m is the following, when R_m is the resistance of the meter:

$$i_m = \frac{E_{th}}{R_e + R_m},$$

where R_e is the internal resistance of the thermo-couple.

Connection of the measuring instrument to the thermocouple

We may continue with our considerations along two lines.

a) We may begin with a given thermocouple with resistance R_e, and ask what meter we must use in order to obtain the greatest deflection with a definite primary current i. To answer this we must calculate the voltage on the meter:

$$E_m = E_{th}\,\frac{R_m}{R_m + R_e}\,;\;\frac{E_m}{E_{th}} = \frac{R_m/R_e}{R_m/R_e + 1} \cdot (6)$$

Fig. 4 gives a graphic representation of (6). The requirement is that the sensitivity to voltage of the meter (*i.e.* the deflection at a definite voltage on the meter) shall be so great that the meter has the desired deflection with the voltage calculated from (6). When $R_m = R_e$ the thermo-

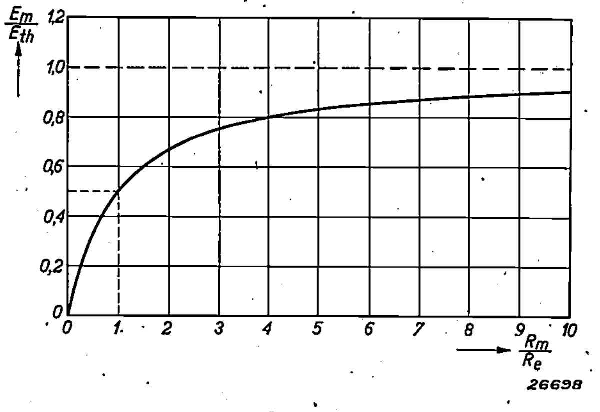

Fig. 4. Drop in terminal voltage of the thermocouple, when current is taken off by the meter. E_m/E_{th} is plotted as a function of R_m/R_e. When $R_m = R_e$ the most favourable energy transfer takes place.

couple gives off the greatest amount of energy to the meter. A meter with this resistance therefore may be the least sensitive to energy ($\frac{\text{deflection}}{\text{energy}}$ may be smallest); it is usually also the cheapest meter. A more sensitive meter may, however, also be used.

b) We may begin with a given meter, and ask how the thermocouple must be constructed in order to give the greatest deflection on the meter. This problem is not so simple. It is clear that with a given thermal electromotive force, and therefore a given temperature of the junction, the voltage on the meter is highest when the resistance of the thermocouple is zero (see *fig. 5a* which is the same as fig. 4, but with a different abscissa). We have, however, seen that the loss of heat through the thermocouple becomes greater as the resistance is lowered, while the temperature increase of the junction is decreased by greater loss of heat through the thermocouple. The thermal electro-motive force is therefore diminished if the resistance of the thermocouple is reduced (see *fig. 5b*). The combination of these two effects is shown graphically in *fig. 5c*, where the voltage on the meter is plotted as a function of the resistance of the thermocouple at a constant current through the filament. If we pass from a high value of the resistance to a low value, the current through the meter at first rises because the total resistance falls, while the thermo-electromotive force changes very little; the current then falls again because the thermo-electromotive force falls and the total resistance no longer changes sufficiently. Fig. 5c is valid for a definite meter resistance; if the meter has a higher resistance, the maximum is shifted to the right, if the resistance is lower, it is shifted to the left.

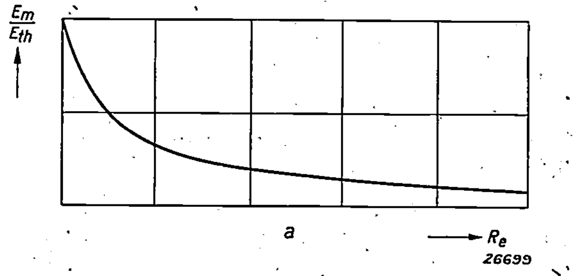

Fig. 5. *a*) Voltage on the meter as a function of the resistance of the thermocouple at a constant thermo-EMF.

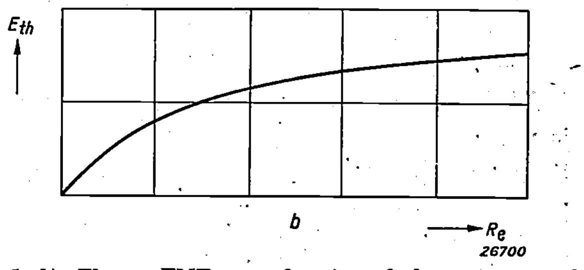

Fig. 5. *b*) Thermo-EMF as a function of the resistance of the thermocouple at a constant current through the filament.

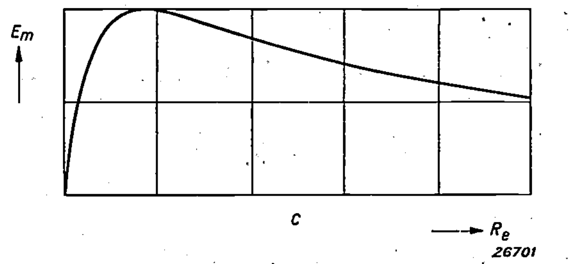

Fig. 5. *c*) Voltage on the meter as a function of the resistance of the thermocouple at a constant current through the filament.

Further discussion of the characteristic of the thermo-ammeter

In the deviation of the formulae it is assumed that ϱ, λ and δ do not depend upon the temperature. The derivation is therefore only valid for small increases in temperature. If these increases are not small, the dependence on temperature of these quantities must be taken into account. In that case, however, the differential equation is insoluble, so that the influence of this dependence must be estimated in some other way. This can be done very roughly by first giving the constants in the solution of the differential equation the values at room temperature, and then the values at the temperature of the middle of the filament; the difference between the two results indicates in any case the direction in which deviations may be expected. We shall discuss this for the various quantities which may be dependent on temperature.

With each individual quantity it will be found that the result of its dependence on temperature is that the characteristic is no longer truly quadratic. In favourable cases the deviations due to the different quantities may compensate each other.

ϱ. The temperature coefficient of the specific resistance has very different values for different materials; it is for example positive and about 0.4 per cent per degree centigrade for copper, platinum and iron, and practically zero for constantan and manganin. If ϱ is positive the resistance increases with increasing current through the wire, so that with a high current relatively too much energy is developed. The deflection of the galvanometer will thus be greater than proportional to the square of the primary current.

λ. The temperature coefficient of heat conductivity is much smaller than that of electrical resistance; it is weakly negative for the pure metals and positive for constantan and manganin. Experience has shown that its influence is slight; the reason for this lies in the above-mentioned fact that a small value of λ on the one hand improves the heat insulation of the filament, and on the other hand it prevents the conduction of heat to the thermocouple. A change in λ therefore will in the first approximation only influence the variation of temperature along the filament, but not the temperature of the middle. This does not of course hold for the loss of heat through the thermocouple, which is proportional to the heat conductivity of the thermocouple. If the latter has a positive temperature coefficient the result is that the deflection increases more slowly than proportional to the square of the current.

σ. The Stefan-Boltzmann law holds for the heat radiation of a black body. According to this law the amount of energy radiated per second and per square centimetre of the body is proportional to the fourth power of the absolute temperature. If the radiation takes place in surroundings at the temperature T_0, the radiation is then:

$$R = S\,(T^4 - T_0^4)\,\frac{\text{cal}}{\text{cm}^2\cdot\text{sec}\cdot\text{degree}^4}\,.$$

If the temperature differences under consideration are not great, this expression may be replaced by:

$$R = S\,a\,(T - T_0) = S\,a\,\tau.$$

The quantity a, however, now depends upon the temperature difference, and as a second approximation we may write:

$$a_\tau = a_0\,(1 + \alpha\,\tau).$$

The constant α is approximately independent on the temperature difference, and between 20° C and 100° C it is equal to + 0.6 per cent per degree Centigrade. The dependence of δ on temperature results in the fact that the increase in temperature is less than the formula indicates, and therefore the deflection of the galvanometer changes less than proportional to the square of the primary current.

In addition ϱ, λ and σ, the thermo-e.m.f. per degree, the resistance of the thermocouple and the heat resistance from the middle of the filament to the junction of the thermocouple may also depend upon the temperature. In many cases the temperature coefficient of the thermo-e.m.f. may be neglected. The resistance of the thermocouple forms merely a part of the resistance of the meter circuit, so that its influence may be slight. It is very difficult to make an estimation of the heat resistance of the connection between filament and thermocouple, and of the way in which this depends upon the temperature; its influence becomes less according as the heat resistance of the thermocouple is increased. With poorly made junctions, however, this influence may very well be felt, so that the connection must be very carefully made.

From the above discussion it follows that particularly the radiation can cause great deviations from the quadratic variation of the characteristic. It is therefore desirable to make the radiation as slight as possible. We have seen in the discussion of the expression for the temperature at the middle of the filament that this temperature is independent of the dimensions for a given resistance of the filament. The radiation, however, is proportional to the surface of the wire; to make the radiation small therefore the dimensions must be kept small.

At small values of K the radiation constant σ no longer occurs in the formulae. The value of K is therefore a measure of the influence of radiation. K is proportional to $l/\sqrt{d}$; the energy sensitivity on the other hand is proportional to l/d^2.

With constant energy sensitivity ($l/d^2 = C_1$), therefore, $C_1\,d^{3/2}$ must be made as small as possible; d must therefore be small.

The same holds for the dimensions of the thermocouple. Wires with different dimensions but with the same electrical resistance cause the same loss of heat; if the dimensions, and therefore the surface, are small, the radiation has the least influence. Therefore short thin wires must be used to construct thermojunctions with a quadratic characteristic. By this means the heat capacity of the thermo-couple is decreased at the same time, so that the final deflection of the meter is more quickly attained. A limit is set to the reduction of the dimensions, however, because such reduction diminishes the ability to withstand overloads.

Overloading and sensitivity

We have seen that the highest temperature of the filament does not occur in the middle of the wire, but somewhere between the middle and the terminals. If the current through the filament is made too high, the filament will burn through at those spots. It is therefore important that the ratio of this maximum temperature to the temperature at the middle should not be too great. A measure for this ratio, which is not easy to calculate, is the ratio of the average temperature of the wire to the temperature at the middle. For small values of K it is:

$$\frac{\bar{\tau}}{t_M} = \frac{2/3 - 4/15\,K^2 + \dfrac{2\,M}{\sqrt{\sigma\,\lambda\,u\,q}}\,(1/6\,K - 13/180\,K^3)}{1 - 5/12\,K^2}$$

At a given value of K (thus at a given value of $l/\sqrt{d}$) the ratio becomes greater if d is decreased; the dependence of d, however, is less, the smaller the value of K. Further it is clear that a thin wire is more quickly damaged than a thick one at a given temperature; this is a reason for not using too thin a wire.

With large current thermo-junctions K is so small that the maximum temperature is practically equal to the temperature at the middle. In this case the permissible current is not limited by the melting of the wire, but by the overheating of the cement with which the thermocouple is fastened to the filament. It is very difficult to find a cement which is attacked only at a higher temperature than that which injures the filament.

The overloading of thermo-junctions for weak currents is therefore determined by the velocity of evaporation or the melting point of the filament, that of those for heavy currents by the properties of the cement.

In direct relation to overloading is the indication of the current for which the thermo-junction is intended. Since the degree of attack on the filament upon overloading depends on the length of time during which a given current is sent through the wire, it is impossible to indicate a maximum permissible current, if the time is not also indicated. A suitable measure is for example the current at which the filament remains absolutely undamaged for one minute. At the same time it is desirable to know the current at which the instrument can be used continuously. In practice it is then desirable to use a still lower current in order to have a margin of safety.

In the case of the Philips thermo-junctions the current is indicated at which the thermoelectric force is 12 mV; in addition the maximum current is given at which the instrument may be used continuously and the current at which it is quickly burnt through. We may call a thermo-junction which gives 12 mV at 10 mA, a 10 mA instrument.

It is therefore only possible to speak of a thermojunction for a definite maximum current; the current at which the full deflection of the meter is obtained depends upon both thermocouple and meter.

The following must also be kept in mind. Suppose that we have a certain meter and the corresponding thermocouples for 10 mA and 20 mA. The resistance of the filaments of these couples are respectively 75 and 23 ohms. In a given circuit the resistance of the 10 mA couple may be too high, for instance when the couple must form part of a tuned circuit. In this case the resistance of the 20 mA couple is perhaps permissible. That couple may then be used if a more sensitive meter is available which gives the full deflection at 10 mA, when used with the 20 mA couple. In certain cases therefore a dial instrument will not be used but a very sensitive mirror galvanometer, in order to be able to use a thermocouple with low resistance for low currents also.

Reversal effect

The first thermo-junctions constructed exhibited a strong reversal effect, *i.e.* the deflection changed when the direction of the current of the primary direct current was changed. There may be two causes for this phenomenon:

a) One of these may be the Peltier effect, which is the reverse of the thermo-electric effect. When current is sent through a thermocouple the junction becomes colder if, upon heating it, a current would

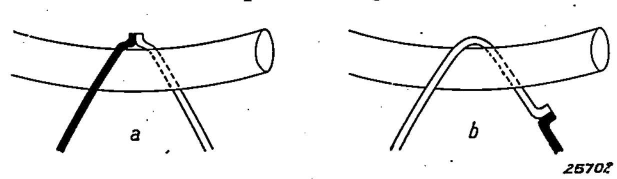

Fig. 6. *a*) Peltier effect in thermocouples. The junction of the couple lies on the filament and therefore part of the heating current passes through it.
b) Reversal effect in thermocouples. The junction of the couple lies near the filament, but the heating current passes through a short section of one of the wires of the thermo-couple.

have occurred in the same direction, and vice versa. The Peltier effect therefore occurs in the original thermo-electric ammeter by Klemenčič, and also in the case of the commonly used bridge circuits which we shall not discuss here. With modern thermocouples also the effect may appear when the thermo-junction is not electrically insulated from the filament, and when it lies exactly on the filament (see *fig. 6a*). In this type of construction part of the primary current also passes through the junction of the thermocouple and gives rise to the Peltier effect. This may be avoided by placing the junction near but not on the filament.

b) Even when the junction does not lie on the filament a reversing effect may also appear (see *fig. 6b*). At the point where filament and thermocouple touch each other, part of the primary current will always pass through part of the thermocouple. Due to the resistance of the latter a potential difference will arise which leads to a small extra current through the meter whose direction is reversed when the direction of the primary current is reversed. In order to avoid this the thermocouple must be insulated from the filament; when this has been done the reversal effect will be found to have disappeared and the thermocouple can be calibrated with direct current.

Speed of indication

The final deflection is not rapidly reached with a thermo-ammeter. The main reason for this is that a temperature equilibrium must be established in the filament and the thermocouple. Since the thermocouple itself does not carry the primary current, the heating up of the junction must take place from the filament by conduction, for which some time is necessary. A type of construction is possible in which the thermocouple also serves as filament (see *fig. 7*); in this case the heat is developed in every element of volume of the thermocouple, so that only a small amount of heat needs to be transported by conduction, and the final deflection is quickly attained. An elaboration of this construction is formed by the bridge arrangements in which a number of thermocouples are joined to form a square; the disadvantage of this method is the presence of a strong Peltier effect.

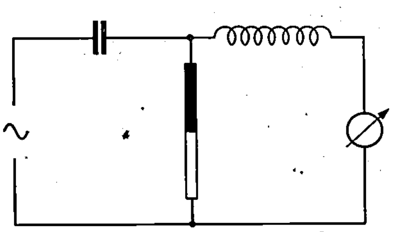

Fig. 7. Thermocouple with direct heating. The condenser is introduced so that the direct voltage given by the couple may not be short-circuited by the source of alternating current. The self-inductance is introduced to prevent the alternating current from passing through the direct current metre.

As was noted above, the heat capacity also has some influence on the speed of indication; therefore small dimensions are desirable. The resistance of the thermocouple may also effect the speed of indication because with too low resistance the damping of the metre is considerable.

Dependence of the indication on frequency

The heat development in a wire is independent of the frequency of the alternating current used for heating up to very high frequencies. At very high frequencies, however, the so-called skin effect appears, and the resistance, and therefore the heat development also, are greater than at low frequencies. For the thickness of wire used the influence of this effect is only appreciable at wave lengths of less than one metre, so that it need not usually be taken into account.

It is quite another question with the influences of various capacities (see *fig. 8*). The capacity in parallel with the filament is the cause of part of the current to be measured not passing through the filament, and the result is that the deflection of the metre is too small. This is sometimes compensated by the self-induction of the filament. The capacity between the supply leads has the same effect. There is further capacity between the filament and the thermocouple which is connected with earth capacitatively through the galvanometer. This capacity is further increased by the fact that the leads of the filament and those of the thermocouple are close together in the pinch

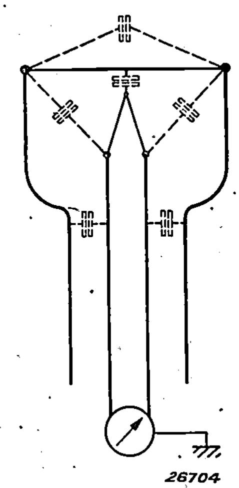

Fig. 8. Diagram of a thermocouple with the various parasitic capacities.

of the bulb and in the base. This capacity may be partially reduced by using thermocouples without bases in measurements at very high frequencies. The third kind of capacitive influence, the capacity of the leads to earth, is hereby also very much reduced. In order to overcome the last-mentioned influence it is absolutely essential that one side of the filament be earthed.

The influence of these parasitic capacities becomes greater the higher the resistance of the filament. A thermocouple with a filament having a resistance of 20 ohms gives reliable indications at a wave length of 6 metres within 1 per cent; when the thermocouple is used without a base the indication is still reliable within 1 per cent at wave lengths of 3 meters.

Advantages in the use of thermo-ammeters

In conclusion we give the following list of the advantages connected with the use of thermocouples for measuring alternating current.

a) The effective value of the alternating current is measured because use is made of the development of heat, and therefore of a mechanism which varies essentially with the square of the current. For this reason the measurement is very little affected by the presence of higher harmonics in the current measured, and it is not at all affected by the phase of these harmonics. This is a common property of all thermal instruments.

b) Up to a very high frequency the indication is independent of the frequency of the current measured.

c) The energy consumption is small compared with that of dynamometers and soft-iron instruments.

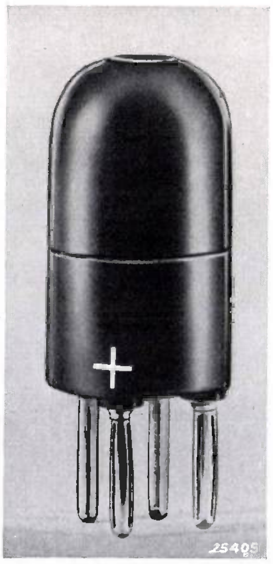

Fig. 9. Thermocouple assembled.

d) The delicate part of the measuring arrangement, the thermocouple itself, is easily renewable; in this connection the quadratic, or at least reliably constant, characteristic is also important.

Survey of the Philips thermojunctions

The Philips thermo-electric meters are constructed with the above considerations in view. *Fig. 9* is a photograph of a fully assembled thermocouple and *fig. 10* shows the interior of such an instrument. The cover serves not only for decoration, but also for shielding the instrument from heat radiation from the surroundings.

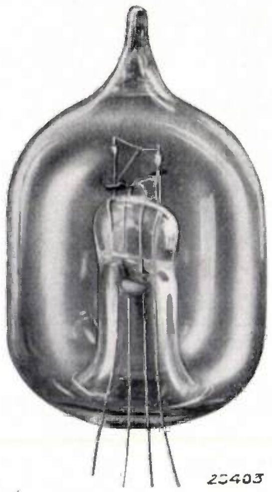

Fig. 10. Interior of a thermocouple.

The *table* gives a survey of the various properties of the instruments. With all types the temperature coefficient of the resistance of the filament is small; for most measurements it may be neglected. The indication is practically quadratic up to half the filament current, where the thermo-electric force is 12 mV; the deviation from the quadratic variation is 2 per cent at the most in this range, the thermo-electromotive force is about 4 mV at that current, so that a full deflection can still be obtained with sensitive dial instruments. The thermocouple is insulated electrically from the filament; the instruments can therefore be calibrated with direct current.

Type number	12 mV EMF at approx. mA	resistance of the filament ohm	resistance of the thermocouple	continuous use up to mA	maximum permissible current mA
Th1	10	75	5	15	20
Th2	20	23	3	30	40
Th3	40	7.5	3	75	100
Th4	100	2.2	3	150	200
Th5	200	1.1	3	300	350

ASSEMBLY OF WIRELESS RECEIVERS

The photograph shows two conveyor belts at which radio receivers are being assembled. The belts convey parts from one worker to the next. The girl takes the arriving piece of work from the belt, performs the necessary operations which consist in the mounting of parts, taken from the stock at her side, into the chassis by welding, soldering or screwing. For a smoothly running process the girl should just have finished her task when the next piece of work arrives. In this way the chassis is gradually assembled and at the end of the belt it can be mounted in its cabinet.

THE ILLUMINATION OF THE NEW LEAGUE OF NATIONS PALACE IN GENEVA

by L. C. KALFF. 628.972

The League of Nations Palace is intended for the housing of all League organisations (except the Bureau de Travail which is already definitely established elsewhere), which up to now have been accommodated in various buildings in Geneva. The different parts of this organization are naturally very diverse in character. There are administrative departments which are in daily use, assembly halls which are scenes of activity now and again for several days or weeks when there are congresses and meetings of international committees, and a considerable portion of the building is intended for the very momentous and festive gatherings on important occasions, for which banquet halls and monumental rooms are necessary. The extensive group of buildings includes four sections, as may be seen from the ground plan (*fig. 1*), namely the offices of the secretarial department, the section for the League Council, in the main axis the hall of the representatives of the Assembly with committee rooms surrounding it, and finally the library.

The illumination of the secretarial department is mainly ultilitarian, and its design is for the most part determined by the dimensions of the mains. The latter are not very heavy so that in the offices, which are in light colours, diffuse direct lighting had chiefly to be used, with one diffuser (type DM; *fig. 2a*) with a lamp of 150 decalumens (111 W) per desk, so that an illumination of about 65 lux is obtained on the desk, and at the same time a sufficient general illumination (*fig. 3*). Only in several low-ceilinged rooms in which the typists sit, and where one would be too much restricted in the placing of the desks if direct lighting were used, are indirect reflectors (type WX; *fig. 2b*) used, with a uniform illumination of about 110 lux. In this way the placing of the desks is left quite free (*fig. 4*). In the evening this wing of the building with its many lighted windows makes a very fine effect from the „Cour du Sécrétariat", as may be seen from the vignette at the head of this article.

In the section for the League Council the rooms are lighted in the same way with the exception of several rooms for which special plans were made in collaboration with the architects. In the first place there is the "Salle du Conseil" which offered particularly difficult problems. The entire hall is

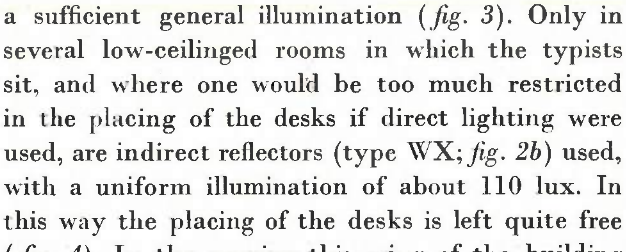

Fig. 1. Plan of the League of Nations Palace.

executed in polished grey marble and so called "Camaieu painting" in grisaille on a gold background. The result is poorly reflecting, dark walls and ceiling, with locally strongly reflecting surfaces. be in a poor light as seen from the hall, namely only in silhouette.

The committee rooms all open on the approximately 160 metre long gallery which is 10 metres high.

Fig. 2. *a* DM Diffuser

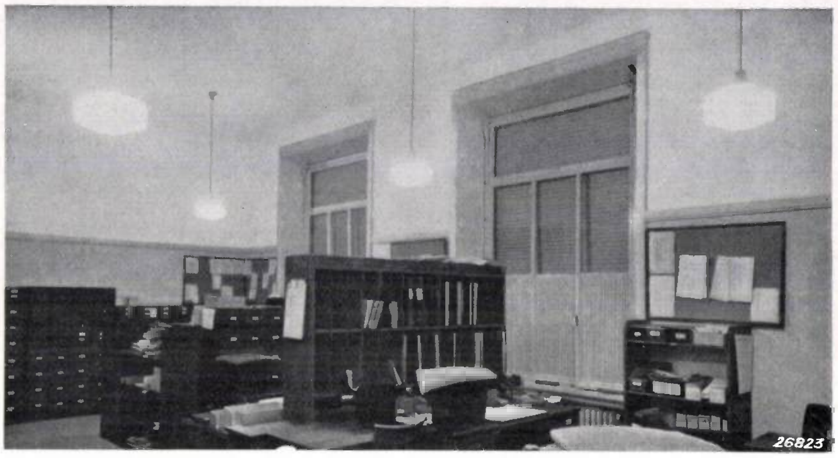

Fig. 3. Illumination with diffusing fixtures DM of one of the rooms of the secretarial department of the League of Nations.

The only possible illumination was the installation of extensive ornamental lighting fixtures with large diffusely radiating surfaces with as low brightness as possible. For this purpose Perzel of Paris constructed large bronze bowls 2 metres in diameter with thick glass rings behind which the lamps were placed in white reflectors (*fig. 5*). The light-coloured curtains before the five enormous windows are illuminated decoratively from above with silver reflectors, but not too intensely, since otherwise the Council, with its back to the windows, would

The gallery ends in square chambers which are lighted indirectly by four large bronze standards by Perzel, with lamps of 500 watts which light the ceiling (*fig. 6*), and are provided with a frosted glass rim to avoid sharp-edged shadows on the walls. An

Fig. 4. Indirect illumination with WX fixtures; the placing of the desks is here independent of the lighting.

Fig. 2. *b* WX Indirect lighting fixture.

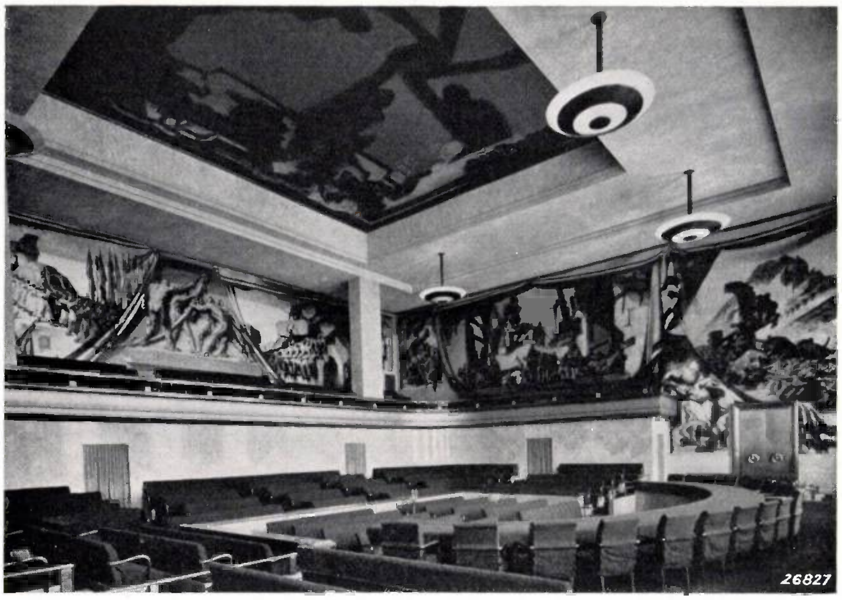

Fig. 5. View of the "Salle du Conseil".

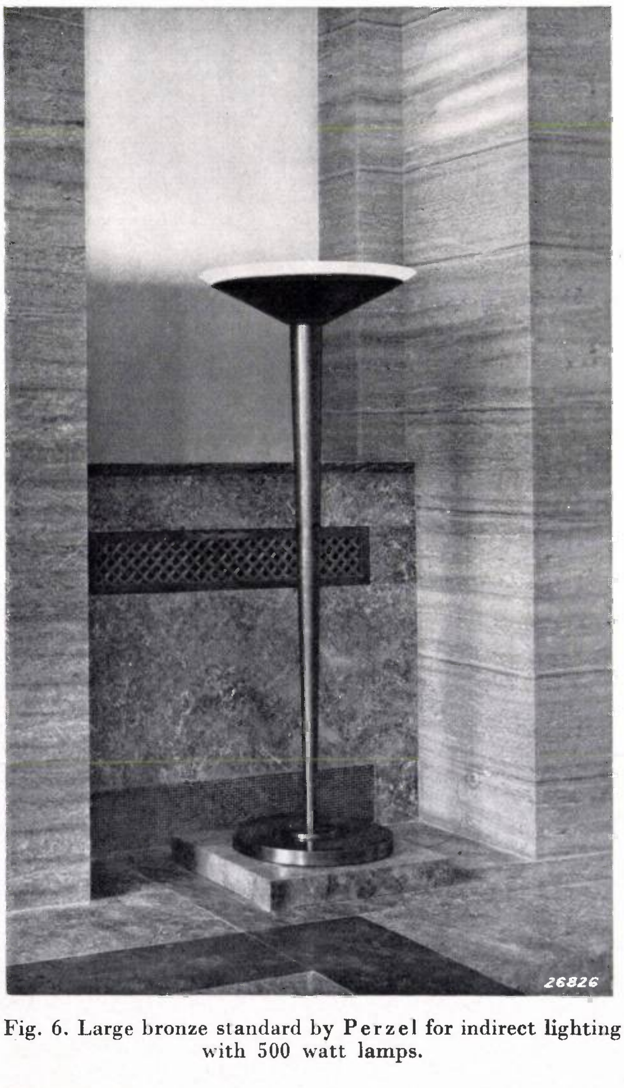

Fig. 6. Large bronze standard by Perzel for indirect lighting with 500 watt lamps.

indirect illumination with small 40 watt lamps at distance of 20 cm in a continuous chromium plated reflector forms the main illumination of the gallery. Most of the committee rooms are lighted indirectly from vaulted portions of their ceilings (*fig.* 7). Three small silvered parabolic reflectors per linear metre are introduced in recesses on rails previously fixed. Each reflector contains a lamp of 40 watts and a very pleasant uniform illumination is obtained. In front of every reflector a convex, partially frosted glass had to be introduced to avoid light spots, while at the same time the light reflected from the recesses themselves further promotes this uniformity. Since many of the recesses had already been finished before the lighting was projected, the correct proportioning of the lighting system was very difficult. In some cases, instead of the silvered reflectors, a continuous chromium plated reflector had to be mounted with five lamps per metre to get a uniform illumination.

The "Salle des Pas Perdus" (*fig. 8*) which is treated in a particularly decorative manner, opens off the middle of the great gallery, and serves as a foyer to the great "Salle de l'Assemblée". The dark polished granite floor and the marble columns (*fig. 9*) here again presented great dificulties. The main illumination was finally designed as indirect lighting from a built-in recess in the gallery which runs along the 56 metre long hall (see cross section *E* in fig. 8). Five reflectors (type SC 170; *fig. 10*) per linear metre with 60 watt lamps were installed in this recess and light the ceiling (16.8 kW). To limit direct vision of the lamps from the gallery white metal partitions were introduced between the reflectors, while the whole was made dust proof with bent glass plates which are slightly frosted on the inside to eliminate light streaks. The short sides of the long hall have no reflectors as the portions of the ceiling at the ends would otherwise be more intensely lighted than the rest, but for the sake of uniformity of the whole only bare 40 watt lamps with five per linear metre are mounted there behind the bent glass plates. The other long side of the hall has nine enormous bronze window frames 12 metres high, each frame enclosed by a bronze moulding. In this moulding (detail D in fig. 8) there is a continuous row of tubular lamps "Philinea", 80 watts per metre) with white enamelled reflectors which light the white surface of moulded plaster between the marble columns. This gives a very dignified but also festive character to the hall. For this illumination 9 times 42 "Philinea" lamps 50 cm long were necessary altogether (15.12 kW). Between the windows are eight bronze standards (fig. 6) by Perzel with 500 watt lamps which give light accents (cf fig. 9). In total 16 + 4 + 15 = 35 kW is installed.

Fig. 7. Indirect lighting of one of the committee rooms.

The great "Salle de l'Assemblée" (*fig. 11*) presented a particularly complicated problem because in this huge hall the illumination had to be entirely part of the architecture (*fig. 12*), and not only on the podium and the speakers chair, but also on the floor, was a very good illumination necessary, while also in the two deep galleries the public and the press had to be able to read and write. The main illumination was obtained in imitation of the daylight illumination by 132 silvered reflectors with lamps of 300 and 500 watts above the glass middle section of the ceiling. This section of the ceiling consists of so-called thermolux glass, two layers of clear glass with a layer of glass wool between, and is very fine in appearance and gives good diffusion.

Fig. 9. View of the "Salle des Pas Perdus".

Fig. 10. Reflectors, types SC 170 and SC 130.

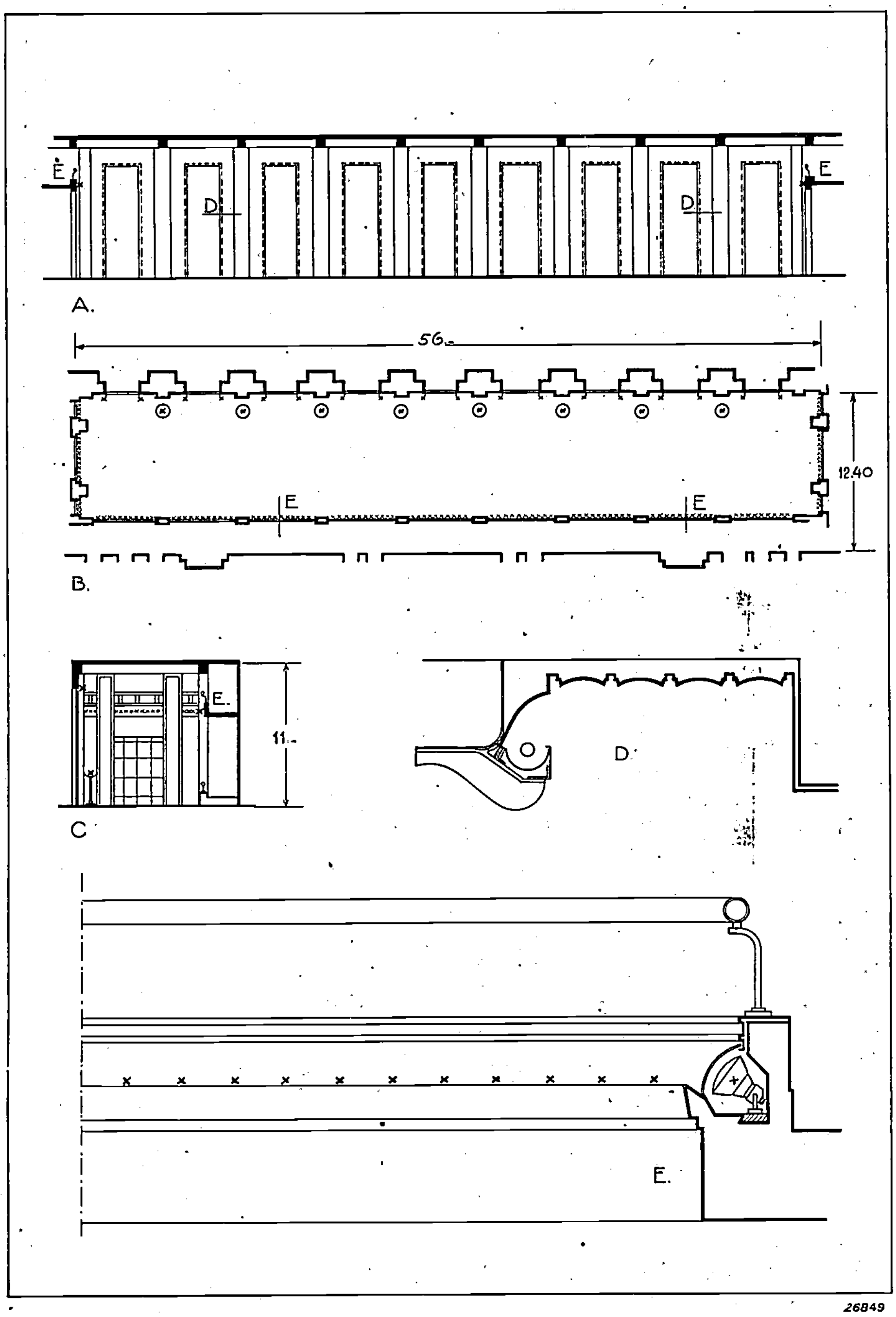

Fig. 8. Construction scheme of the "Salle des Pas Perdus".
A longitudinal cross section of the hall,
B ground plan of the hall,
C cross section through the width of the hall,
D horizontal cross section of the light recesses around the windows,
E detail of the indirect ceiling illumination in the gallery, in vertical cross section.

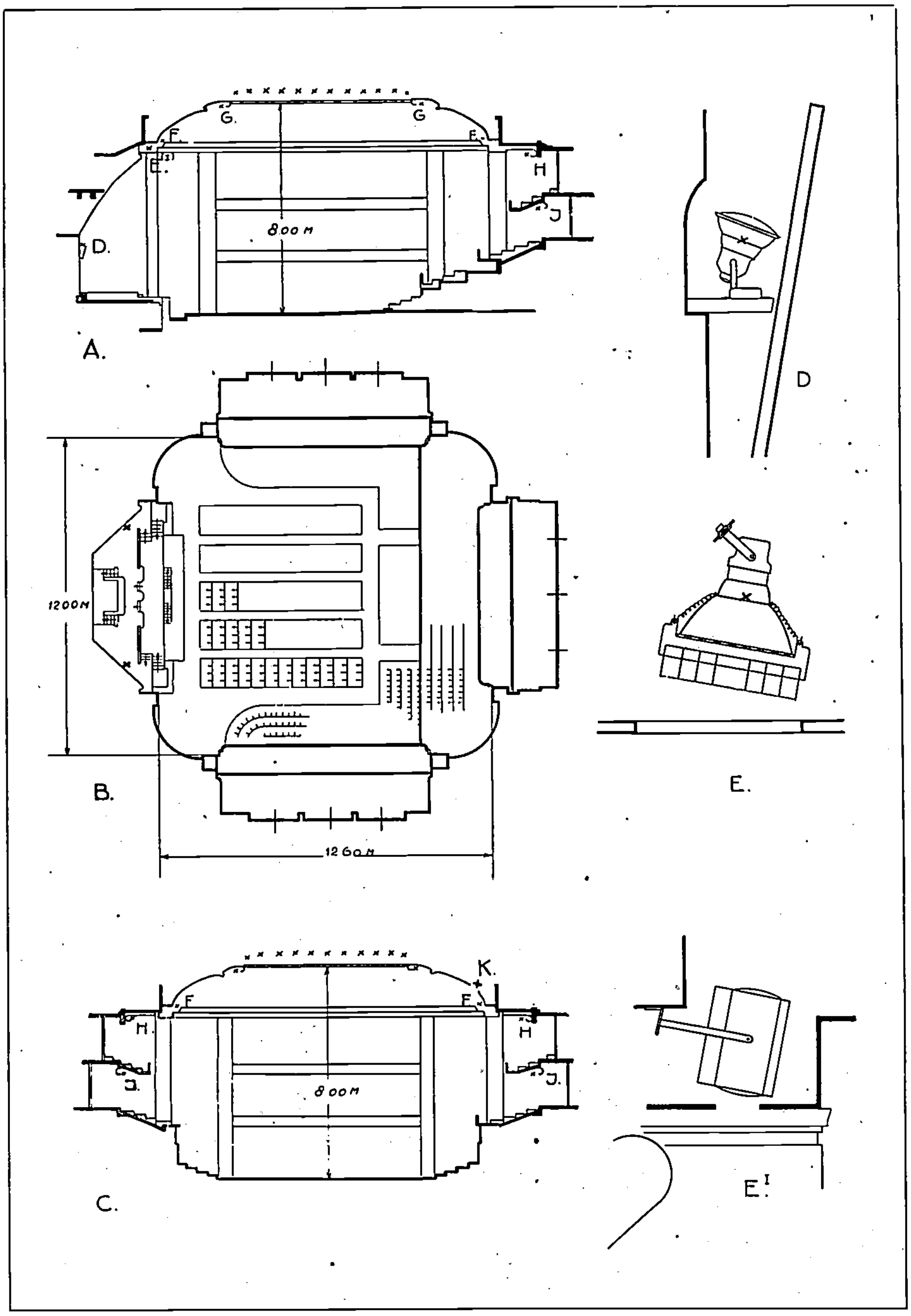

Fig. 11. Construction scheme of the "Salle de l'Assemblée".
A longitudinal section of the hall,
B ground plan of the hall,
C crosswise section of the hall,
D detail of the rear wall of the podium in vertical cross section,
E and E^I, normal and film lighting installations in the light recess above the podium; in fig. 12 the latter is visible as four light spots directly under the loudspeakers.

A flat white band runs around this glass surface and is lighted by 25 watt lamps in a bronze recess (at G in fig. 11). Outside of this there is a wide curved border of plaster which has been sprayed

Fig. 12. Vieuw of the „Salle de l'Assemblée".

with a soft sound-absorbing material. This border is uniformilly lighted by a row of silvered reflectors with 150 watt lamps hidden in a plaster recess which is gilded on the outside (at F in fig. 11). These lamps can be reached for repair above the hall on travelling bridges above the glass and

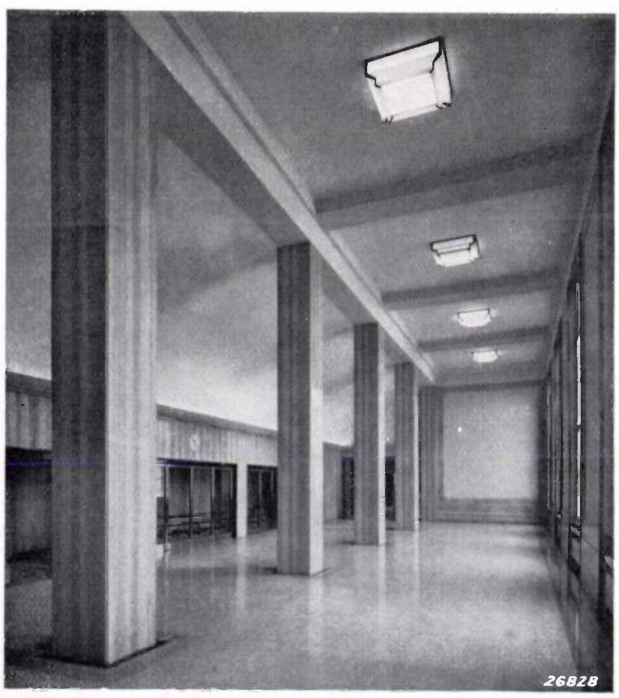

Fig. 13. Lighting of the side galleries of the "Salle de l'Assemblée".

behind shutters for the recesses. These shutters are necessary primarily for the ventilation system. The ordinary illumination of the podium is by five silvered reflectors type SC 255 with 200 watt lamps which are hidden in the gilded recess above the podium. To prevent direct vision into the reflectors from the hall they are provided with louvres (detail E in fig. 11). In addition to this ordinary illumination of the podium there is also an installation for film lighting (detail E^I in fig. 11), which may be seen as four spot lights in fig. 12 directly under the loud speakers. The marble wall behind the podium has a small light recess with 25 watt lamps in reflectors type SC130 (at D in fig. 11), for the purpose of placing a light accent on the chief detail of the hall. The galleries have indirect illumination from long strips of white plaster in the ceiling which are lighted with 25 watt lamps in chromium-plated metal reflectors (at H and I in fig. 11).

In the preceding paragraph we mentioned the installation for film lighting. To facilitate news reel reporting during meetings and discussions it was necessary to be able to illuminate intensely but for brief periods the podium with the speaker, and also certain sections of the hall where the delegations from the different member countries sit, very intensely for short times. In order to do this in a way which is the least disturbing for the speakers it must be possible to raise the level of illumination at the place to be photographed to 1500 lux or higher for several minutes whenever desired. For this purpose four permanent spot lights with mirrors and condenser lenses are installed above the podium, each of which has a power of 3 kW. These lamps light the chairman's table, and the speakers position. Since the light comes from very high up, it is not disturbing, however, side lighting from the hall must also be provided since otherwise the lighting of the subjects to be photographed would be very poor.

In addition there are five round openings in the ceiling on the right side of the podium which can be closed (see fig. 12, in the lighted part of the ceiling at the left). Behind these openings are five large mirror reflectors, two permanently directed on the podium and provided with 5 000 watt lamps, and three with 10 000 watts which can be directed on 25 different sections of the floor of the hall. The operator of these lamps is in telephonic communication with the film cabins which are situated under the first gallery.

The side galleries of this hall have one part which

Fig. 15. *a* FLC Fixture.

is lighted indirectly from a marble alcove by 25 watt lamps in reflectors of type SC 130 (*fig. 13*). These lamps provide a uniform illumination of the sloping portion of the ceiling formed by the underside of the first gallery. The rest of the space is lighted by four large square ceiling ornaments by Perzel of Paris. Behind the white diffusing glass of each ornament there are 18 lamps of 65 watts.

It would lead us too far afield if we were to describe in detail the lighting of every hall and room in the building. We shall therefore conclude with a description of the temporary flood-light installation of the front of the building giving on the "Cour d'Honneur" which was used at the opening (*fig. 14*). The requirement in this case was that on evenings of festive occasions the outside of

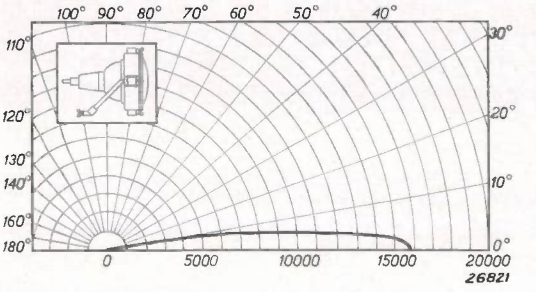

Fig. 15. *b* Light distribution curve of a FLC mirror reflector.

the building should be well lighted, and in fact so well lighted that it might be seen from the other side of Lake Geneva, which is fairly broad at this point. At the same time, however, it was required that persons in the building should be able to reach the gardens by way of the grand staircase without being blinded. It was of course difficult to find a single solution for both problems. Fig. 14 shows fairly well the result achieved.

For this purpose three different batteries of 24 mirror reflectors, type FLC (*fig. 15*), with searchlight lamps of 1000 watts were set up on the roof of the wing leading to the secretarial department. The intention is to make a permanent installation some time in the future, which will give not only a good illumination for a distant view of the building, but which will also illuminate the "Cour d'Honneur" for those present in it.

Fig. 14. Lighted exterior giving on the "Cour d'Honneur".

A SIMPLE ARRANGEMENT FOR THE MEASUREMENT OF THE SPECIFIC RESISTANCE OF LIQUIDS

by A. CLAASSEN.

621.317.73 : 537.311.3

A measuring arrangement is discussed which consists of the simple "Philoscope"[1]), universal measuring bridge, an alternating current generator for 1000 cycles per second and a measuring vessel. This set-up is intended especially for carrying out the control measurements on the conductivity of liquids which are regularly made in many industries.

Measurement of the resistance of liquids

In many industries the specific conductivity, *i.e.* the reciprocal of the specific resistance, of liquids must be regularly checked. This may be so in the case of the waste water, the boiler water and the rinsing water in laundries where the progress of the washing process must be determined. In the sugar technique also it is simpler, instead of the elaborate determination of the ash of the sugar, to measure the specific resistance of the sugar solution which is directly connected with the desired weight of ash.

In this article an arrangement is described for the measurement of the specific resistance in which use is made of the simple electrical measuring bridge previously described in this periodical[1]). The problem of the measurement of specific resistance comes down to the measurement of the resistance of a liquid in a measuring vessel which has been calibrated with certain standard liquids. In the measurement of the specific resistance of liquids in practice the difficulty arises that it is not easy to determine the correct length and cross section of the column of liquid to be examined. If for example two equally large electrodes at different potentials are placed at a given distance apart in a liquid, the lines of force will not run absolutely mutually parallel from one electrode to the other, but a certain bending of the lines of force will occur, so that the effective cross section of the column of liquid will be greater than the area of the electrodes themselves, unless the cross section is restricted in space by a non-conductor. If, however, the specific resistances of several liquids have been determined by means of a vessel in which the bending of the lines of force has been made impossible in some way, these liquids can be used for the calibration of a more simply constructed vessel, which at the same time may be better adapted to the requirements of practical measurements. With this calibrated vessel the specific resistances of other liquids can then easily be measured.

[1]) Cf.: Philips techn. Rev. 2, 270, 1937.

The measurement of the resistance of a liquid is a problem which presents various difficulties. It is impossible to measure this resistance with direct current as may be done with a metal conductor, since when a direct current is sent through a liquid situated between two electrodes immersed in the liquid certain changes take place at the electrode (for instance the development of gas or precipitation), which are accompanied by the appearance of definite potential differences between metal and liquid and acting against the current. The voltage applied between the electrodes, therefore, is not the same as the true potential difference between the ends of the column of liquid, and it is impossible to deduce the resistance of the column of liquid from the voltage applied and the curernt measured. The potential differences occurring in such a case are called **polarization voltages** and the phenomenon itself is called **polarization**.

The influence of polarization is obviated by measuring the resistance with an alternating current of sufficiently high frequency. Also if the measuring vessel is suitably constructed the influence of polarization can be practically neglected, since the polarization voltages have no opportunity of building up in the short time during which the current flows in a given direction. We shall first study the influence of polarization in the alternating current circuit.

The decomposition products which are formed at the electrodes due to the current, result in the appearance of an opposing electromotive force V_p which we shall assume to be proportional to the amount of electricity which has passed through, q:

$$V_p = -Pq = -P\int i \, \mathrm{d}t \,, \quad \cdots \quad (1)$$

which is entirely permissible at not too low frequencies. This formula reminds us of the formula for the voltage V on a condenser C with a charge q:

$$V = -\frac{q}{C} = -\frac{1}{C}\int i \, \mathrm{d}t \,. \quad \cdots \quad (2)$$

The polarization voltage V_p corresponds exactly in its behaviour with the voltage which occurs between two condenser plates due to a charge q, when the capacity is:

$$C = \frac{1}{P} \cdot \quad \cdot \quad \cdot \quad \cdot \quad \cdot \quad \cdot \quad \cdot \quad \cdot \quad (3)$$

If R is the resistance of the column of liquid, the ohmic loss of voltage is iR and the voltage on the liquid is equal to the sum of iR and V_p. The measuring vessel therefore behaves like a resistance R and a capacity equal to the reciprocal of the polarization factor P connected in series (*fig. 1*),

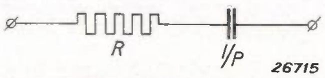

Fig. 1. Substitution diagram for the measuring vessel as part of an alternating current circuit.

so that its impedance for sinusoidal alternating current with an angular frequency ω is:

$$Z = \sqrt{R^2 + \frac{P^2}{\omega^2}} \cdot \quad \cdot \quad \cdot \quad \cdot \quad \cdot \quad \cdot \quad (4)$$

The influence of polarization is, according to formula (4), small when:

1) the frequency is sufficiently high,
2) the polarization factor is kept small, which can be done by giving the electrodes a large surface, since P is inversely proportional to the area of that surface,
3) the resistance of the liquid is high.

Care must be taken that the alternating voltage be as free of harmonics as possible. According to formula (4), the impedance of the measuring vessel will no longer be the same for the different harmonics as soon as the polarization is appreciable. This means that it is then impossible to bring the measuring bridge into equilibrium for a non-sinusoidal alternating current, since the different harmonics would require different adjustments.

The measuring arrangement

The arrangement which has been developed for the measurement of the resistance of liquids consists of:

1) the "Philoscope" universal measuring bridge type GM 4140 (*fig. 2*), which has previously been described [1]),
2) a generator for 1000 cycles: GM 4260 (*fig. 3*),
3) the measuring vessel GM 4221 (*fig. 4*).

In order to use the measuring bridge with a separate 1000 cycles generator the normal bridge voltage should be disconnected from the input terminals. The generator GM 4260 offers the possibility of choosing between a frequency equal to the mains frequency and a constant frequency

Fig. 2. "Philoscope" universal measuring bridge, type GM 4140.

of 1 000 cycles independent of the mains frequency for supplying the bridge part of the measuring bridge. The second frequency is especially intended for the measurement of resistance of liquids, since it is sufficiently high to allow one to neglect polarization with the type of measuring cell used, and on the other hand it prevents disturbance from unfavourable capacitative influences. The oscillator valve in the generator is provided with a variable gridleak, so that it is possible by adjusting for weak oscillation to ensure that the alternating voltage is purely sinusoidal at all loads.

The measuring vessel (see fig. 4) is constructed as an immersion cell. The platinum electrodes, each with a surface of about 1 cm², are placed

Fig. 3. GM 4260 Generator for the excitation of alternating current with a frequency of 1000 cycles per sec.

vertically at a distance of about 8 mm from each other and joined to the copper leads by means of platinum wires fused into the glass. The copper leads in turn make contact with the terminals. The outer surfaces of the electrodes are entirely covered with glass which has been applied in the

25776

Fig. 4. Measuring vessel GM 4221 for the measurement of the specific resistance of liquids.

molten state, and are held rigidly in place by a glass support. The electrodes are completely protected by the robust outer glass jacket of the measuring cell. This outer jacket has openings at the bottom and at the side to make it easier to fill it with the liquid.

The following advantages are obtained by this method of placing the electrodes:

1) no air adheres to the electrodes upon immersion, as may be the case with horizontal electrodes and which might produce large errors in the determination of the resistance;
2) the lines of force in the liquid are quite fixed, since they are bounded by the outer glass wall, while the liquid automatically reaches a sufficient height above the upper edge of the electrodes upon immersion.

The specific resistance ϱ of a liquid is proportional to the resistance R of the measuring vessel filled with the liquid, and we may write:

$$\varrho = cR \quad \cdots\cdots\cdots \quad (5)$$

in which c is a constant characteristic of the measuring vessel (cell constant).

This cell constant is determined by calibrating the vessel with a solution of known specific resistance. The value of the cell constant is engraved on its neck. (For cells of the above construction it is about 2 in most cases).

An expedient which is very important for the practical avoidance of polarization is the platinizing of the electrodes. The electrodes are covered with a thin layer of platinum black (or platinum sponge) by electrolysis, so that their effective surface is many times enlarged. The platinizing of the electrodes has one objection, namely that the very finely divided platinum sponge adsorbs dissolved substances from the solution very strongly (especially acids and bases, to a lesser degree salts), and does not immediately give them up in more dilute solutions. Platinizing is especially recommended for the measurement of solutions with a low specific resistance. If one is concerned only with solutions which do not have a very low specific resistance, greater than 1 000 ohm cm for instance, it is better to carry out the measurements with bare metal electrodes, which can be obtained by dissolving the platinum black in *aqua regia*; the danger of polarization in such cases need not be feared.

With electrodes platinized in the correct way it is possible to carry out accurate measurements in the range extending from solutions with a specific resistance of about 20 ohm cm to solutions with the highest specific resistance occurring (several hundred thousand ohms cm). With solutions having a specific resistance less than 20 ohm cm polarization still occurs to a slight extent; this polarization could only be avoided by using very much larger platinum electrodes and thereby making the cell very expensive, or by making the distance between the electrodes very great, so that in the first place the cell becomes awkward in shape, and in the second place the measurements of higher specific resistances can no longer be carried out accurately. It is, however, possible to take into account the influence of polarization, which is manifested in a decrease of the cell constant (10 - 15%), by calibrating the cell with solutions of very low specific resistance, as for example sulphuric acid of maximum conductivity (sp. resist. 1.3 ohm cm) or a saturated solution of sodium choride (sp. resist. 5 ohm cm). If the absolute value of the conductivity in this range is of less importance, relative measurements of great accuracy are always possible.

In the measurement of the conductivity of poorly conducting solutions it is necessary to take account of the conductivity of the water used. If the conductivity of the water is χ_{H_2O} and the measured conductivity of the solution $\chi_{sol.}$, then the amount contributed by the dissolved substances to the conductivity of the solution is:

$$\chi_w = \chi_{sol.} - \chi_{H_2O}\ .$$

Good distilled water has a specific resistance of about 200 000 ohm cm. With water distilled very carefully in the absence of carbon dioxide specific resistances of 10^6 ohm cm can be attained.

The influence of the temperature on the specific resistance of liquids is very great. In practically all aqueous salt solutions it becomes 2.0 - 2.5 per cent lower for every degree Centigrade increase of temperature. With most acids this value is 0.9 - 1.6 per cent, with alkali solutions 1.9 - 2.0 per cent. For accurate measurements an accurate measurement of the temperature within 0.2° C is necessary.

THE EXAMINATION OF THE MACROSTRUCTURE OF MATERIALS AND PRODUCTS WITH THE HELP OF X-RAYS. V

by J. E. DE GRAAF.

539.26

When a manufactured article, the product of various processes, is found in the end to possess flaws, these flaws may have arisen in the material and may be independent of the form of the article. Such flaws are most likely to appear oftenest when the material has attained its final form or position by way of the state. Examples of such cavities and enclosures have been given in this series in the discussion of the examination of cast and welded articles [1]). On the other hand the faults may have arisen during the process of giving the material a particular shape. The more trivial faults, such as for example due to the fact that the correct measurements have not been kept to in the process of turning, seldom offer material for an X-ray examination, because such faults are usually discovered in the inspection preceding delivery.

A structural process which, however, certainly calls for X-ray examination is rivetting which has also been treated in this series [2]). It was pointed out that defects in the crystal structure which may arise during these processes do not fall within the reach of absorption investigation.

Faults in assembly

In certain cases faults may arise without change in the material and independently of its form. These may be faults in the assembly, *i.e.* in the combination of perfect parts to a whole by soldering, cementing, fastening with bolts, connecting with wires, etc. In this case X-ray examination is again useful in order to sift out the faulty pieces at intermediate stages, and to trace the causes of faults in returned pieces. In addition X-ray investigation alone may make a certain process possible as a manufacturing method.

An example of the latter case is offered in the use of X-ray photography in the Philips industry in the manufacture of X-ray tubes and of certain transmitter valves. The construction of both of

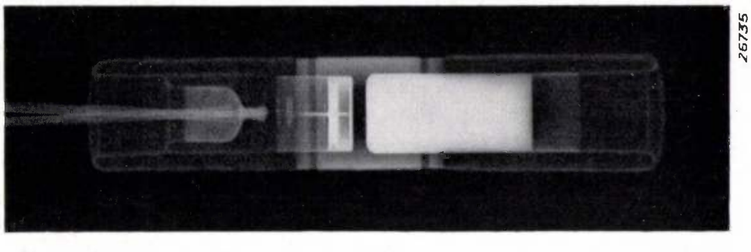

Fig. 1. Reduced X-ray photograph of an X-ray tube; the relative position and separation of the electrodes in the chrome-iron cylinder can only be seen in the X-ray photograph.

[1]) Philips techn. Rev. 2, 377, 1937; 3, 93, 1938.
[2]) Philips techn. Rev. 2, 350, 1937.

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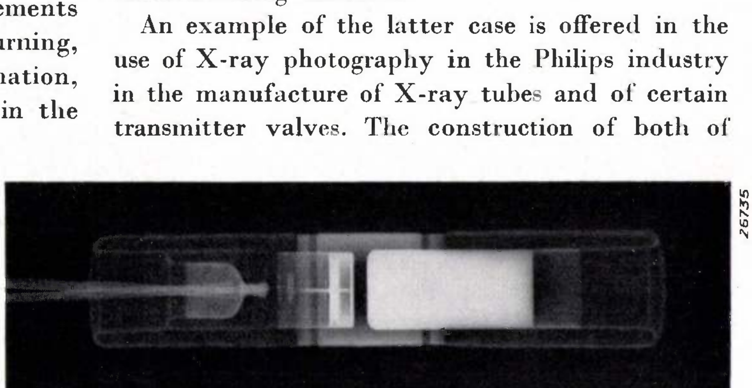

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[2]) Philips techn. Rev. 2, 350, 1937.

these is distinguished by the use of an optically opaque chrome-iron cylinder in which the vital parts of the tube are housed. These parts are then fused with glass to the cylinder in a joint which is proof against high vacuum. *Fig. 1* gives the reduced X-ray photograph of an X-ray tube. On the basis of this photograph it is possible to discover whether the electrodes are at the proper distance from each other. The same procedure is followed in the manufacture of transmitter valves, where in addition the X-ray photograph is used to discover any possible change in the position of the electrodes which might have come about during the shock test [3]).

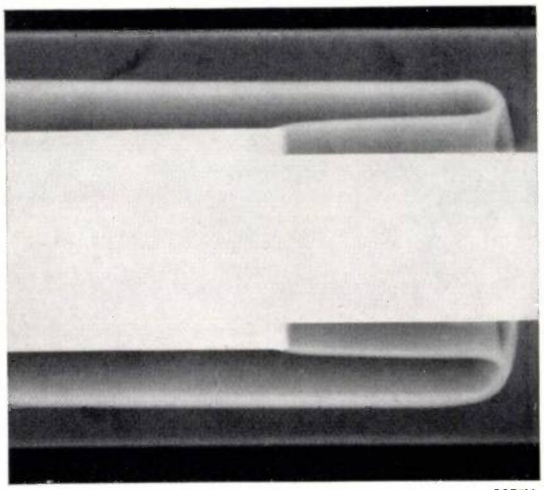

Fig. 2. Terminal assembled with bad cement for the connection of a high tension cable with an X-ray tube (gas bubbles and cracks in the cement).

Fig. 3. Breakdown in a terminal for the connection of a high tension cable with an X-ray tube. The cause (a gas bubble in the cement) and the track of the breakdown may be clearly seen (at →).

Another important sphere of application in assembly check tests is that of high voltage insulators and cables. No air spaces may occur between the parts of such insulators or cables, because with the high field strengths used, the air spaces by their ionization, would initiate breakdown. The components are therefore fastened together with a cement in which, however, gas bubbles are sometimes formed. *Fig. 2* shows the gas bubbles and even cracks in a terminal fastened with a quite inefficient cement. The terminal was intended for the connection between a high voltage cable and an X-ray tube. *Fig. 3* shows how the electrical breakdown of such a terminal appears in a photograph. The cause of the breakdown may here be seen (an air bubble in the cement), whereas if the cement had been melted out it would have disappeared entirely. With high tension insulators of artificial resin the X-ray photograph [4]) makes it possible to check the position of the pressed-in metal parts.

A third sphere of application lies in the examination of electrical apparatus. It may be a question of a simple cord in which a strand of wire is broken or of a complicated piece of apparatus which consists of many parts. In the filter shown in *fig. 4*, for example, a short circuit occurred which disappeared every time the wax was melted out, but reappeared again when it was poured in. By means of an X-ray photograph taken with the wax in place, the short circuit could be traced. Broken or

[3]) Philips techn. Rev. **2**, 115, 1937.

[4]) Philips techn. Rev. **1**, 257, 1936.

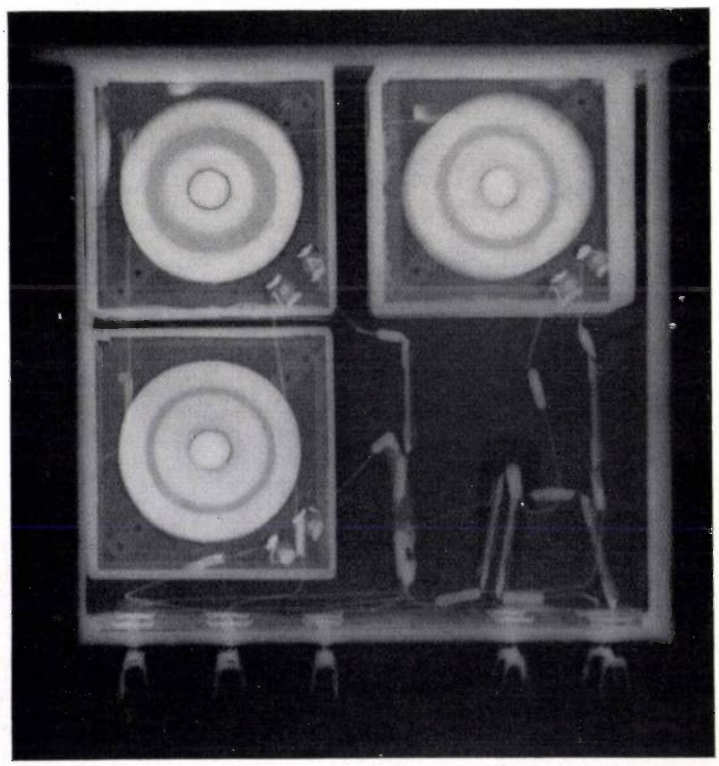

Fig. 4. Electrical filter. Short circuits, faulty soldering and air bubbles in the mass (lower right) can be observed in such a photograph.

badly soldered contacts also, as far as they are not obvious from the functioning of the apparatus, can often be found most simply in an X-ray photograph. How the correct placing of coils in a can may have suffered by the process of impregnation is shown in *fig. 5:* the primary coil of a transformer has sunk down out of the secondary coil so that the transformer no longer functions properly.

Finally assembly inspection with the help of X-rays is also applied to mechanical problems. In this case naturally it is the finer pieces of apparatus rather than large machines in which X-ray inspection is valuable. *Fig. 6* gives an example in the form of a carburetter in which a detail, such as the position of the petrol supply tubes, may be seen very clearly. The presence and position of stop pins, the tightness of nuts and the presence of self-locking nut washers are other examples of the application of X-ray examination in this branch of engineering.

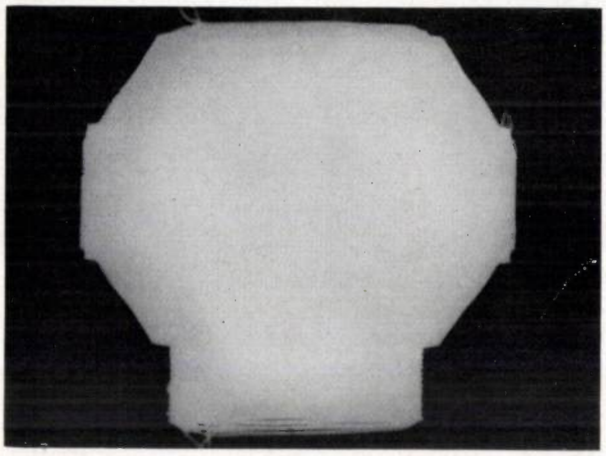

Fig. 5. The primary coil of a transformer impregnated when in its housing , has sunk down out of the secondary coil, as this X-ray photograph of the two coils in their relative position plainly shows. The iron core has been removed.

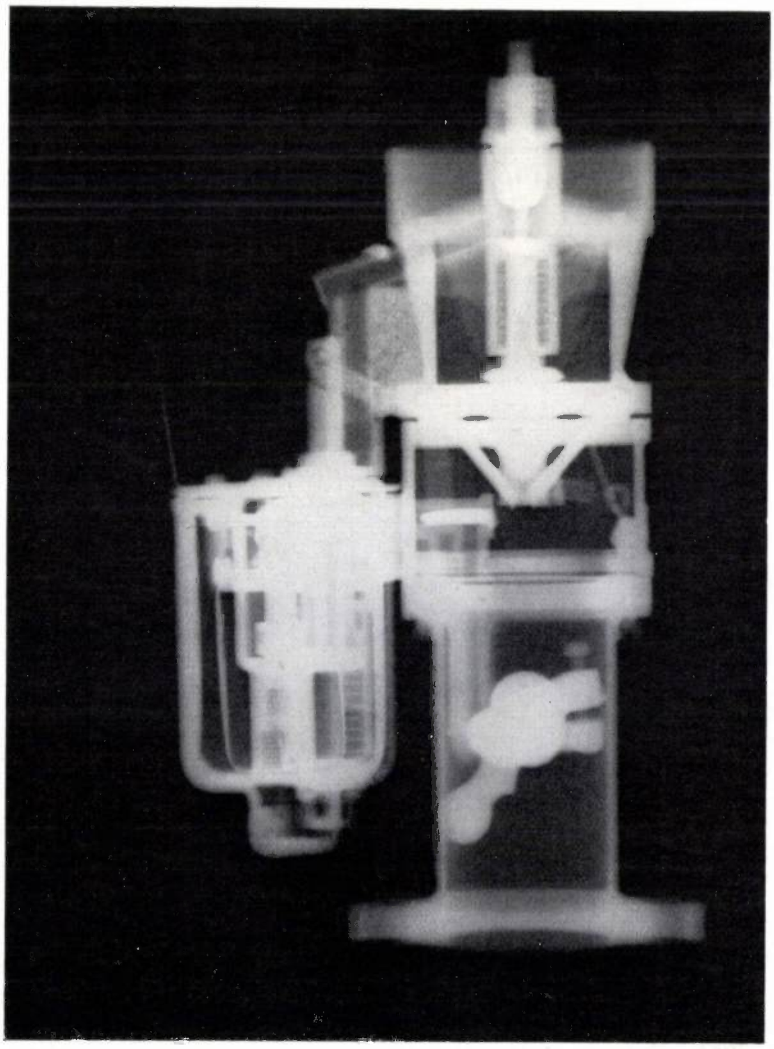

Fig. 6. X-ray photograph of a carburetter in which many details of the complicated structure may be plainly observed.

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DEALING WITH TECHNICAL PROBLEMS
RELATING TO THE PRODUCTS, PROCESSES AND INVESTIGATIONS OF
N.V. PHILIPS' GLOEILAMPENFABRIEKEN

EDITED BY THE RESEARCH LABORATORY OF N.V. PHILIPS' GLOEILAMPENFABRIEKEN, EINDHOVEN, HOLLAND

NOISE IN RECEIVING SETS

by M. ZIEGLER. 621.396.822

The noise in receiving sets in first treated in this article as an acoustic phenomenon. The chief sources of noise in radio sets are then dealt with: the first tuning circuits and the first amplifier valves. As suitable quantities for judging a radio set as to noise the concepts "noise voltage" and "specific noise voltage" are introduced. Measures are studied which may be taken to keep the noise level low, as well as the cases where the use of low-noise valves offers advantages. After several remarks on the permissible ratio of noise to signal voltage, the variation of the noise voltage is studied as a function of the signal strength, with special attention to the influence of automatic volume control on this variation.

Introduction

As a consequence of the thermal motion of electricity in conductors irregular voltage fluctuations occur between two points of an electric circuit. Fluctuations of the anode current about its mean value occur in amplifier valves, due to the corpuscular nature of electricity. It has already been shown in this periodical that these voltage and current fluctuations are entirely random and may be represented by a spectrum in which all frequencies are present in equal intensity. These alternating voltages and currents are very small and cannot be observed directly. When they are sufficiently amplified, however, for instance a million times, and passed through a loudspeaker they will give rise to irregular vibrations of the air; the impression which these vibrations make upon the air is called noise. More familiar examples of a similar sound are the rustling of leaves or the sound of gentle rain in a wood.

Noise as an acoustic phenomenon

As already mentioned all frequencies are represented in equal intensity in the above-mentioned electrical vibrations; this is, however, not the case with the air-vibrations. All frequencies are not amplified to the same extent, and the loudspeaker itself is not equally sensitive for all frequencies.

The sound impression which the irregular air vibrations make upon the air depends upon the distribution of the intensity of the components of the acoustic noise spectrum; the impression is for example clearly different according to whether the high frequencies are more or less strongly represented or the low, and it is plain that in such a case a distinction can be made between a high, sharp noise and a low, dull noise. If a filter is introduced between amplifier and loudspeaker which passes a certain narrow frequency band, the noise gives rather the impression of a definite tone which is clearer the narrower the band chosen.

The impression of loudness which the noise makes on the ear is also very dependent on the spectral distribution of the air vibrations. A high and a low noise, which represent the same vibration energy, will not make the same impression of loudness on the ear. Estimations of the loudness of the noise will moreover vary very much among individuals, since the curves for the sensitivity of the ear also vary very much among different individuals.

The reproduction of a signal, music or speech, is in principle always accompanied by the above-described noise, the intensity of which depends chiefly on the amplification necessary for the signal. When the noise is not very weak compared with the desired signal, it will form an audible, more or less disturbing background, it may even render the spoken word unintelligible and music disagreeable. It is obvious that every attempt should be made to keep this disturbance as slight as possible: in other words every attempt should be made to make the impression of loudness of the noise as slight as possible compared with that of the desired signal. We have just seen that the impression of loudness of the noise is influenced by its spectral distribution of intensity; therefore this distribution, *i.e.* the frequency characteristic of amplifier and loudspeaker together, will exert an influence on the audibility and disturbing effect of the noise. Since a detailed treatment of this subject would lead us too far afield, we shall limit ourselves to the most general considerations.

If we look for the most favourable frequency characteristic, it is obvious in the first place that the frequencies which never occur in the desired signal need not be amplified; the addition of new frequencies to the spectrum of the noise will, when the energy of the remainder remains unchanged, only be able to increase the noise impression, in that case without any advantage to the reproduction of the desired signal.

This principle can be further extended, and frequency ranges which are not of essential importance in the reproduction of music or speech may be omitted entirely. It is known that speech may be reproduced intelligibly with only the frequencies between 300 and 2 000 c/s, and music often retains its natural character when frequencies below 150 and above 3 000 c/s are cut off. If the high frequencies are cut off, the high sharp noise disappears, the energy of the speech or music is thereby often only slightly less, and one obtains the impression that the set has less background noise. Every owner of a receiving set can observe this when he reduces the amplification of the high frequencies with the help of the tone control in listening to a weak station.

This changing of the characteristic in order to make the impression of the ratio of signal to noise intensity more favourable is carried out at the expense of the quality of reproduction. The intensity relation between the different frequencies should of course preferably be exactly the same in the reproduction as it was in the original music or speech, so that the sensitivity of the whole transmitting and receiving installation, including amplifier and loud speaker, must be equally great for all frequencies, even for the frequencies (the highest and the lowest) which only occur to a smaller extent.

By weakening the high tones of a set which reproduces the high tones correctly (for instance by introducing a suitable filter passing only low tones, or, more simply, by placing a condenser across the terminals of the loudspeaker), the set will undoubtedly produce less noise, but this will be achieved at the expense of faithful reproduction. If one is particularly concerned with the noise, the set may be considered better in its changed conditions; there must therefore be a compromise between quality and absence of noise, or better, the quality (fidelity) of the set must be adapted to the circumstances: if the level of noise is so low that it is not heard, then the highest requirements of faithful reproduction may be demanded; if, however, the noise level is so high that it becomes annoying, it is desirable that the frequency spectrum of the reproduction be restricted to the most essential frequencies.

In order to decide which of two sets is more favourably constructed with respect to noise, it is obvious that it is not correct to compare directly the noise of a set which gives a good reproduction of the highest and the lowest frequencies with the noise of a set which only reproduces a limited range of frequencies. The noise of the first set will certainly become less when its frequency characteristic is made the same as that of the second, which can always be done by cutting off the highest and the lowest frequencies. Therefore, *in order to compare the noise of two sets we must make their reproduction characteristics the same:* then the differences in position and intensity of the sources of noise in sets of different types of construction are clearly shown. The influence of these differences, which are generally more important than the influence of the reproduction characteristic, will now be discussed.

Chief sources of noise

As in the previous articles in this periodical on this subject [1]), we shall not consider abnormal sources of noise, but confine ourselves to those sources whose presence is inherent in the amplifiers used, namely the thermal motion of electricity in electrical circuits and the shot effect of the anode current of amplifier valves. The influence of both effects can be represented by supposing that, in series with every resistance R, there is a fluctuating E M F, whose mean square $\overline{V^2}$ is proportional to the resistance, and that upon every anode current, a fluctuation is superimposed whose mean square is proportional to the mean anode current $\overline{I}$. For every frequency band $\Delta\nu$ the same contributions are obtained:

[1]) M. Ziegler: The causes of noise in amplifiers, Philips techn. Rev. 2, 136, 1937; Noise in amplifiers contributed by the valves, Philips techn. Rev. 2, 329, 1937.

$$\Delta \overline{V^2} = 4\, kT\, R\, \Delta\nu, \quad \cdots \cdots \quad (1)$$

$$\Delta\, \overline{(I_a - \overline{I}_a)^2} = F_a{}^2\, 2\, e\, \overline{I}_a\, \Delta\nu \quad \cdots \cdots \quad (2)$$

In an amplifier which consists of electrical circuits and amplifier valves there are therefore sources of noise at many points. All these fluctuations are entirely independent of each other, and the mean square of the total fluctuations is consequently equal to the sum of the squares of the contributions made by the individual sources of fluctuation. Not all of them will contribute the same amount to the noise issuing from the loudspeaker, since some noise voltages are amplified much more than others. In most cases, the effect of the thermal fluctuations in the elements of the circuit preceding the first amplifier valve and of the spontaneous anode current fluctuations of the first valve dominate over the others, because

The influence of the shot effect corresponds to that of an alternating voltage V_b on the grid of the first amplifier valve:

$$\Delta \overline{V_b{}^2} = F_a{}^2\, \frac{2\, e I_a}{S^2}\, \Delta\nu, \quad \cdots \cdots \quad (4)$$

where S represents the slope of the valve [3]).

The square of the total voltage on the grid of the first valve with which we are here concerned is the sum of the two contributions above:

$$\Delta \overline{V^2} = \Delta \overline{V_k{}^2} + \Delta \overline{V_b{}^2} = \left(4\, kT R_k + F_a{}^2\, \frac{2\, e I_a}{S^2}\right) \Delta\nu\,. \quad (5)$$

Instead of the anode current I_a, the slope S and the noise factor $F_a{}^2$, one may use in equation (5) the noise resistance R_b, a quantity which has already been explained in this periodical [3]), as the value of the resistance between grid and cathode

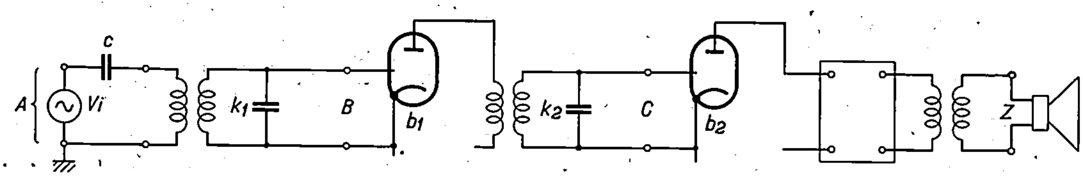

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Fig. 1. Position of the sources of noise in a receiving set. The principal sources of noise are: first circuit k_1 and first amplifier valve b_1, second circuit k_2 and second amplifier valve b_2. The EMF induced in the receiver V_i acts in A. C represents the capacity of the aerial. The box represents the part of the apparatus not drawn. V_i causes the voltages $V_i\, \gamma^{AB}$, $V_i\, \gamma^{AC}$ and $V_i\, \gamma^{AZ}$ in B, C and Z respectively.

these fluctuations are amplified by the greatest number of valves. Suppose for example that the first valve amplifies ten times, then the contribution of a noise voltage between the grid and the cathode of the first valve to the total noise energy is 10^2 times as great as the contribution of an equally high voltage on the anode of this amplifier valve. The contribution of the latter voltage may then be neglected. In the case of a receiving set constructed according to the general details of *fig. 1* account need only be taken of the fluctuations of electricity in the first circuit and the shot effect in the first amplifier valve. The former give rise to a voltage on the tuned circuit which, for a narrow frequency range on either side of the resonant frequency, is given by:

$$\Delta \overline{V_k{}^2} = 4\, kT\, R_k\, \Delta\nu, \quad \cdots \cdots \quad (3)$$

in which R_k represents the parallel resistance of the tuned circuit [2]).

[2]) For a narrow range of frequencies on both sides of the resonant frequency the circuit behaves like a resistance R_k so that the formula (3) may be applied directly. Exactly the same result is attained by beginning with the fluctuating EMF of the coil resistance and calculating by means of Kirchhoff's laws the voltage thereby occurring across the terminals of the condenser.

of an amplifier valve whose thermal fluctuating EMF contributes to the noise an amount equal to that contributed by the valve itself. Equation (5) then has the form:

$$\Delta \overline{V^2} = 4\, kT\, (R_k + R_b)\, \Delta\nu \quad \cdots \quad (6)$$

Ratio of signal to noise

In order to study the disturbing effect of the above discussed sources of noise at a certain strength of the desired signal, we should have to compare the sound impressions of disturbance and signal from the position of the observer. The practical conclusions with respect to the construction of the set will, however, as we shall see later, be fundamentally the same whether we take into account the frequency characteristic of the loudspeaker and the curve for the sensitivity of the ear, or whether confine ourselves to a comparison of the electrical energies of noise and signal which are conducted to the coil of the loudspeaker.

Let us imagine that the receiver is tuned to a carrier wave of the frequency ν which induces a voltage V_i in the aerial. This carrier wave is

[3]) Philips techn. Rev. 2, 329, 1937.

modulated sinusoidally: depth of modulation m, frequency of modulation s. This modulation may, as is known, be represented by the presence of two side bands with the frequencies $\nu + s$, and $\nu - s$, and the amplitude $^1/_2\ mV_i$. These side bands are amplified together with the carrier wave. If superheterodyning is employed, the frequencies of carrier wave and side bands become larger or smaller by the same amount. After detection and low-frequency amplification, as a result of the collaboration of the carrier wave and one of the side bands, a sinusoidal alternating voltage with the frequency s and the amplitude

$$^1/_2\ m\ V_i\ \gamma_s{}^{AZ}.$$

occurs at the output terminals. In the expression for the amplitude $\gamma_s{}^{AZ}$ is a factor which we may call the amplification between the points A and Z of fig. 1. This amplification depends on the modulation frequency (signified by the subscript s) and will in general depend also on the strength of the carrier wave.

The low frequency oscillations from the two side bands are generally in phase, and we may assume that the amplification is the same for both. The total amplitude of the signal voltage is then exactly double the above amplitude, namely:

$$V_s = m\ V_i\ \gamma_s{}^{AZ} \cdot \cdot \cdot \cdot \cdot \cdot \quad (7)$$

The voltage fluctuations of the noise are amplified in exactly the same way as the side bands of the carrier wave received: they may also be considered as a type of modulation of the carrier wave. In contrast to the side bands, however, the low frequency voltage fluctuations, from the frequencies $\nu + f$ and $\nu - f$ in the fluctuations, are not in phase, but have an arbitrary phase relation. The mean square of the fluctuating voltage at the output side is therefore equal to the sum of the mean squares of the amplified voltage fluctuations for all frequency intervals on both sides of the carrier wave, and is given by:

$$\overline{V^2} = \Sigma\, 4\, kT\ (R_k + R_b)\ (\gamma_f{}^{BZ})^2\ \Delta f \cdot \cdot \quad (8)$$

γ^{BZ} represents the amplification between B and Z in fig. 1 [4]).

If we now assume, as in the case of the sinusoidal modulation, that the amplification is the same for the frequency $\nu - f$ as for the frequency $\nu + f$, we may replace the summation over all frequencies by twice the integral taken from 0 to ∞. The result may now be written in the form:

$$\overline{V^2} = 8\, kT\ (R_k + R_b)\ (\gamma_s{}^{BZ})^2 \int_0^\infty \left(\frac{\gamma_f{}^{BZ}}{\gamma_s{}^{BZ}}\right)^2 \mathrm{d}f \cdot \quad (9)$$

The integral in equation (9) has the dimensions of a frequency; it is connected with the "fidelity" of the receiver and is indicated by the letter G.

If we wish to take into account the frequency characteristic of the loudspeaker and the curve for the sensitivity of the ear, these two factors may be included in the amplification $\gamma_f{}^{BZ}$. The value of G may be slightly changed thereby, but there is no other change in the above calculation.

By dividing (9) by (7) we obtain finally the ratio of the electrical energy of noise to that of the signal:

$$\frac{\overline{V^2}}{V_s{}^2} = \frac{8\, kT\ (R_k + R_b)}{m^2\ V_i{}^2} \left(\frac{\gamma_s{}^{BZ}}{\gamma_s{}^{AZ}}\right)^2 \cdot G = \frac{8\, kT\, (R_k + R_b)}{(\gamma_s{}^{AB})^2\, m^2\ V_i^2}\ G, \cdot \cdot \cdot \quad (10)$$

in which γ^{AB} represents the amplification between the aerial and the source of fluctuation (the grid of the first amplifier valve). The ratio of noise to signal evidently does not depend upon the magnitude of the amplification γ^{BZ}. For the random fluctuations which cause the noise, therefore, just as for every other form of interference, it is true that their ratio to the desired signal remains the same no matter what the amplification applied to signal and interference.

Noise voltage

We have seen that at the output side of the apparatus the ratio of the signal to the noise due to the first valve and the first circuit does not depend upon the amplification γ^{BZ}. Besides the input signal V_i and its depth of modulation only the square of the following quantity occurs:

$$\sqrt{8\, kT\, (R_k + R_b)\, G}/(\gamma^{AB}).$$

The above quantity has a simple meaning: it is exactly equal to the R.M.S. value of the fluctuating voltage of the first valve and the first circuit divided by the amplification between the aerial and the grid of the first valve, and therefore indicates the magnitude of a fluctuating E M F which, acting at the same place as the signal to be amplified, would cause noise equal to the contribution of the first circuit and first valve. We call this value the noise voltage V_r of the set.

4) γ^{BZ}, like γ^{AZ}, depends upon the value of the input signal. If the signal is very weak the detector still works in the quadratic part of its characteristic, and an increase in V_i then improves the detection, so that γ^{BZ} increases. If V_i increases still more the detection will finally cease to improve, and, due to the automatic volume control which usually commences to operate then, the amplification will decrease. If a very weak signal is fed to a receiving set, and then allowed to increase, an increase of the noise will at first be observed followed by a decrease.

Until now only the voltage fluctuations in the first valve and first circuit have been taken into account. In the following we shall also take into account the contribution of the second valve and the second circuit (at C in fig. 1), and we thus obtain for the noise voltage the following expression:

$$V_r^2 = 8\,kT\left[\frac{R_{k1}+R_{b1}}{(\gamma_s^{AB})^2}\,G_1 + \frac{R_{k2}+R_{b2}}{(\gamma_s^{AC})^2}\,G_2\right] \quad (11)$$

In general the amplification between B and C will be independent of the frequency in the range of frequencies which contribute to G.

Then $G_1 = G_2 = G$.

Further by taking into account that:

$$\gamma^{AC} = \gamma^{AB}\,\gamma^{BC},$$

equation (11) may be simplified to:

$$V_r^2 = 8\,kT\,G\,\frac{\left[R_{k1}+R_{b1}+\dfrac{R_{k2}+R_{b2}}{(\gamma_s^{BC})^2}\right]}{(\gamma_s^{AB})^2} \quad (12)$$

The expression within brackets has the dimensions of a resistance and for the sake of simplicity it will be represented by $\mathfrak{R}$.

Experimental determination of the (specific) noise voltage of a receiving set

The practical importance of the noise voltage or specific noise voltage will be made clearer by means of a discussion of the way in which it is determined experimentally.

In the arrangement of *fig. 2* aerial and earth terminals of the set to be investigated are connected *via* an artificial aerial with a signal generator which produces a sinusoidal voltage of suitable amplitude and of a frequency to which the set is tuned.

Instead of the loud speaker coil there is a square law meter of equal impedance. Due to noise fluctuations the meter will show a deflection which, according to equation (9) with the simplifications introduced subsequently, corresponds to the square of a voltage of the value:

$$\overline{V^2} = 8\,kT\,G\,\mathfrak{R}\,(\gamma_s^{BZ})^2.$$

If the carrier wave is now modulated sinusoidally with a signal of frequency s, the deflection is increased by the added tone energy:

$$V_s^2 = (m\,V_i\,\gamma_s^{AZ})^2.$$

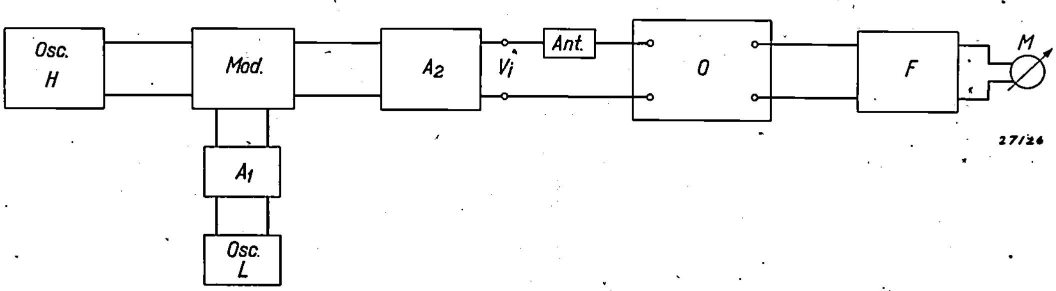

Fig. 2. Diagram of the layout for measuring the specific noise voltage of receiving sets. In the modulator the high-frequency voltage of the oscillator H is modulated by means of a 500 cycles alternating voltage from the oscillator L whose amplitude can be adjusted at any value with the attenuator A_1. The modulated voltage is reduced with A_2 to the desired value V_i and is conducted to the aerial terminals of the receiving set over the artificial aerial. The square law meter M is connected to the receiver *via* a bandpass filter F (passing frequencies from 300 - 1 300 c/s), and indicates the electrical energy of the modulation tone and of the noise within the band passed.

Specific noise voltage

At the beginning of this article it was explained that we shall concern ourselves chiefly with the difference in noise which may still occur in different sets after the frequency characteristics of the reproduction have been made the same. This means that the noise voltages refer to the same value of G. We shall always use for G the value 1 000 sec^{-1} and call the noise voltage corresponding to this value the specific noise voltage V_{rsp}.

From equation (12) it follows that:

$$\begin{aligned} V_{rsp}^2 &= 8000\,kT\,\mathfrak{R}/(\gamma_s^{AB})^2 \\ &= 0.322\cdot 10^{-4}\,\mathfrak{R}/(\gamma_s^{AB})^2\,(\mu\mathrm{V})^2 \quad (13) \end{aligned}$$

The depth of modulation can now be so adjusted that the deflection due to noise and tone energy together is just double the deflection due to noise alone. Then the tone energy is equal to the noise energy. The depth of modulation at which this occurs is called the noise modulation depth m_r. From this definition it follows that:

$$m_r\,V_i\,\gamma_s^{AZ} = \sqrt{8\,kTRG}\;\gamma_s^{BZ},$$

$$m_r\,V_i = \sqrt{8\,kTRG}/\gamma_s^{AB} = V_r.$$

We see from this that the noise voltage is equal to the product of the aerial signal and the noise modulation depth. Measurement of the noise voltage thus becomes the determination of the noise modulation depth at a known aerial signal.

In order to obtain the specific noise voltage, the noise voltage must be multiplied by $1\,000/G$. The determination of G which is necessary for this is rather elaborate. A simpler method consists in the use of a bandpass filter which passes only those frequencies between 300 and 1 300 c/s, and which is connected between the output terminals of the set and the quadratic meter. Since the fidelity curves of receiving sets are generally fairly flat in this frequency range, the value of G of the whole setup will be 1 000 c/s in sufficiently close approximation.

An additional advantage in the use of such an output filter is that the 50 cycles alternating voltage and several of its harmonics are eliminated. These fluctuations from the mains may otherwise easily lead to a disturbance of the noise measurements in alternating current receivers.

The influence of amplifier valves and oscillatory circuits on the noise of a set

The attempt to obtain as low a noise voltage as possible leads, according to equation (13) to two requirements:

1. The gain γ^{AB}, *i.e.* the voltage amplification between the aerial and the grid of the first valve should be as great as possible,
2. The total noise resistance:

$$\mathfrak{R} = R_{k1} + R_{b1} + \frac{R_{k2} + R_{b2}}{(\gamma_s{}^{BC})^2} + \dots$$

must be as small as possible.

A large value of γ^{AB} is obtained by having a closely coupled aerial circuit with a sharp resonance curve. This requirement of a large gain is also necessary for the sensitivity of the set, but we shall not discuss this point here.

The value of the noise resistance at a given amplification depends upon the noise resistances of the amplifier valves and the parallel resistances of the circuits used. If definite requirements are made of amplification, selectivity, etc. then the characteristics of the circuits, including R_{k_1}, R_{k_2}, etc. are thereby determined; with equal amplification, however, the contributions of the various amplifier valves to the noise voltage may be very different.

Previously in this periodical (Ziegler, *loc.cit*) it was derived that the noise resistance of an amplifier valve is given by:

$$R_b = 20\,000\ F_a{}^2\ \frac{I_a}{S^2}\ \text{ohm}, \quad \cdots \quad (14)$$

where I_a is the anode direct current in mA and S the slope in mA/volt. The factor $F_a{}^2$ has a value which varies between 0.05 and 1.

In the following table may be found the values calculated for the noise resistance of the octode EK 2 (oscillator frequency-changing valve), of the high-frequency pentode EF 5 and of the "silentode" EF 8.

	I_a	S	$F_a{}^2$	R_b
	mA	mA/volt	—	Ω
EK 2	1.5	0.5	0.67	80 000
EF 5	8	1.7	0.27	15 500
EF 8	8	1.8	0.062	3 200

When the circuit resistances are high compared with the noise resistances of the valves, the replacement of these valves by valves of a type having lower noise resistance will give much less improvement than when $\mathfrak{R}$ is mainly determined by the valves.

For the sake of illustrating the various ratios occurring in practice, the specific noise voltage will be calculated for three sets, on the medium wave at 1 000 kc/s when the circuit resistance may be 100 000 ohms, and on the shortwave at 10 000 kc/s where 8 000 ohms is a normal value for R_k.

The first set has an octode EK 2 as first valve, the other two begin with a stage of high-frequency amplification preceding the frequency-changing valve. The valves used are the pentode EF 5 and the silentode EF 8, respectively.

Additional data of importance for our considerations will be found in the above table, where the calculated values of the noise energy in the different cases are given.

With the assumed value of γ^{BC} the second valve and the second circuit do not have an influence on the noise voltage more than a few per cent.

It may be seen from the table that on short waves where the circuit impedance is comparitively small, the application of a high-frequency amplifier stage preceding the frequency-changing valve, and particularly the use of a silentode (set C) gives an important decrease in the noise energy, which is essential for good reception on short waves. On the medium waves the contribution of the noise resistance of the amplifier valves to the total noise appears from the table to be comparitively small, so that in the case in question the improvement obtained by the use of high-frequency amplification is not great.

The contribution of the amplifier valves becomes much greater, however, if the valves are used at a lower anode current. This may easily be seen from formula (14): the noise factor F_a does not change much, but S^2 decreases much more rapidly than I_a, so that R_b increases. On decreasing the slope of the first valve by a factor of ten the noise resistance of the octode EK 2 used in set A is about 800 000 ohms and the noise resistance of the silentode used in set C amounts to 30 000 ohms. The noise voltage of the first stage of set A will then, at 1 000 kc/s be 2.24 times as much, while that of the first stage of set C only increases by 14 per cent. It may therefore be seen that the advantage of amplifier valves with low noise resistance can be realized even at high values of the circuit resistance. because the anode current does not decrease as rapidly as the square of the slope (see equation 14)). In the second place the amplification γ^{BC} decreases so that in equation (12) the contribution of $(R_{k_2} + R_{b_2})/(\gamma^{BC})^2$ to the noise resistance increases. Both of these changes result in the fact that the noise voltage increases when the automatic volume control begins to work upon increase in the input signal. On the other hand the signal strength also increases, so that the ratio of signal to noise will in general be more satisfactory. Since the noise voltage is not constant, the quality of a receiving set with respect to noise cannot be determined by a single measurement of the noise voltage. One should investigate how the noise

Calculated values of the specific noise voltages of three sets at 1000 kc/s

No.	R_{k_1}	first valve	R_{b_1}	γ^{BC}	R_{k_2}	second valve	R_{b_2}	R	γ^{AB}	V_{rsp}	Noise *) energy
	Ω		Ω		Ω		Ω	Ω		μV	
A	100 000	EK 2	80 000	50	100 000	EF 5	15 000	180 046	5	0.48	1.72
B	100 000	EF 5	15 000	10	100 000	EK 2	80 000	116 800	5	0.39	1.14
C	100 000	EF 8	3 000	10	100 000	EK 2	80 000	103 800	5	0.366	1.00

Calculated values of the specific noise voltage at 10,000 kc/s

No.	R_{k_1}	first valve	R_{b_1}	γ^{BC}	R_{k_2}	second valve	R_{b_2}	R	γ^{AB}	V_{rsp}	Noise *) energy
A	8 000	EK 2	80 000	50	8 000	EF 5	15 000	88 092	1.5	1.12	7.38
B	8 000	EF 5	15 000	10	8 000	EK 2	80 000	23 880	1.5	0.585	2.00
C	8 000	EF 8	3 000	10	8 000	EK 2	80 000	11 880	1.5	0.413	1.00

*) The noise energy of set C has been chosen as the unit of noise energy.

The variation of the noise resistance as a function of the aerial signal. Noise graphs.

If the adjustment of the amplifier valves to maximum steepness of slope, for which the above values of the noise voltage have been calculated, was retained for every value of the aerial signal, the noise voltage would be constant and independent of the value of the signal.

In modern receiving sets the adjustment of the amplifier tubes is, however, influenced by the strength of the signal; when it becomes greater the negative bias of the control grids increases automatically, and the amplification becomes less, so that the output changes only little, and overloading with the accompanying interferences and distortions is avoided.

The increase of the negative grid voltage has two results with regard to the noise voltage. In the first place the noise resistances of the amplifier tubes increase very much, as we have already noted, voltage changes with the amplitude of the aerial signal. The results of measurements on several types of apparatus are given in the noise graphs of *fig. 3* where the noise voltage is plotted against the aerial signal on a logarithmic scale. The continuous lines refer to an apparatus with high-frequency amplification and a silentode, and are lower than the dotted curves which refer to an apparatus with an octode as first valve. At low signal strengths the curves are horizontal, since here the automatic volume control is not yet in operation or has no appreciable effect; at high signal strength the curves have an upward trend.

Permissible ratio between noise and signal

In order to draw practical conclusions from the noise voltage calculated and measured it is important to know what is the permissible ratio between noise and signal. According to formulae (10) and (12) this ratio, measured electrically, is equal to $V_r^2/m^2V_i^2$. The disturbing impression of the noise will therefore depend mainly upon the relation between the noise voltage V_r and the modulation voltage mV_i. Further the frequency

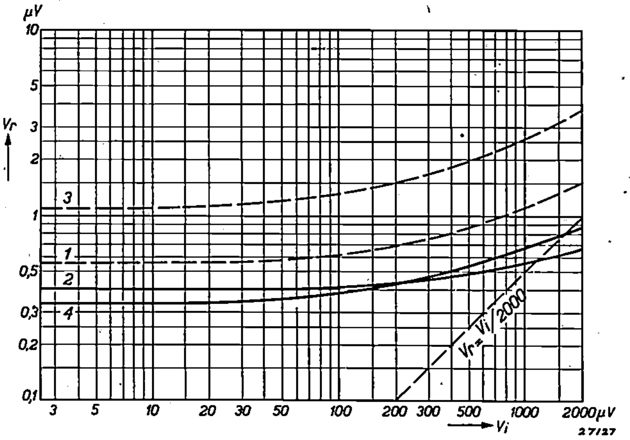

Fig. 3. Noise graphs. The specific noise voltage is plotted as a function of the aerial signal for two characteristic sets and at two frequencies.

Curve 1: first valve a frequency-changer; frequency 1 000 kc/s
Curve 2: first valve a silentode; frequency 1 000 kc/s
Curve 3: first valve a frequency-changer; frequency 10 000 kc/s
Curve 4: first valve a silentode; frequency 10 000 kc/s.

The straight line at 45° is the line $V_r = V_i/2\,000$. For the values of V_i to the right of the intersection of this straight line with the curves of the noise voltages, the noise is practically inaudible.

characteristic and the reproduction will be important because they affect the character of the noise. The absolute value of the noise energy also plays a part of course in connection with the threshold sensitivity of the ear.

The question as to the permissible value of V_r/mV_i is not precise enough in its simple form: whether a value is permissible or not depends upon many circumstances and considerations; in the first place it depends upon the nature and strength of interferences other than background noise. A more precise formulation of the question is to ask the ratio between noise and signal at which, in the absence of other disturbances, the noise is so nearly inaudible that there is no necessity to diminish it further.

For a given apparatus (Philips 695A) the rather divergent data of a number of observers have given as an average result that with the level of reproduction normal for an average room the above requirement was fulfilled when:

$$\frac{V_{rsp}^2}{(mV_i)^2} = 2.8\cdot10^{-6} \quad . \; . \; . \; . \quad (15)$$

The specific fluctuation energy should therefore be 55.5 decibels lower than the energy of modulation. Since this result only establishes an order of magnitude it may also be used for rough estimations when one is concerned with sets having a different reproduction characteristic.

From (15) it follows that:

$$\frac{V_{rsp}}{mV_i} = \frac{m_{rsp}}{m} = \frac{1}{6}\,\%.$$

The specific noise modulation may not be more than $^1/_6$ per cent of the average signal modulation.

If we base our calculations on an average signal modulation of 30 per cent [5]) we find that for noise-free reception the specific noise voltage may not be more than 1/2 000 of the signal received.

If this rule is applied to the three sets studied in the previous section, and if one assumes that the noise voltage does not increase due to automatic volume control, it is found that on 1 000 kc/s and 10 000 kc/s the signal strengths in the table below are a necessary minimum for noise-free reception.

No.	Carrier wave 1000 kc/sec	Carrier wave 10.000 kc/sec
A	0.96 mV	2.24 mV
B	0.78 mV	1.17 mV
C	0.73 mV	0.83 mV

In the graph of fig. 3 the straight line $V_r = 1/2\,000\ V_i$ is drawn as a continuous line. The points of intersection of this straight line with the curves for the noise voltage indicate for these practical cases the minimum strengths of aerial signal necessary for noise-free reception.

[5]) It should be noted here that the depth of modulation of good transmitters is often much less than 30 per cent, since this improves the quality of broadcasting, transmission and reception. For a modulation depth of 10 per cent, noise-free reception is only reached at $V_i = 6\,000\ V_r$.

EMISSION OF LIGHT IN THE POSITIVE COLUMN AT LOW PRESSURE

by W. UYTERHOEVEN.

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As an introduction the phenomenon of self-absorption is briefly discussed with its influence on the occurrence of excitation and ionization by steps, as well as the Doppler effect in the emission of light. The light emission of the positive column in neon, sodium vapour and mercury vapour at low pressure is dealt with, and the conditions for maximum efficiency are considered. Finally several peculiarities of low pressure columns in alternating current discharges are studied

In the following we shall consider in detail the most striking property of the positive column, its *emission of light*. We shall confine ourselves to discharges in sodium vapour, neon and mercury vapour at low pressure which are used technically for illumination: the first in sodium lamps and the other two in red and blue illuminated advertising signs.

It has already been mentioned in a previous article [1]) that the excitation of atoms by electrons, which is the cause of light emission, is only a byproduct of the ionization necessary for obtaining the charged particles. It is obvious that we wish to produce this important "by-product" with the greatest possible efficiency. The conditions for this efficiency are dependent on the nature of the gas or vapour atoms which are used for the transformation of electrical into light energy. In the case of sodium a low pressure is used for the discharge, with mercury a high pressure is an advantage.

In studying the light emission of the positive column we repeatedly encounter the phenomenon of self-absorption, which we shall first discuss briefly.

Self-absorption of resonant radiation

By the term self-absorption we mean the property of a gas (or vapour) of being able to absorb the radiation which it naturally emits. Let us suppose that a sodium atom is brought from the zero state to the 2.1 V state by electron excitation (cf *fig. 8*). After an average of about 10^{-8} s the excited atom gives up the energy absorbed in the form of yellow sodium light and returns to the zero state. If this radiating atom were alone in the tube, or among atoms of a rare gas, the radiation would immediately leave the tube (*fig. 1a*). Actually, however, such a radiating atom is always surrounded by a large number of non-excited sodium atoms. These latter atoms easily absorb the radiation of wave lengths 5 890 and 5 896 Å and become excited atoms in the 2.1 V state [2]). It then takes an average of 10^{-8} s again before the newly excited atom radiates its energy and absorption again occurs by still other non-excited atoms. This process of emission and reabsorption is repeated many times before the radiation leaves the tube in the case of lamps which emit resonant radiation (*fig. 1b*). For sodium

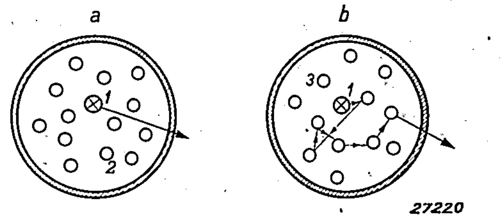

Fig. 1. Diagrammatic representation of self-absorption: *1* = excited sodium atom, *2* = rare gas atoms, *3* = normal, non-excited sodium atoms. In case *a*) the radiation leaves the tube directly, in case *b*) it is absorbed repeatedly before it leaves the tube.

lamps this number may be estimated at 10 000 under certain conditions. In this case therefore it is not 10^{-8} s before the radiation excited is passed on to the surroundings outside the tube, but $10^4 \times 10^{-8} \doteq 10^{-4}$ s, so that there is an apparent lengthening of the life of the excited atoms. We shall see later from the spectra that for an economical production of sodium light self-absorption should be avoided as much as possible, while for the production of mercury light a high self-absorption is favourable. It is clear that the chance of excitation in steps is very much increased by the phenomenon of self-absorption. If both of these processes are to be promoted, self-absorption may on the one hand be increased by increasing the pressure, that is by increasing the number of absorbing atoms, and on the other hand the number of collisions between electrons and excited atoms may be increased by increasing the current density, *i.e.* by increasing the concentration of electrons. Both methods are used simultaneously in the high-pressure mercury discharge.

Upon further study the phenomenon of self-

[1]) Philips techn. Rev. 3, 161, 1938.

[2]) The 2.1 volt state consists of two states differing very slightly in energy content, called a doublet.

absorption is found to be very complex, due in part to the width of the spectral lines. With precise instruments it is found that a spectral line is never absolutely monochromatic, but always has a certain width, *i.e.* the light is not radiated in a single wave length, for example 5 889.963 Å, but in a region between 5 889.988 Å and 5 889.938 Å.

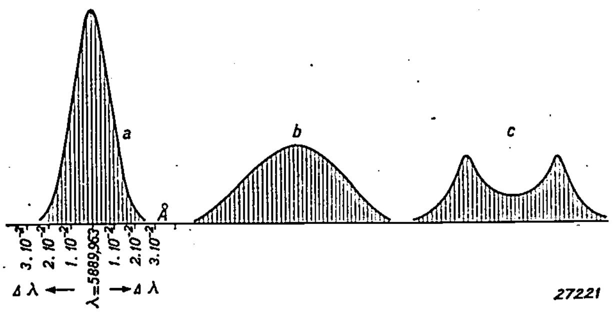

Fig. 2. *a*) Doppler broadening of the sodium line 5890 Å at 277° C (550° K). *b*) Form of the spectral line after the yellow beam has passed through an atmosphere of sodium vapour at lower temperature or in ionized condition.

One of the most important factors which causes this broadness of the lines is the Doppler effect which also appears in the propagation of sound. In the case of a source of sound, a tuning fork for instance, with a given frequency, the tone observed becomes higher when the source is moving toward the observer, and it becomes lower when it moves away from the observer. This phenomenon may be observed in the case of the whistle of a fast locomotive passing an observer: at the moment the locomotive passes the observer, the tone of the whistle suddenly becomes lower. In the positive column the spectral line which is observed is due to the simultaneous radiation of a large number of excited atoms. These atoms take part in the random thermal motion of the atoms, so that upon radiation a number of them are moving toward the observer and an equal number in the opposite direction. The thermal velocities of the atoms occurring in a low pressure column discharge are of the order of several hundred metres per sec. Although these velocities are small compared with the velocity of light ($c = 3 \times 10^8$ m/s), nevertheless they are sufficiently high to cause a measurable change in the frequency observed (comparable to height of a tone) and in the wave length. The resulting distribution of intensities as a function of the wave length due to this Doppler broadening may be seen in *fig. 2a* for sodium vapour at a temperature of 277° C (550° K).

The difficulties in the calculation of self-absorption arise because of the fact that the frequencies at the middle of the line are more easily absorbed than those at the edges. This changes the form of the line. If a beam of sodium light whose spectral line has the form *a*) of fig. 2 is sent through sodium vapour which, like the radiating vapour, has a temperature of 550° K, the spectral line after emergence will have become broader and have the form of *fig. 2b*. If on the other hand the beam passes through sodium vapour at a lower temperature, or through partially ionized sodium vapour in which cumulative effects may occur due to the presence of electrons, the line takes on the form of *fig. 2c*: the middle frequencies have been absorbed.

Finally we would point out that when there is sufficient self-absorption in a gas Lambert's law holds, according to which the brightness of a source of light is independent of the direction of observation. When strong self-absorption occurs the thickness of the layer through which the light passes when it leaves the tube directly is very small, so that the depth occupied by the discharge behind the wall has no influence. For a sodium lamp Lambert's law is approximately valid, this is, however, not so with a neon or a mercury lamp although self-absorption of the visible spectral lines also takes place in the two latter cases to a somewhat smaller extent since the absorbing atom is already in an excited state (*v. fig. 3*, and *fig. 13*). The number of excited atoms in the discharge is, however, always much smaller than the number of neutral non-excited atoms.

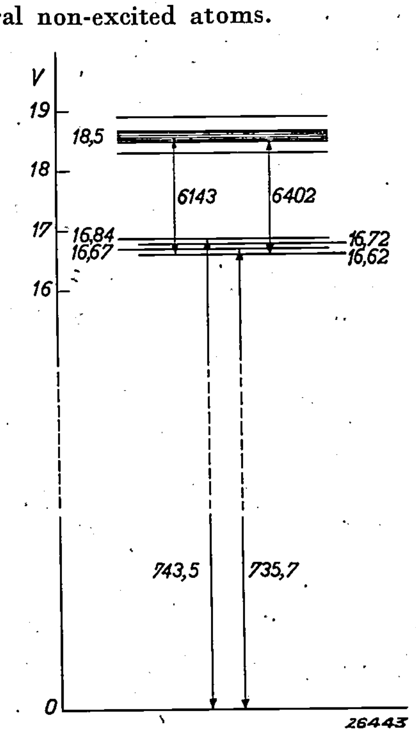

Fig. 3. Schematic representation of the stationary states of the neon atom which are important in the light emission of the positive column.

Neon column discharge

The complete spectrum of the neon atom is very complicated, but the portion of it which is important in the light emission of the positive column is given schematically in fig. 3 with the excited states. The four levels from 16.62 to 16.84 V we shall indicate collectively as the 16.6 V state; the higher levels in the neighbourhood of 18.5 V, which differ from each other by fractions of a volt only, we shall refer to collectively as the tenfold 18.6 V state. The visible lines are emitted here upon transitions from the 18.6 V group to the 16.6 V group. Several of the strongest lines are indicated, the accompanying number gives the wave length in Å. The resonance lines, which correspond to transitions from two of the 16.6 V levels to the zero state lie in the far ultraviolet at 744 and 736 Å because of the great difference in energy. The other two 16.6 V levels are both "metastable": The transition from these to the zero state is very improbable, so that it takes a much longer time before the energy is given off as radiation. Metastable states are usually destroyed by collisions either with the wall or with other atoms and electrons, while their life is of the order of 10^{-4} s.

Since two of the 16.6 V states are starting levels for the resonance lines, and the other two are metastable, it may be expected that the influence of cumulative effects will be strong in neon. On the other hand experience shows that the 16.6 V levels are very rarely excited by electrons. We therefore have the two following possibilities for light emission in the neon column:

a) The atom is excited from the zero state into one of the 18.6 V levels, passes over into one of the 16.6 V levels with emission of visible light and then returns to the zero state (*fig. 4a*).

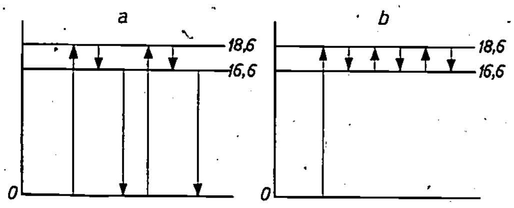

Fig. 4. Excitation of transitions in the neon spectrum which produce visible lines. *a*) from the zero state, *b*) from the 16.6 volt states.

Of the energy supplied, therefore, 18.6—16.6 = 2 V is radiated in the visible region. With the help of the curve for the sensitivity of the eye and the relative energy distribution in the neon spectrum one may calculate how much light is radiated when the energy of the visible neon lines taken together amounts to one watt; the result is found to be 137 lm. Therefore $\frac{2}{18.6} \times 100 = 10.8$ W are given off as visible light in the column per 100 W supplied. With the given energy distribution this gives $137 \times 10.8 = 1480$ lm, so that the "efficiency" of the column would be 1 480/100 = 14.8 lm/W [3]).

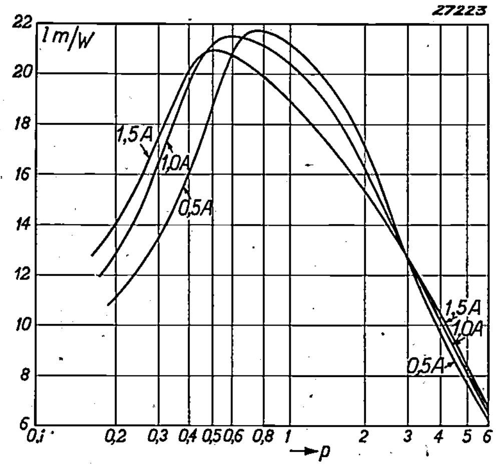

Fig. 5. Efficiency of the neon column with a diameter of 35 mm for various currents as a function of the pressure *p*.

b) The atom is again brought from the zero state to one of the 18.6 V levels and passes over into one of the 16.6 V levels with radiation of visible light. Before it returns to the zero state, however, it is again brought to an 18.6 V state by a collision with an electron for which 2 V are necessary and which are entirely transformed into visible light (*fig. 4b*). With a large number of successive excitations from the 16.6 V levels the efficiency becomes 137 lm/W.

In the actual discharge values are found which lie between the two extremes of case a) and case b), up to 30 lm/W. From this it follows that the process described under b) must play an important part, *i. e.*, the 16.6 V levels must act as a pseudo zero state. This phenomenon will be promoted by an increase of the self-absorption and especially by a lengthening of the life of the metastable atoms. As we have already mentioned these metastable atoms are destroyed by collisions with the wall or with other atoms. If the pressure is low they diffuse quickly to the wall, if the pressure is high many collisions with other atoms occur whereby they pass over into one of the two 16.6 V states which are not metastable. With a given diameter of tube and current, a maximum is found at a given pressure for the efficiency as a function of

3) The energy used for ionisation has not been taken into account and cannot be neglected. The true efficiency should therefore be considerably less.

the pressure (*fig.* 5). Gases other than the rare gases have a very unfavourable influence on the life of the metastable states, so that traces of such gases considerably lower the efficiency.

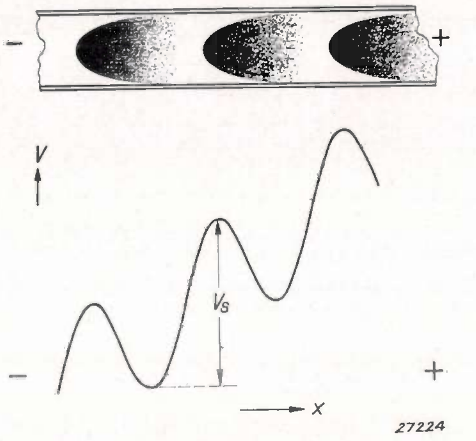

Fig. 6. Variation of the voltage and light intensity in a striated column (light-emitting parts are drawn darker).

The remarkable behaviour of neon upon excitation by electrons, when, instead of the lowest group of the 16.6 V levels being excited, the higher 18.6 V group is excited is encountered in the striae. Relative minima and maxima of the voltage (see *fig.* 6) which are superposed on the normal linear variation correspond to the successive dark and light spaces in the column. The increase in voltage V_s from minimum to maximum amounts to about 18.6 V. At the light spots the electrons give up their energy, and after having passed through about 18.6 V they may again become excited in the next light spot.

In the foregoing we have confined ourselves to direct current; since, however, practically all lamps work on alternating current we shall make a few brief remarks about the alternating current column in neon. *Fig.* 7 gives the variation of current I_e, lamp voltage V_e and light flux F as functions of the time in the course of a full period. The oscillograms refer to a neon lamp for the irradiation of plants, which is supplied with 220 V through a choke (type 4311). The current is practically sinusoidal, the voltage however has a different form. When the current is zero, the lamp goes out and must then be re-ignited. For this purpose a higher voltage, the so-called re-ignition voltage, is necessary at the beginning of every cycle. Further the voltage is low in the middle of the cycle when the current is high, a consequence of the negative characteristic of the positive column. The light flux F varies approximatively as the current I_e.

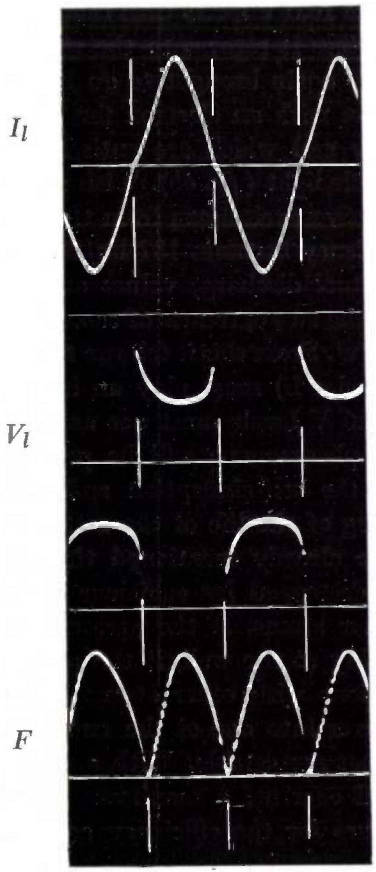

Fig. 7. Variation of lamp current I_l, lamp voltage V_l and light flux F for a neon lamp.

Sodium column discharge

In the case of sodium lamps we are concerned with a typical source radiating resonance light. *Fig.* 8 gives a scheme of several of the important energy levels; the transitions which lead to visible lines are represented by continuous lines and the wave lengths are given in Å. Spectral lines

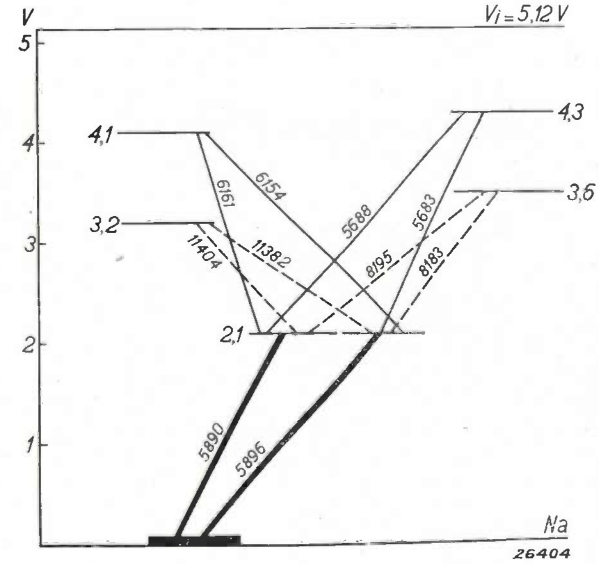

Fig. 8. Diagram of several of the energy levels of the sodium atom.

outside the visible region are indicated by dotted lines. It may be noted that, except for the double line 5 683—5 688, the resonance line 5 890—96 has the most favourable position of all lines with respect to the curve for the sensitivity of the eye. The relative sensitivity of the eye for $\lambda = 5\,890$—96 Å is about 0.765, so that when 1 W is radiated on this wave length, $0.765 \cdot 621 = 475$ lm is received [4]). The ideal would of course be to transform the electrical energy which is given to the electrons in the positive column entirely into yellow sodium radiation. This would mean a source of light with a yield of 475 lm/W. This is impossible for various reasons: there must always be some ionization to make possible the transport of current, and furthermore it is impossible to excite only the resonance line in a column. However, under favourable conditions about 90 per cent of the theoretically possible maximum for the sodium line can be attained, *i.e.* 425 lm/W.

These high values were obtained in the laboratory with a very low current density, the discharge tube being kept at the desired temperature of about 280° C in a special oven. This temperature is essential since otherwise the vapour pressure of the sodium would be insufficient. A lamp for practical use must, however, fulfil certain requirements, which need not be fulfilled by a lamp for laboratory tests. For example the lamp should be made as small as possible for a given light flux, both for the sake of decreasing the danger of breakage and so that the fixtures in which the lamp is to be mounted do not become too large. For this purpose the current density will be increased, and, because of the cumulative effects this is done at the expense of efficiency. With increasing current density the light flux only increases proportionally up to a certain limit, at higher currents the increase of light flux is less. *Fig. 9* shows this graphically.

Furthermore it is desirable that the lamp shall be

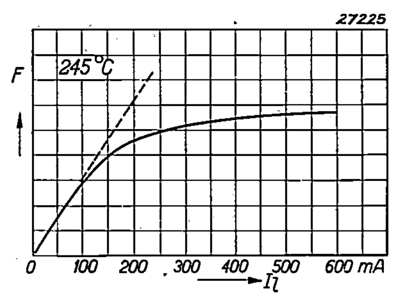

Fig. 9. Light flux F of the yellow sodium lines as a function of the current I_l at constant tube diameter and temperature.

[4]) It is known that 1 W radiated on the wave length $\lambda = 5\,550$ Å, which is the wave length of the maximum sensitivity of the eye, produces 621 lm.

selfheating and that it shall maintain its temperature without a special heating arrangement. As we have already stated the temperature of the coldest spot in the lamp must be about 280° C in order to obtain the necessary pressure of sodium vapour. At this temperature the heat dissipation to the surroundings is very important so that quite a large amount of energy is necessary per square centimeter of wall surface. This can be very much decreased by surrounding the lamp with an evacuated space since then heat dissipation by conduction and convection is practically eliminated and only heat radiation remains.

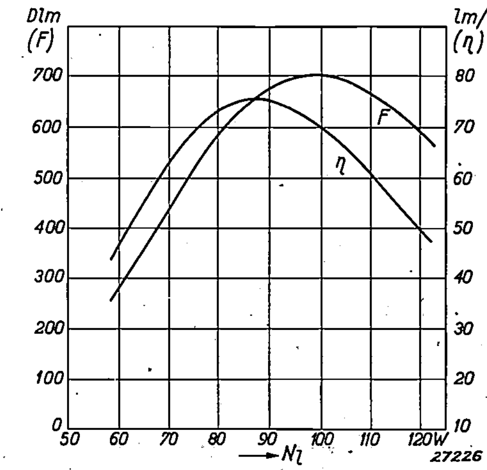

Fig. 10. Light flux F (in decalumens) and efficiency η (in lm/W) for a sodium lamp SO 650 as a function of the power on the lamp N_l.

Without special means a lamp containing only sodium would not ignite at ordinary temperature on the usual mains voltage, since the vapour pressure is only 10^{-11} mm. By the addition of a rare gas, usually neon, this difficulty is avoided. Upon ignition the discharge takes place in neon, the tube becomes warm and the sodium begins to vaporize. After several minutes so much metal is evaporated that the neon light is practically superseded by sodium light. With a good sodium lamp the neon light still contributes several per cent to the light flux, since the discharge takes place partly in neon and partly in sodium. By increasing the working temperature the vapour pressure of the metal can be made so high that the column discharge takes place wholly in the metal vapour, but the efficiency is then appreciably lower. *Fig. 10* illustrates this by means of the characteristic of a sodium lamp (SO 650) in which the light flux F is plotted as a function of N_l, the power in the lamp. At small values of N_l the light flux is also small, since due to the low temperature and pressure of sodium only neon light is radiated; at larger values of N_l the temperature and with it the vapour pressure of the sodium becomes greater so that more sodium light is also radiated. If, however, the value of N_l becomes greater than 100 W the light flux decreases because the efficiency decreases sharply due to the harmful cumulative effects. The range of values of N_l in which the light flux is a maximum is just the range where the neon as well as the sodium takes part in the discharge.

Let us assume that in a neon column there is sodium on the walls at such a temperature that it vaporizes in the discharge. As soon as the metal atoms enter the discharge, they are excited and ionized. The excited atoms will give off the energy as yellow radiation and diffuse further toward the axis, the ions, however, under the influence of the electrical transverse field return to the wall and are neutralized there by recombination. If now the current density is sufficiently high all the evaporating sodium will be ionized in the neighbourhood of the wall, and will be unable to reach the axis of the tube. At the centre of the discharge therefore the neon will be excited and ionized, and it is to this that the red-radiating core is due. There are therefore a considerable number of neon ions formed, with V_i = 21.5 V, and one is inclined to assume that it would be advantageous to replace them by sodium ions, with V_i = 5.1 V, whereby with constant current the potential gradient and the consumption of energy ought to decrease and the efficiency ought consequently to increase. The fact that this is not so according to fig. 10, may be explained by keeping in mind that the harmful cumulative effects become stronger in such a case. In good sodium lamps the presence of sodium is limited by a suitable combination of current density, wall temperature and tube diameter to a thin layer near the wall. The light emission therefore takes place in this thin layer (*v.* fig. 12), and since there is no sodium at the core, self-absorption and loss of light by cumulative effects are restricted.

This behaviour of the column is very obvious in sodium lamps working on alternating current. *Fig. 11* gives the variation of lamp current I_l, lamp voltage V_l and light flux F as functions of the time. The current I_l again has a practically sinusoidal form due to the self-induction put in series for the purpose of stabilization. The voltage here, however, is very much deformed since the concentration of the available sodium is changed very much in the course of the cycle by ionization. At the beginning of the cycle the re-ignition peak may again be distinguished but is lower there than in the case of a cold lamp where the discharge takes place in neon. The voltage now varies at first in a way analogous to that of the neon discharge of fig. 7, except that the values occurring are lower since

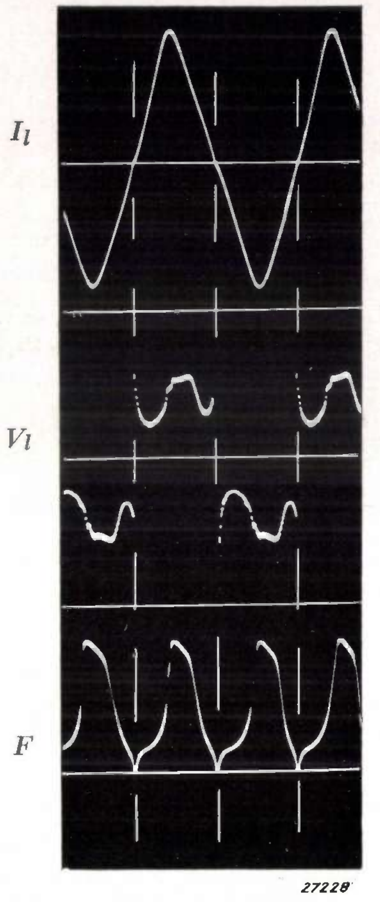

Fig. 11. Variation of lamp current I_l, lamp voltage V_l and light flux F as functions of the time for a sodium column lamp SO 650.

in the beginning the discharge takes place only in sodium. The current and the degree of ionization also increase with the time, in the axis of the tube all the sodium is ionized and at the wall temperature prevailing the rate of vaporization and of diffusion of the metal atoms are insufficient to supply the deficiency of sodium atoms, so that the neon now begins to take part in the discharge. This may be seen not only from the light emission, but also from the variation of the voltage which rises suddenly since neon with its high ionization potential (V_i = 21.5 V) is now being ionized. In this case the light flux does not vary directly as the current I_l in the phase (fig. 11) as in the neon discharge (fig. 7) but varies more irregularly due to the changing concentration of sodium.

Finally *fig. 12* gives a series of photographs for comparison of the transverse cross section of a neon column and a sodium column taken at corresponding moments with an alternating current discharge. The upper series which refers to neon show clearly that the production of light, like the current density, has a maximum at the axis. In the middle of the cycle when the current I_l has reached its maximum the light intensity is also a maximum, and the light is excited mainly at the axis. The lower series refers to the column of two excited levels. In such a case the promotion of cumulative excitation by increase of pressure and current density is favourable for the efficiency. In the case of mercury, just as with neon, the stimulation of cumulative excitation is facilitated by the fact that 6.7 and 4.8 V are resonance levels and that 5.45 and 4.7 V are metastable levels with a longer life. With a given diameter of tube

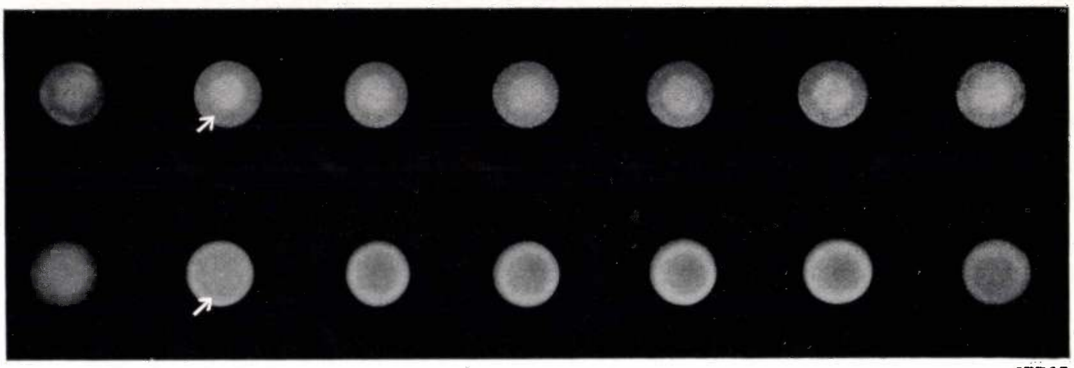

Fig. 12. Light in the cross section of a neon column (upper) and of a sodium column (lower) at different instants during the cycle of an alternating current discharge. The light border whose inner limit is indicated by the head of the arrow may be attributed to reflection on the inner surface of the tube. With neon the core is bright, with sodium it is dark.

a sodium lamp in ordinary use. At the beginning of the cycle, at a low current, the sodium light fills the whole cross section and it is at a maximum value on the axis. Half way through the cycle, at a high current, the sodium light is concentrated in a thin layer near the wall; neon light, however, now appears at the axis although of a much lower intensity. This is the so-called red core. At the end of the cycle, when the current has again become small, the sodium light again fills the whole cross section.

Column discharge in mercury vapour at low pressure

Just as with sodium we are here concerned with a discharge in a metal vapour, mercury, in which however the added rare gas no longer takes part in the discharge during normal functioning of the lamp, since at the ordinary working temperature, about 80° C, the vapour pressure of the mercury is already quite high, namely $\approx$ 0.1 mm.

The main levels of the mercury spectrum are indicated in *fig. 13*, the continuous lines represent the transitions which correspond to spectral lines in the visible, the accompanying numbers indicate the wave lengths in Å. Transitions which give ultraviolet lines are here again indicated by dotted lines. As may be seen from the figure the two resonance lines lie in the ultraviolet (2 537 and 1 850 Å). All the lines in the visible are so-called higher lines, they occur upon transitions between

and constant current the efficiency of the column discharge as a function of the mercury pressure shows a maximum at a definite pressure (*fig. 14*). Upon increasing the pressure the efficiency at first decreases but later increases again when contraction occurs upon transition from the low pressure to

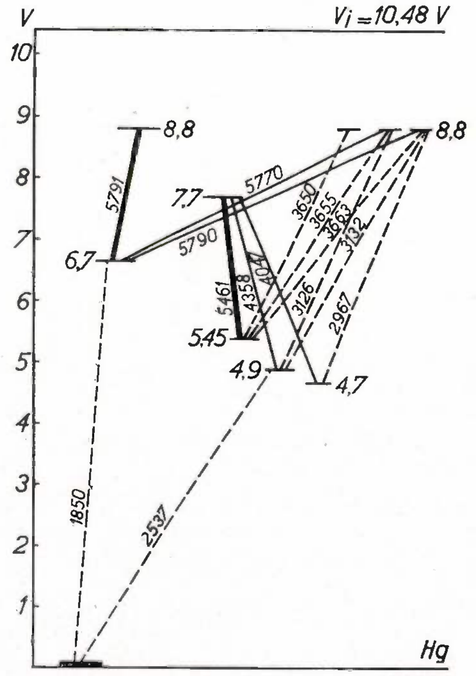

Fig. 13. Energy diagram of the mercury atom. Spectral lines in the visible region are continuous, those in the ultraviolet are broken. The numbers indicate the wave lengths in Å.

the high pressure discharge. The visible lines of the mercury vapour discharge at low pressure: 5 790 - 5 791 - 5 770 yellow, 5 461 green, 4 358 blue and 4 047 violet, are the same as appear in the high pressure discharge, the relative intensity is, however, different, so that the low pressure discharge appears bluer.

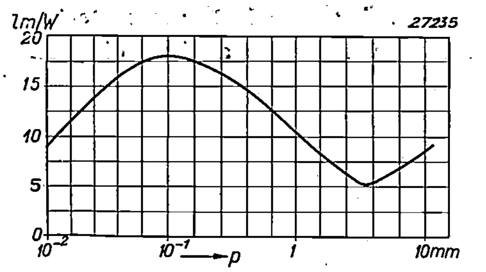

Fig. 14. Efficiency of a column discharge in mercury at low pressure, with a constant tube diameter (17 mm) and constant current, as a function of the pressure p.

The variation of lamp current I_l, lamp voltage V_l and light flux F as functions of the time with a low pressure mercury column is analogous to that of the curves given in fig. 7 for neon. As in the case of sodium the rare gas filling, usually a mixture of neon and argon, facilitates ignition in the cold state. It also plays an important part in the so-called re-ignition as is the case with all low pressure metal vapour lamps working on alternating current. Re-ignition occurs more readily since a greater number of charges are left over from the preceding cycle. Due to the presence of the rare gas the disappearance of the residual charges on the wall is retarded and the re-ignition voltage is thereby lowered.

COMPRESSION AND EXPANSION IN TRANSMISSION SOUND

by V. COHEN HENRIQUEZ. 534.322 : 534.86

The contrasts in speech and music which are reproduced electrically are limited by the ratio of intensity between the strongest sound which can be reproduced without distortion and the inevitable unwanted interference sound (noise and disturbances). This ratio is in general much smaller than the intensity ratios which appeared in the original sound. A faithful reproduction is therefore only possible by "compressing" the sound in one of the first links of the transmitting system and "expanding" it at the end in the reproduction apparatus. Several problems which occur in this process are discussed in this article, and a number of circuits are indicated which are suitable for this purpose

The sense of hearing enables us to observe air vibrations of very different intensities. On the one hand air vibrations with an amplitude down to 2×10^{-9} cm and a pressure of 3×10^{-4} dyne/cm^2 can just be heard, while on the other hand intensities which are 10^{13} times greater may be perceived by the ear without difficulties.

In the technique of electrical acoustics where the problem is to reproduce certain acoustic oscillations by electrical and mechanical means at a different spot or at a different time, the transmission of such a wide range of intensities is accompanied by difficulties. When we attempt to reproduce the acoustic oscillations through the different links of an electrical or mechanical system (*fig. 1*), it is fundamentally impossible to avoid the addition to the oscillations of a greater or lesser amount of irregular fluctuations which are commonly called noise, interference or hum, according to the nature of the sound which they produce. These fluctuations have been dealt with repeatedly in this periodical [1]).

For satisfactory reproduction it is necessary that the weakest acoustic oscillations emerge sufficiently above the level of disturbance, while the strongest oscillations must not lead to overloading. This condition is harder to satisfy the greater the ratio of intensities.

Attempts have of course long been made to limit the differences in intensity in the reproduction. For satisfactory reproduction it has been found unnecessary to be able to command the above-mentioned intensity range of 130 db between the barely audible sound and the limit imposed by feeling pain. In everyday life also we are always surrounded by a certain level of disturbance due to noises in which we are not directly interested such as the click of typewriters, the noise of engines, of the wind, etc. The intensity range commanded by an orchestra will not usually be more than 60 to 80 db. An intensity range of 60 db will in general be sufficient for the transmission of music.

1) See articles by M. Ziegler on noise, Philips techn. Rev. 2, 136 and 329, 1937; 3, 193, 1938.

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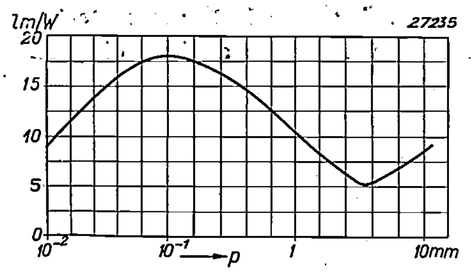

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But even this intensity range is greater than that permitted by ordinary systems for electrical transmission of sound. The maximum intensity range usual in practice is 35 to 45 db for broadcasting transmitters, and 25 to 35 db for gramophone records.

This limitation is usually taken into account by artificially reducing the contrasts in the signal to be transmitted, for instance by regulating the amplification between the microphone and the transmitter (or the apparatus for cutting gramophone records) in such a way that the amplification is weak at loud passages and strong at weak passages. This regulation, which is called compression, may be done in various ways which we shall mention later [2]). At the receiving end, after the signal has passed the different links where disturbances may be added to it, the original intensity relation may be restored by applying a method which is the inverse of compression, *i.e.* a higher amplification of the strong and a lower amplification of the weak signals.

In order to obtain precisely the correct intensity relations it is of course necessary that the second process, called sound expansion, is the exact inverse of the previous compression.

Compression and expansion characteristics

If it were always possible to match compression and expansion to each other, it would obviously be desirable to use them both to the greatest possible degree. Actually, however, the majority of receivers are not equipped with any device for expansion, and it is therefore necessary to choose the compression characteristic so that the reproduction remains as natural as possible even without expansion.

Let us imagine that a curtain were hung between an orchestra and the auditorium, and that the curtain had the property of reducing the strength of loud passages and increasing that of weak ones. The effect intended by the composer for the contrast between piano and forte would then be lost. It might perhaps be heard from the sound quality of the instruments that the musicians were playing loudly, but it would not be observed directly. Furthermore the subjective harmonics and combination tones which occur in the ear itself at great intensities would be missing. Compressed music would be less satisfying than normal music because the distribution of intensities intended by the conductor is destroyed.

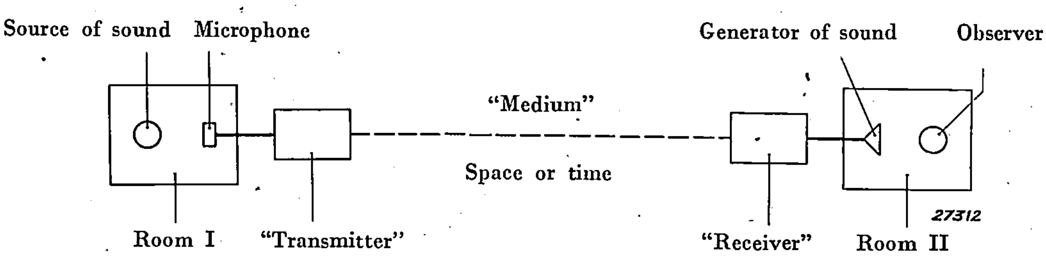


Fig. 1. The various links in sound transmission. "Transmitter" in its most general sense may for example also be an apparatus for the making of gramophone records or sound film. The concepts "Medium" and "Receiver" should also be taken in the same general sense. Each of these links may be the source of disturbances, for example:

Room I	: Insufficient acoustic insulation with the result that street noises and the like penetrate.
Microphone	: Noise of the microphone itself, insufficient shielding of connections sensitive to disturbances or interference.
"Transmitter" "Receiver"	: Noise from resistances and amplifier valves.
"Medium"	: Atmospheric disturbances in radio transmission. Scratching of the needle on gramophone records.

Just as we may speak of distortion in the transmission of music when all the frequencies are not reproduced in the correct relation, compressed music must also be qualified as distorted. An attempt must therefore be made to compress the music as little as possible.

If we assume that the signal to be transmitted varies between 20 and 80 phons, while the signal passed by the apparatus may only be within an intensity range of 45 db, the compression charac-

[2]) A form of compression upon sudden large variations in intensity is carried out by the ear, this is the so-called protective reflex of Kreide and Kato. If the ear is adjusted to the perception of weak noises and is then suddenly strongly stimulated, the *musculus tensor tympani* contracts as a reflex. The ear drum is thus drawn into such a position that the transmission of sound to the inner ear is very much hindered. A more familiar example of compression applied by an organ of the human body is the variation in the diameter of the pupil of the eye, which serves to lessen the variations in intensity of illumination of the retina.

teristic must be chosen so that the strength of the signal sent out increases only by 45 db between 20 and 80 phons. The characteristic may have various forms. Curves a and b of *fig. 2* are alike in that the

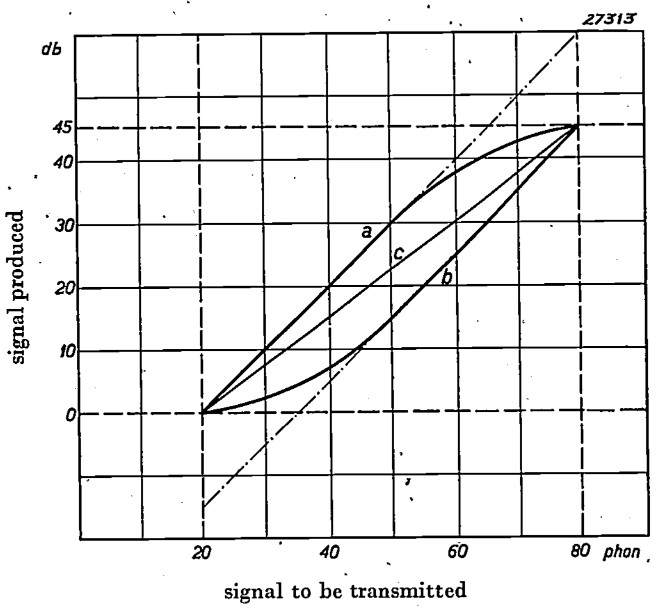


Fig. 2. Different compression characteristics which reduce variations in intensities from 60 db to 45 db.

intensity of reproduction is distorted only in a portion of the intensity range. Curve a more nearly resembles a "limiter", in which the amplification is decreased above a certain strength of signal, and forms a contrast to curve b in which the amplification is increased as soon as the signal falls below a certain limit. Such characteristics may be satisfactory when the fluctuations in intensity of the sound to be reproduced are so small that in general the right-hand portion of the characteristic is used,

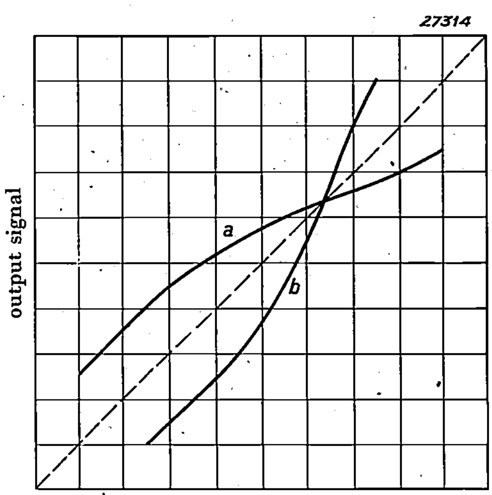


Fig. 3. A compression characteristic (a) and the corresponding expansion characteristic (b) may be derived from each other by drawing the respective mirror images in relation to a line at 45°.

and compression only comes into action in exceptional cases. When, however, the intensity continually varies over the whole region of 60 db, it is better to distribute the compression uniformly as is done for instance in curve c of fig. 2. The corresponding expansion characteristic may always be found by exchanging ordinate for abscissa, in other words by drawing the mirror image of the curve with respect to a line at 45°. This is shown in *fig. 3*.

The most satisfactory characteristic depends, as we have seen, upon the character of the music. If the same characteristic must be used for every case a uniform distribution of compression is to be preferred.

Compression is very often not carried out automatically but manually: the amplification is regulated by hand according to the intensity of the music. In this case there need be no permanent relation between the intensity of the original and of the compressed sound; it is then possible to take into account the special characteristics of the score [3]), and in this way to make the compression less noticeable. Against this is the objection that it is impossible to cancel this kind of compression accurately by automatic expansion in the receiver.

Time factor

In the foregoing we have seen that the application of compression or expansion means that the output intensity of a link in the transmission system (an amplifier for instance) does not increase proportionally to the input intensity. Before showing how this effect is achieved, we shall first point out the distinction between two fundamentally different possibilities:

1) The output amplitude at any given instant does not increase proportionally with the input amplitude at the same instant; the amplification is thus essentially non-linear.
2) The relation between output voltage and input voltage is linear; the amplification factor, however, is a slowly varying quantity which depends upon the average intensity over a certain interval of time.

The first system is practically unusable because the non-linear amplification of the signal leads to overtones, and a greater frequency range must therefore be transmitted. Moreover, with an inaccurately matched expansion characteristic overtones and combination tones will occur in the

[3]) Regulation with reference to a score was discussed in an article by R. Vermeulen: The relationship between Fortissimo and Pianissimo, Philips techn. Rev. 2, 266, 1937.

reproduction which are much more disturbing than an inaccurate reproduction of intensities.

In the second system non-linear distortions need not occur when the change in the amplification factor is made to take place so slowly by means of a retarding element that during a period the amplification of the lowest frequencies which are reproduced remains practically unchanged. On the other hand the regulation may not take place too slowly for the following reasons:

1) Upon a very sudden increase of intensity the retardation of compression may result in a temporary overloading of the apparatus, and therefore in temporarily distorted reproduction. The quicker the regulation, the less noticeable is this distortion.
2) If compression and expansion are not the inverse of each other, the regulation will itself become audible when regulation is too slow. For example after the beginning of a loud passage an increase in the amplification will be noticed due to increasing expansion. This will also be the case when compression is by hand and expansion automatic.

In order to make a suitable choice for the rate of regulation, we must consider several characteristics of the ear which are important in this connection. There exists for the ear a subjective time interval for the building up and dying out of sound. That is to say, there is no difference in observation between the sudden and the gradual beginning of a tone if the latter is not too slow. Let the intensity be represented as a function of time by:

$$I = I_{\text{final}} \left(1 - e^{-\frac{t}{a}}\right),$$

it will then be found that one observes no difference between $a = 0$ (sudden beginning) and $a = 0.07$ s. This interval of 0.07 s which is the subjective initial sound lag, is independent of the intensity of the sound.

If one listens to the difference between a tone which ends abruptly and one which decreases more gradually according to the equation

$$I = I_0\, e^{-\frac{t}{a}},$$

here again no difference is observed between $a = 0$ and values of a up to 0.1 s for weak, and up to 0.3 s for strong initial intensities. This shows that the sound impression continues to exist for a short time after the tone has been interrupted, and that the time necessary for its disappearance amounts to several tenths of a second.

The foregoing leads to the following choice of regulation times:

The initial time delay (increasing compression or expansion) should be less than 0.07 s; the final time delay should be within several tenths of a second.

These times are fortunately so long that signals with low frequencies (40 to 50 c/s) are reproduced without unpleasant distortion.

Meanwhile it was found desirable in practice to choose the final time delay somewhat longer since otherwise, with manual compression and automatic expansion, a disturbing effect appeared in the reproduction of percussion instruments. The sound of such instruments, that of the piano for instance, dies out in a damped oscillation. If the final time delay of the "expander" is chosen small, the amplification decreases during the dying out of the vibration of the note struck so that the note reproduced seems to die out more rapidly. With high expansion this effect may be very disturbing. With a final time delay of 0.5 to 1 s and not too high expansion the phenomenon is unnoticeable.

Adaptation of compression and expansion

If we wish to remove completely the distortion caused by the compression of the sound, not only must we choose the characteristic of the expander correctly, but we must in addition make the process of regulation in the expander such that the effect of regulation in the compressor is completely cancelled. This is actually possible, and consists of making the time lags of compressor and expander the same. If alternating voltage is supplied to the input

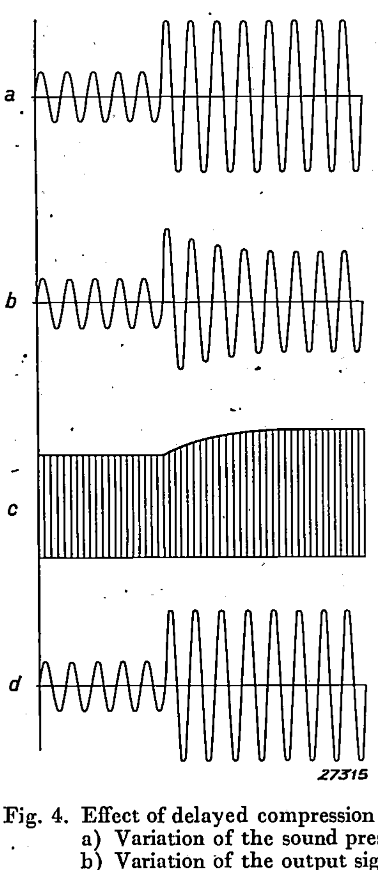


Fig. 4. Effect of delayed compression and expansion
a) Variation of the sound pressure;
b) Variation of the output signal of the transmitters;
c) Variation of the amplification factor of the receiver;
d) Variation of the output signal of the receiver.

side of a compressor and this voltage suddenly changes in amplitude (*fig. 4a*) the output voltage may be represented as a function of time by fig. 4*b*. If such a signal is now supplied to an expander the amplification factor of the receiver will increase gradually after the sudden change in signal strength and vary according to fig. 4*c*. The signal produced by the receiver is equal to the product of the input signal (curve *b*) and the amplification (curve *c*), and with suitable construction of the retarding element it may be made to correspond exactly with curve *a*.

Circuits

The usual circuits for automatic compression or expansion may be divided into two main groups:

A) Circuits in which the amplification is influenced by the alternation of the voltage relations of one or more amplifier valves,

B) Circuits in which use is made of the change in resistance of a filament upon being heated by an electric current.

We do not intend to give a complete survey of the possible circuits; we shall only discuss several possibilities in a general way.

A. Influence exerted on the amplification by a change in the voltage relation of amplifier valves

For changing the amplification a voltage is necessary which is dependent on the intensity of the signal. This may be obtained by rectifying the signal; the result of this detection, the regulation voltage, is then fed to the regulating amplifier over suitable retarding elements. By the choice of the polarity of this voltage a compressor or an expander may be obtained as desired.

In the regulating amplifier a valve is usually used whose slope varies with the grid voltage. In order to avoid non-linear distortions the input alternating voltage which such a valve receives must be amplified with a practically constant slope. The variations of this voltage must therefore be much smaller than the variations in the grid bias used for regulation. In general therefore the regulating amplifier will only be able to handle signals which are too small to give proper detection. It is therefore necessary to amplify the signal before it is detected, and to use for this either the amplifier proper or a separate detector amplifier. This brings us to the block diagrams of *figs. 5a* and *b*.

The difference between 5*a* and *b* is that the amplification factor in case *a* is determined by the input voltage and in case *b* by the output voltage of the regulating amplifier. The use of the latter circuit may, however, be accompanied by difficulties, because regulation by means of the output voltage may lead to instability. This may be explained as follows.

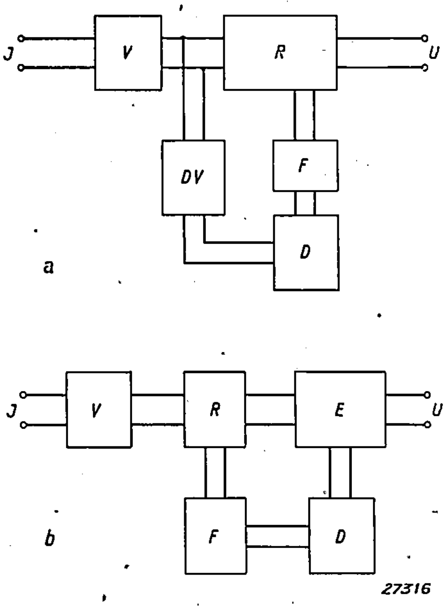


Fig. 5. Diagram of a compressor or an expander.
a) The amplification factor is determined by the input signal;
b) The amplification factor is determined by the output signal.
V pre-amplifier, *R* regulation stage, *F* filter, *D* detector, *DV* detector amplifier, *E* final amplifier.

If P_i is the level of the input signal (in nepers), P_u the level of the output signal, then:

$$P_u = P_i + \varphi ,$$

where φ is the amplification factor (also in nepers). Now the amplification factor of the regulating amplifier changes with the output signal. For small variations we may write:

$$\varphi = \varphi_0 + a\,P_u ,$$

where a is positive in the case of expansion and negative in that of compression. From the above equations it follows that:

$$P_u = P_i + \varphi_0 + a\,P_u ,$$

$$P_u = \frac{P_i + \varphi_0}{1 - a} .$$

Stable functioning is only possible when the output voltage remains finite. a must therefore be negative or positive but less than one. In other words: stable working occurs only when the percentage increase of the amplification factor is less than the percentage increase of the output voltage [4]). If the amplification factor is determined by the input signal as in fig. 5*a*, such a phenomenon cannot occur. The output signal is then always finite, and is determined by an equation of the form:

$$P_u = (1 + a)\ P_i + \varphi_0.$$

Among the possible methods of construction of the regulator the following may be mentioned: change in amplification by change in the slope of a valve (this possibility was discussed above), and change in amplification by change in the external resistance of a valve. For this purpose a second valve may be used as loading resistance whose internal resistance depends closely on the grid voltage. The regulating voltage is supplied to the grid of this second valve.

Two regulating valves are often used in a push-pull circuit to combat secondary effects in regulation, such as non-linear phenomena and voltage surges.

B. Influence exerted on the amplification by changing the resistance of a filament.

If use is made of a change in the resistance of a filament with the temperature, a material must be chosen for the filament with which a wide variation of resistance is possible. For this condition it is necessary not only that the material should have a large temperature coefficient of specific resistance, but also that it should be able to withstand a high temperature without changing its shape or melting. It is found that the maximum changes in resistance may be obtained with tungsten, although its temperature coefficient is smaller than that of iron; tungsten, however, can easily withstand temperatures of 2 100° C, while iron melts at 1 528° C. Moreover, tungsten is very suitable for the making of the filament which when sealed into a bulb in the form of a lamp forms an object which can easily be handled. Before going into the properties which the lamp should possess we shall first discuss several examples of circuits for expansion. Compression circuits may also be developed on the same principles; they may often be derived from the expansion circuit in a simple manner.

The latter possibility is offered by the bridge circuit as in *fig. 6*. The bridge consists of two lamps in two opposite branches and of two constant resistances R_1 and R_2. If the resistances are so adjusted that for room temperature of the

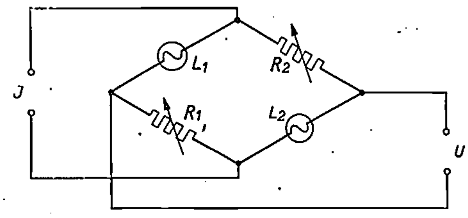


Fig. 6. Bridge circuit for obtaining the regulation voltage of an expander. The resistances R_1 and R_2 are so adjusted that the bridge is in equilibrium at room temperature of the filaments L_1 and L_2. When an alternating voltage is applied the temperature of the filaments increases. The bridge is then no longer in equilibrium and a voltage acts across the output terminals which increases more rapidly than the alternating voltage.

filaments L_1 and L_2 the bridge is in equilibrium, the ratio between the output voltage and the input voltage is small for voltages which scarcely heat the filaments; for higher voltages, which do heat the filaments, the bridge is not in equilibrium and V_u/V_i will become greater.

In this way an expander is obtained. By adjusting the bridge for equilibrium with hot filaments it is just possible to make V_u/V_i become greater for weaker signals; this results in a compressor.

Another possibility is the circuit of *fig. 7*. In this

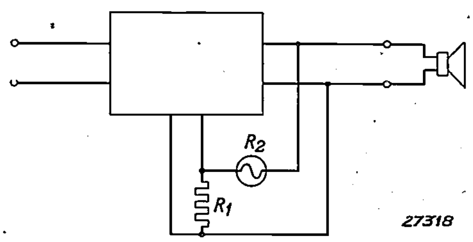


Fig. 7. A very simple expansion circuit, which uses a varying inverse feedback. With an increasing output-voltage the resistance R_2 of the lamp becomes greater. In this way the strength of inverse feedback is reduced and the amplification factor increases.

use is made of an inverse feed-back which is dependent on the output voltage. The amplification factor μ of an inverse feed-back amplifier may be represented by [5]):

$$\mu = \frac{a}{1 + a\beta}.$$

a is here the amplification, *i.e.* the ratio of input to output voltage without inverse feed-back, and β is the ratio of the inverse feed-back voltage to the output voltage. In the case of fig. 7, therefore, $\beta = R_1 + (R_1 + R_2)$. By increasing resistance R_2,

[4]) The condition is necessary but not sufficient by itself. Instability may for example also occur when a circuit as in fig. 5*b* is used for compression and a is negative. This possibility occurs, when due to incorrectly chosen filters, signal frequencies return to the regulator over the detector, thus forming a closed system which can continue to work by itself with sufficient amplification.

[5]) See the article "Inverse Feed-back" in Philips techn. Rev. 2, 289, 1937.

β becomes smaller and therefore the amplification μ becomes greater.

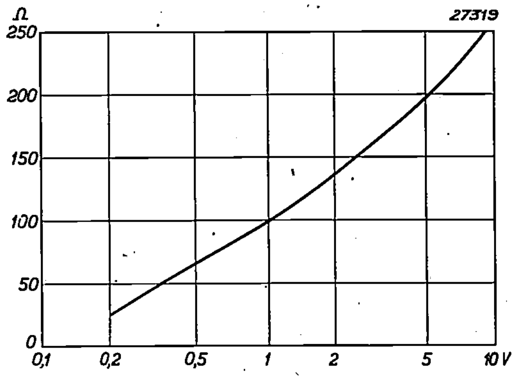


Fig. 8. Resistance of the lamp used for expansion circuits as a function of the voltage.

This circuit has the advantage of great simplicity; in order to obtain the desired form of expansion characteristic, however, several special features are required of the lamp:

1) In the region of the normal voltages which occur at the output end of a radio receiver the resistance of the lamp must change considerably.
2) The lamp should not take up too much energy since it does so at the expense of the energy delivered by the loudspeaker.
3) The times for heating and cooling of the filament must be such that the correct regulation times are obtained. This condition applies of course to every expansion circuit which uses heated filaments.

These requirements could only be satisfied by using a very thin wire wound in the shape of a spiral; one of the purposes of the spiral form is to increase the temperature of the wire with as little energy as possible. The filament is situated in a vacuum so that it can give off heat only by radiation. In order to give the resistance of the lamp a suitable value and to provide that it consumes only a small amount of energy it has been found necessary to use a wire of only 9 μ diameter.

The time necessary for heating and cooling of the lamp were found to correspond very well with the desired values; the time for heating up amounts to about 0.1 s and that for cooling off to about 0.6 s.

Fig. 8 gives the relation between the resistance of the lamp and the voltage, while *fig. 9* gives an example of an easily obtainable expansion characteristic.

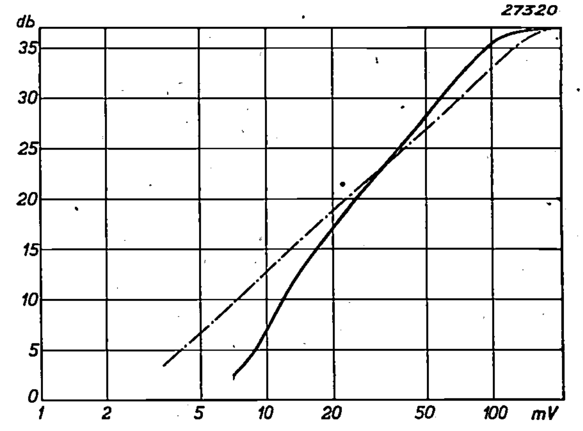


Fig. 9. Expansion characteristic obtained with a circuit as in fig. 7. An intensity range of 24 db at the input side is expanded to 34 db. The chain dotted line gives the characteristic in the absence of expansion.

PHENOMENA IN AMPLIFIER VALVES CAUSED BY SECONDARY EMISSION

by J. L. H. JONKER.

537.533.8 : 621.396.645.3

In amplifier valves secondary emission may occur not only from electrodes but also from non-conductors which are at a positive potential. In this article several phenomena are described which are connected with this secondary emission.
The secondary emission of the anode has an unfavourable effect on the characteristic of tetrodes. Improvement may be obtained by covering the anode with a suitable material or providing it with ribs, and further by employing an extra electrode and by utilizing the space charge in a suitable manner.
The secondary emission of insulators results from the fact that an insulator which is struck by electrons and emits secondary electrons may take on a considerable positive potential. This effect is applied in the fluorescent screen of cathode ray tubes, and may also occur as an undesired phenomenon in radio valves where it causes strong distortion and output damping. Finally several ways are indicated of combatting the secondary emission of insulators.

When a plate is struck by electrons which have been accelerated by a sufficiently high voltage, secondary electrons are freed from the plate as has been described in previous articles in this periodical [1, 2]).

In the second article cited it was explained how deliberate use can be made of the phenomenon of secondary emission to amplify electron currents; in this case specially prepared electrodes are used which give a high secondary emission upon bombardment with primary electrons. Secondary electrons may, however, appear in every amplifier valve when electrons which are accelerated by the voltages common in amplifier valves strike a conducting or insulating surface. In general the number of these secondary electrons, even in the case of surfaces which are not specially prepared, may be of the same order of magnitude as that of the primary electrons.

In this article we shall discuss the phenomena which occur in an amplifier valve as a result of this secondary emission, and we shall also study the available means of combatting secondary emission in cases where it is undesired, or at least of preventing its having an unfavourable effect on the action of the valve.

Secondary emission of conductors

As an example of a conductor on which an undesired secondary emission may occur we shall consider the anode of a tetrode, *i.e.* of a valve with a cathode, a control grid, a screen grid and an anode.

With such a valve, upon large changes of the anode voltage such as occur in final amplifier valves, the minima of the anode voltage may be considerably lower than the screen grid voltage. The continuous line in *fig. 1* represents the anode

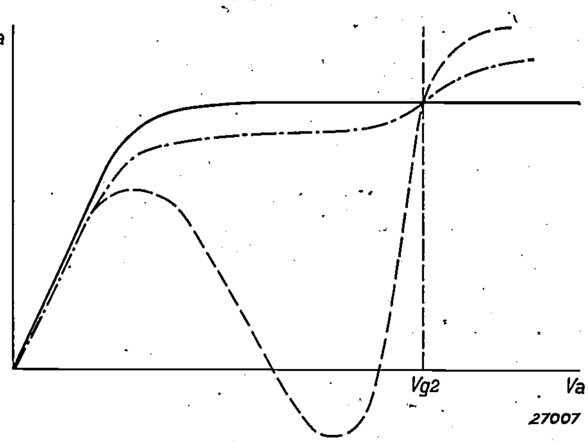


Fig. 1. Current-voltage curves of tetrodes.
-------- anode current of a tetrode as a function of the anode voltage. V_{g_2} is the screen grid voltage.
—·—·— current-voltage curve when secondary emission is partially suppressed by means of a layer of carbon and ribs on the anode.
——— current-voltage curve in the absence of secondary emission.

current which would flow if no secondary emission occurred, as a function of the anode voltage (at constant screen grid voltage). If, however, secondary emission does occur the current varies according to the dotted line.

The difference between the two curves may be explained as follows. If the anode voltage is higher than the screen grid voltage, the secondary electrons which are freed from the anode are driven back to the anode by an opposing field. The primary electrons, much fewer in number, which strike the screen grid will also free secondary electrons. These electrons are drawn to the anode by the same field, so that the anode current is slightly higher

[1]) H. Bruining, Secondary Electron Emission, Philips techn. Rev. 3, 80, 1938.
[2]) J. L. H. Jonker and M. C. Teves, Technical Applications of Secondary Emission, Philips techn. Rev. 3, 137, 1938.

with secondary emission than without it.

If the anode voltage becomes lower than the screen grid voltage, the direction of the field changes, and the secondary electrons from the anode will be attracted to the screen grid. The anode current therefore decreases suddenly and the screen grid current increases, which results in high distortions in the anode alternating current. The means available for combatting these undesired phenomena may be divided into two groups:

1) attempts may be made to keep the secondary emission of the anode low,
2) attempts may be made to make the action of the secondary electrons harmless by suitable construction of the anode or by the use of an extra electrode.

The possibilities of lowering the secondary emission have already been described in detail in this periodical (Bruining, *loc. cit.*).

The lowest secondary emission is obtained from a layer of carbon which is deposited in such a way that the surface is rough or flaky, so that the secondary electrons are for the most part captured in the cavities between the grains of carbon and cannot be drawn away. The secondary emission which has been lowered in this way, however, still amounts to 30 - 50% of the primary electron current. A complete suppression of secondary emission cannot therefore be achieved by this method, and in discharge tubes there are a number of factors acting which make it difficult to retain even these low values when the tube is in use. One of these factors has already been mentioned in a previous article, namely the alteration of the surface by the deposition of evaporated material, barium and barium oxide, from the hot cathode. The high temperature during the outgassing of the electrodes, the gases freed during the process of manufacture and the getter employed may also be factors causing a surface modification which increases the values of the secondary emission.

Considering the fact that s u p p r e s s i o n of secondary electron emission was found to be impossible, other methods were sought of neutralizing the effect of secondary emission. It is possible to do this to a great extent by giving the anode ribs or partitions (see *fig. 2*). The same effect is obtained in this way as by the use of a porous carbon layer. The secondary electrons freed between the partitions are to a large extent recaptured by the partitions and thus return to the anode.

In fig. 1 the line drawn thus:·——·——· represents the anode current of such a valve, which has in addition its anode and screen grid covered with a layer of carbon. As was to be expected the secondary electron current from the anode to the screen grid is considerably reduced but not completely suppressed. The remaining irregularity of the curve proves that the secondary emission is still appreciable.

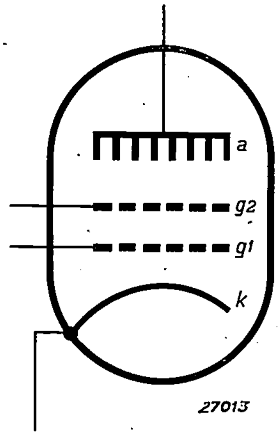


Fig. 2. Tetrode with ribbed anode. The secondary electrons which leave the anode are to a large extent recaptured by the ribs.

A completely effective method of preventing the passage of secondary electrons from the anode to the screen grid is the introduction of an extra electrode, a so-called suppressor grid, which when added to a tetrode converts it into a pentode. This method is used in most radio valves.

How does a suppressor grid work? Its action is based on the fact that the secondary electrons have for the greater part relatively low velocities, so that a counter voltage of 20 - 30 volts is sufficient to send practically all the secondary electrons back to the anode.

If a fine-meshed grid were placed in front of the anode and if it were kept continually about 20 volts lower in potential than the anode, secondary emission would be suppressed. Practically the same effect can be achieved by placing between screen

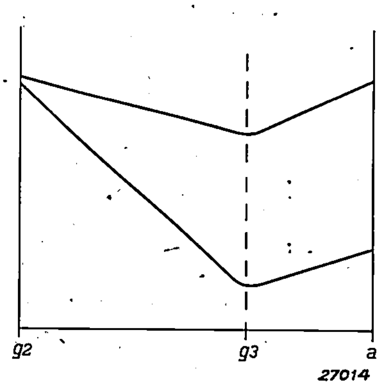


Fig. 3. Variation of the potential between screen grid and anode in a pentode without space charge (plane parallel electrodes).

grid and anode a wide-meshed grid which is connected to a point of low potential, for instance the cathode. By a suitable choice of the dimensions of the mesh and position of this grid the lowest voltage along the path of the electrons between anode and screen grid can always be kept sufficiently lower than the anode voltage to cause secondary electrons to return to the anode. On the other hand this potential must always be positive even with the lowest instantaneous anode voltage, since otherwise all electrons will not reach the anode. In *fig. 3* the variation of potential is given for two values of the anode voltage, space charges between screen grid and anode being neglected.

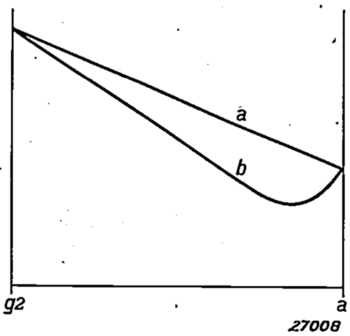


Fig. 4. Variation of potential between screen grid and anode in a tetrode; *a* without, and *b* with space charge. The space charge prevents the passage of secondary electrons from anode to screen grid.

The presence of the space charge has the result that the line representing the variation in potential between two flat electrodes is no longer straight, but curves. *Fig. 4b* shows the variation of potential between the screen grid and the anode as it would be with a large anode current (large space charge) a low anode voltage, and no suppressor grid.

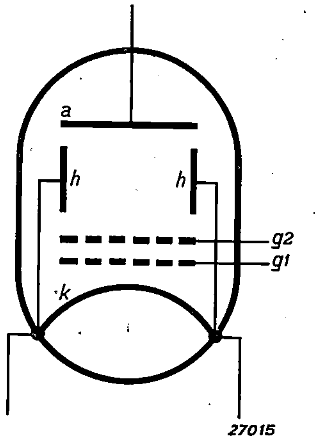


Fig. 5. In order to prevent the transition of secondary electrons from anode to screen grid, the potential in the space between these electrodes must be lowered. This is achieved here by means of the auxiliary electrodes *h* which are kept at cathode potential. At the same time the transition time of the electrons and therefore also the space charge is increased, whereby the potential is still further lowered.

It may be seen that with a large current between screen grid and anode a potential minimum may occur here also, and may be able of preventing the passage of secondary electrons between anode and screen grid.

A practical type of construction which makes use of this effect was obtained by combining the action of the space charge with that of an extra electrode, consisting here of two plates connected with the cathode (*fig. 5*). This construction is simpler than that with a suppressor grid and it entirely suppresses the transition of secondary electrons down to an anode voltage of about 30 V, as may be seen from *fig. 6* in which the variation of the anode current is given as a function of the anode voltage. The anode current remains fairly constant upon decrease of the anode voltage to about 30 V, and then decreases suddenly with a sharp bend.

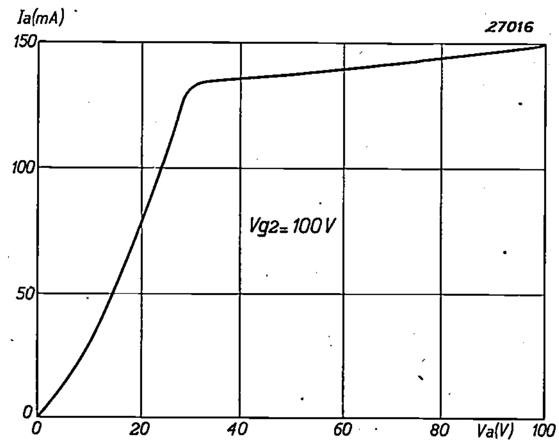


Fig. 6. I_a-V_a characteristic of a tetrode with auxiliary electrode as in fig. 5.

Secondary emission of insulators

Let us consider an insulating part of a radio valve which is struck by electrons, for instance an insulating support. This component, which may easily have a large secondary emission factor, will be connected to the different electrodes over the poorly conducting insulation. A given surface element of the insulating support may now be considered as a secondary emitting plate P (see *fig. 7*), which is connected over a high resistance R with a type of average voltage V_0 of the electrodes, which will for instance be somewhat lower than the anode voltage. This scheme corresponds exactly with that of a dynatron (see article quoted in footnote [2]), and the same is true of the current I_p of the secondary emitting plate as a function of its voltage. The characteristic is given in *fig. 8a*. With increasing voltage I_p at first increases, reaches a maximum and then decreases (it may even become negative), because the coefficient of secondary

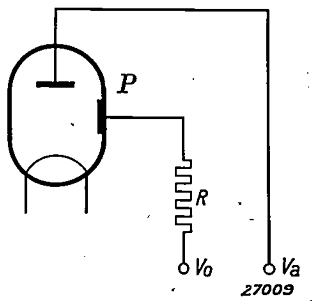


Fig. 7. Equivalent circuit for an insulating part of a radio valve which is struck by electrons.

emission increases so much in a given range of voltages that the number of secondary electrons increases more rapidly than the number of primary electrons. As soon as the plate voltage is higher than the anode voltage V_a, however, the secondary electrons can no longer leave the plate and the current suddenly increases rapidly to the value which it would have in the absence of secondary emission.

What potential will the secondary emitting plate assume in a stationary state? From fig. 7 the following relation may be seen to hold between the current I_p and the voltage V_p on the plate:

$$V_p = V_0 - I_p R .$$

This relation is represented in fig. 8 (straight

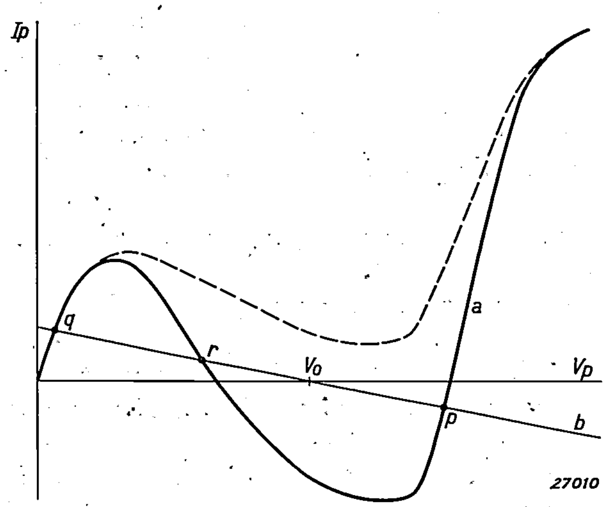


Fig. 8. Curve a: current I_p of a plate with a high secondary emission which is bombarded with electrons as a function of the plate voltage. With a lower secondary emission the current will vary according to the dotted-line curve.
Curve b: voltage V_p of the same plate as a function of the current when the plate is connected with a source of direct voltage V_0 over a high resistance.
The points of intersection of the lines a and b determine the stationary states. The adjustment has two values, given by p and q. Point r is an unstable state. In state p the potential of the plate is higher than the voltage V_0.

line b), together with the characteristic a which also gives a relation between I_p and V_p. The stationary states are given by the points of intersection p, q and r. The intersection q always occurs: the intersection p and r will not occur if the secondary emission is so low that the characteristic has for example the form of the dotted line curve. We see that the secondary emitting plate can adjust itself to different potentials when the secondary emission is high. The lowest potential (q) corresponds approximately to that of the cathode, the highest (p) to that of the anode; the adjustment r is not stable.

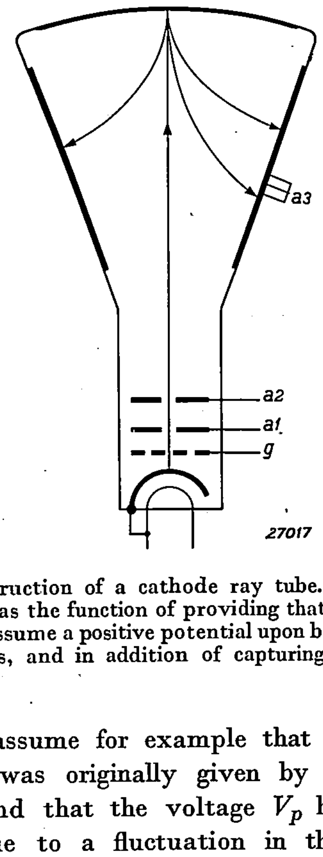


Fig. 9. Construction of a cathode ray tube. The collecting electrode a_3 has the function of providing that the fluorescent screen shall assume a positive potential upon being bombarded with electrons, and in addition of capturing the secondary electrons.

Let us assume for example that the state of the plate was originally given by the working point r, and that the voltage V_p has increased slightly due to a fluctuation in the secondary emission. The result is that the supply of electrons to the plate ($I_{primary} - I_{secondary}$) decreases more rapidly (curve a) than the loss of electrons over the leakage resistance R (curve b). The plate will therefore assume a positive charge, and as a result its potential will increase further until the stationary state p is reached.

The fact that an insulator can reach a high potential and give a high secondary emission upon being bombarded by electrons, sometimes leads to undesired phenomena in radio valves. Before going into this matter we shall show how this same phenomenon may be put to practical use in cathode ray tubes.

The fluorescent screen of cathode ray tubes consists of an insulating material and must be at a high potential in order that the electrons may strike it with sufficient kinetic energy to produce fluorescence. If the secondary emission of the screen were lower than the primary electron current, the screen would assume a negative charge and the high potential could not continue to exist. The screen must therefore have a high secondary emission. Furthermore, care must be taken to ensure that, upon switching on, the screen can reach a potential higher than r in fig. 8, otherwise it would not adjust itself to the working point p but to point q.

In *fig. 9* it is shown how the construction may be carried out practically. Close to the screen the inner wall of the tube is covered by an electrode which is raised to a high potential. This electrode has a two-fold purpose: first that of providing that the potential of the screen shall be such that upon switching on a sufficiently high secondary emission will occur, and second that of being able to draw the secondary electrons away rapidly.

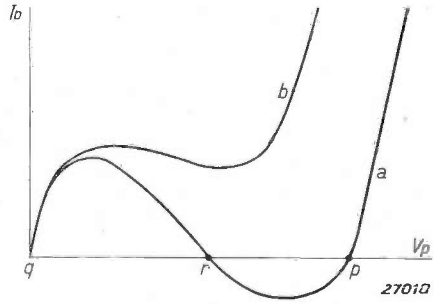


Fig. 10. Current from an insulating component in a radio valve as a function of the potential of that part; *a* at a high anode voltage, *b* at a low anode voltage.

As was stated above, the secondary emission of non-conductors may also occur as an undesired effect on the insulators and glass walls of radio valves. The phenomena may lead in two ways to disturbances: with power amplifier valves distortion occurs, with high-frequency amplifier valves an extra damping occurs in the output circuit.

With power amplifier valves where the alternating anode voltage has a large amplitude, the I_p - V_p characteristic of an insulator in the neighbourhood of the anode will vary very much. In certain cases the characteristic at the highest anode voltages may be such that a working point p of high potential exists, while this is not the case at the lowest anode voltage, because the secondary electrons from the insulator are not drawn away sufficiently rapidly (v. *fig. 10*). In such a case it may happen that the potential of the insulator upon rise and fall of the anode voltage jumps up and down between a high and a low value. Upon each jump in potential a charge may be induced on the modulating grid which causes transients in the anode current v. (*fig. 11*).

In high-frequency valves, where it is desirable that the anode current shall change little with the anode voltage (*i.e.* that the internal resistance shall be high), difficulties may be experienced due to secondary electrons which are emitted from an insulator and reach the anode. Considering the fact that the insulator itself can conduct no current, this extra current is equal to the primary electron current to the insulator. This primary current is in

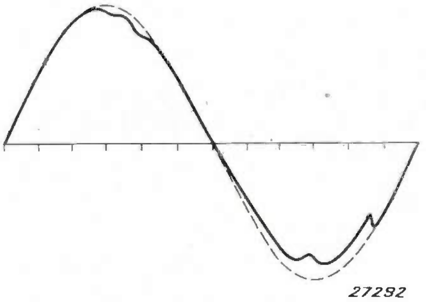
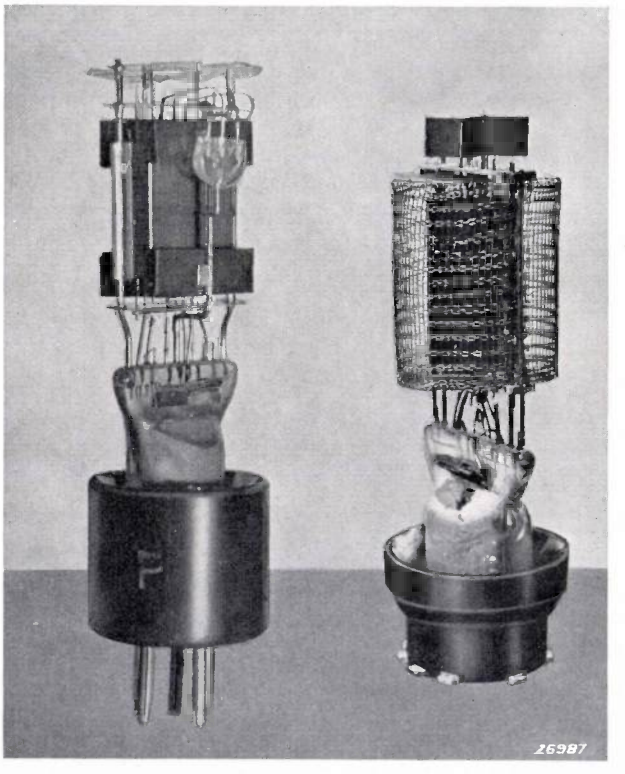


Fig. 11. Variation of the anode current for a power amplifier valve with sinusoidal grid voltage. The curve exhibits a number of irregularities due to a sudden jump in the potential of an insulating support or of the glass wall.



a *b*

Fig. 12 *a*) Power amplifier valve in which the mica supporting plates and the glass wall are shielded against bombardment with electrons by enclosing caps at the ends of the system. *b*) In an amplifier valve with a gauze anode the glass wall is exposed to a strong bombardment by electrons. In order to avoid this the anode is surrounded by a gauze cage which is connected to the cathode.

general very greatly influenced by the potential of the insulator which will change about equally with the anode voltage. The secondary electron current which reaches the anode will therefore also change sharply with the anode voltage, and this means that the internal resistance becomes considerably lower and the amplification of the high-frequency stage diminishes.

In order to avoid these harmful effects here again several methods may be chosen. When, for example, the disturbing phenomenum is due to insulating supports within the system, the most practical solution will be to cover them with a substance whose secondary emission is less than unity (for instance a layer of carbon). If the danger spot is the inner side of the bulb, the jump in potential can also be avoided by earthing this surface through a capacity. In order to do this the outside of the bulb may be covered with a layer of metal which is kept at a low potential (that of the cathode for instance). Finally the charging of the wall of the bulb may be avoided by entirely preventing electrons from leaving the electrode system and reaching the glass wall. This may be achieved by means of shields. *Fig. 12a* shows a power amplifier valve with a massive anode whose ends are closed by caps which make it impossible for electrons to leave the anode. If an anode in the form of a gauze is used as is often necessary in order to improve the heat radiation of the components within the anode, this shielding must also be extended behind the anode. For this purpose a wire gauze cage at cathode potential is suitable. Electrons are unable to pass through this cage to the outside (fig. 12*b*). Shields with a low potential may also be introduced in front of the much used mica supports at the extremities of the system in order to prevent their being struck by electrons.

AN APPARATUS FOR THE MEASUREMENT OF SCANNING SPEEDS OF CATHODE RAY TUBES

by L. BLOK.

621.317.087 : 621.317.755 : 711.3

A simple apparatus is described for the determination of the greatest scanning speed of cathode ray tubes which can be photographed. The light spot on the screen is allowed to describe a logarithmic spiral, at every point of which the speed can easily be determined. The point of least intensity on the spiral which is still sufficiently visible in the photograph determines the maximum scanning speed required.

If, with the help of a cathode ray tube, phenomena are investigated which are not periodic or stationary, but only occur once, it will in general be found practical to record the phenomenon photographically. In the most commonly occurring cases the details cannot be adequately studied by visual observation.

As examples of phenomena which are investigated in this way we may mention the switching on of transformers, break-down in the testing of insulating materials, back-surge in rectifiers, atmospheric discharges, etc.

In the photographic recording of such a phenomenon the blackening of the light-sensitive plate will depend on:

a) the brightness B of the cathode ray light spot,
b) the diameter of the spot D,
c) the velocity v at which the spot moves,
d) the optical enlargement N,
e) the light sensitivity η of the photographic material.

A detailed treatment of the relation between blackening and the factors a) to e) has already been published in this periodical [1]), together with a discussion of measurements relating to the maximum scanning speed. As a control in the manufacture of cathode ray tubes it is desirable to be able to determine this scanning speed in a simple way. In the following a description is given of a measuring apparatus solved for this purpose.

Principle

In order to determine the scanning speed by means of a single photograph it is necessary to allow the spot to describe a path on which the velocity changes continuously, while at every point of the path the velocity must be known with sufficient accuracy. One may then determine on the photograph the point where the path of the spot

[1]) J. F. H. Custers, The recording of rapidly occurring phenomena with the aid of the cathode ray tube and the camera, Philips techn. Rev. 2, 148, 1937.

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is still just sufficiently visible. The velocity at that point is then the maximum scanning speed which can be photographed under the given conditions.

A very simple solution of the problem is obtained when the light spot describes a logarithmic spiral with a constant angular velocity.

This curve may be represented in polar coordinates by the following equation:

$$\varrho = A\, e^{-a\varphi} \quad (1)$$

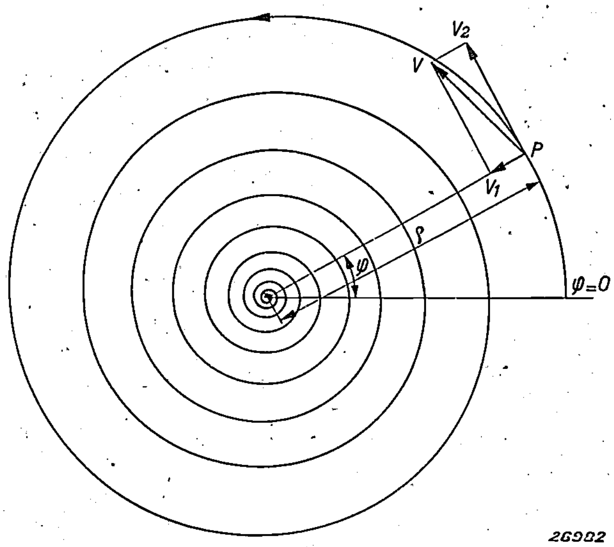


Fig. 1. Logarithmic spiral: $\varrho = A\, e^{-a\varphi}$

The exponent of e is taken negative so that with increasing argument φ the radius vector ϱ decreases (*fig. 1*). If φ is made linearly dependent on the time by setting

$$\varphi = \omega t \quad (2)$$

where ω is a suitable angular velocity we obtain:

$$\varrho = A\, e^{-a\omega t} \quad (3)$$

when $t = 0$, $\varrho = A$, when $t = \infty$, $\varrho = 0$.

The velocity at any point P of the spiral (*v.* fig. 1) is given by:

$$v = \sqrt{v_1^2 + v_2^2},$$

where v_1 is the radial velocity and v_2 the velocity perpendicular to the radius vector. We now have the relation:

$$v_1 = \frac{d\varrho}{dt} \text{ and}$$

$$v_2 = \varrho \frac{d\varphi}{dt} = \varrho\, \omega\,.$$

From equation (3) we find that:

$$v_1 = \frac{d\varrho}{dt} = -a\, \omega\, A\, e^{-a\omega t} = -a\, \omega\, \varrho\,,$$

so that:

$$v = \sqrt{a^2\omega^2\varrho^2 + \varrho^2\omega^2} = \varrho\, \omega \sqrt{a^2 + 1} \quad (4)$$

If a^2 is made small with respect to unity, the radial velocity v_1 becomes small with respect to the velocity v_2, and the distance between successive coils of the spiral becomes small with respect to the radius ϱ. Then for the velocity v we obtain by approximation:

$$v = \varrho\, \omega \quad (5)$$

This expression gives the velocity of a point which describes a circle with an angular velocity ω and a radius ϱ. The approximation employed therefore comes down to this, that the number of windings of the spiral must be so great that an element of the curve may be considered as an element of a circle.

Application of principle

In order to find out how a spiral can be obtained on a cathode ray tube by an electrical method, we pass over to rectangular coordinates.

For this purpose we take:

$$x = \varrho \cos\varphi = Ae^{-a\varphi}\cos\varphi\,,$$
$$y = \varrho \sin\varphi = Ae^{-a\varphi}\sin\varphi\,.$$

With $\varphi = \Omega t$ this becomes:

$$x = A\, e^{-bt}\cos\omega t = A\, e^{-bt}\sin\left(\omega t + \frac{\pi}{2}\right),$$
$$y = A\, e^{-bt}\sin\omega t\,,$$

where $a\omega$ is set equal to b.

x and y as functions of the time t have the form of two damped oscillations which differ in phase by 90°. Since the deflection of the spot on the screen is proportional to the potential difference between the plates, it is sufficient to apply a damped sinusoidal oscillation to each set of plates differing in phase by 90°.

Such a voltage is obtained in a simple manner by giving an LCR circuit an electrical impulse, and allowing it to die out in oscillations of its own frequency according to the well-known equation:

$$V_t = V_0\, e^{-\frac{R}{2L}t} \cos\omega t,$$

where V_0 is the voltage on the circuit due to the impulse at the moment $t = 0$, V_t the voltage at the time t, R the series resistance, L the self-inductance of the circuit, $R/2L$ the damping factor and $\omega = 2\pi f$, in which f is the frequency of the circuit itself.

By applying this voltage to the primary winding P of a transformer which has two separate equal secondary windings S_1 and S_2 (*fig. 2*), two similar secondary voltages are obtained of the same form as the primary voltage.

In order to obtain the desired phase difference, S_1 is loaded with the resistance R_1 and the con-

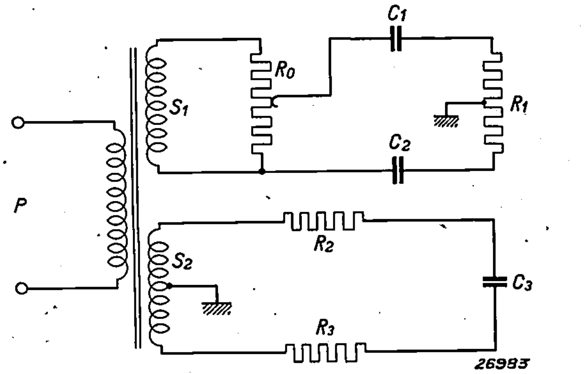
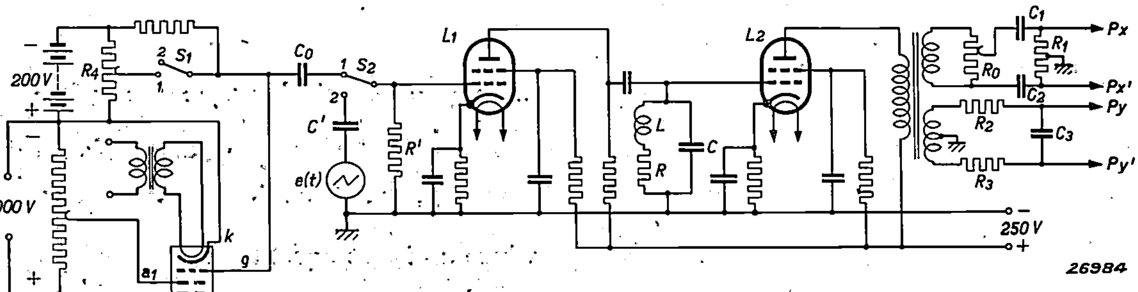


Fig. 2. Transformer with two similar symmetrically loaded secondary windings. The voltages on R_1 and C_3 differ 90° in phase and are symmetrical with respect to earth.

densers C_1 and C_2, and S_2 with R_2, R_3 and C_3, which are chosen so large that:

$$\frac{1}{\omega\, C_1} + \frac{1}{\omega\, C_2} = R_1 \quad \text{en} \quad R_2 + R_3 = \frac{1}{\omega\, C_3},$$



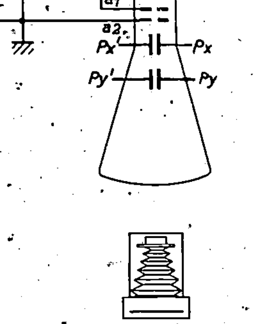


Fig. 3. Circuit diagram for the determination of the scanning speed. L_1 is a pentode AF 7, L_2 a pentode AL 5.

while in addition it is provided that:

$$R_1 = 2\, R_2 = 2\, R_3 \text{ and}$$
$$C_1 = C_2 = 2\, C_3 .$$

In this way the voltage V_{R1} on R_1 is made equal to the voltage V_{C3} on C_3, while both are symmetrical with respect to earth and, moreover, the phase difference between V_{R1} and V_{C3} is exactly 90°.

With a cathode ray tube the proportionality factor between voltage and deflection is not the same for both pairs of deflection plates. The spiral obtained will therefore be "oval". By applying the voltage V_{R1} which can be regulated with the potentiometer R_0 to the more sensitive plates, this difference may be compensated and the spiral becomes "circular".

Description of the circuit

The diagram of the circuit may be seen in *fig. 3.* The positive pole of the high voltage apparatus which supplies the anode voltage is earthed. When the switch S_1 is in position *1*, with the help of potentiometer R_4 the voltage of the grid g of the cathode ray tube can be regulated, and thereby the strength of the electron current. When S_1 is in position *2*, g becomes about 200 volts more negative than the cathode k and the current of electrons is thereby suppressed.

With switch S_2 in position *1*, on changing S_1 from *2* to *1*, the adjusted electron current begins to flow and a voltage impulse occurs on the grid of valve L_1. The oscillating circuit LRC thereby receives an impulse so that on the primary winding of the transformer in the anode circuit of L_2 the desired damped sinusoidal oscillation occurs. There is a potential of more than 2 000 volts on condenser C_0 which must be constructed accordingly.

For good adjustment of the camera it is desirable that it be focussed upon the spiral itself. For this purpose it is also necessary to be able to obtain a permanent image of the spiral on the screen of the cathode ray tube. To do this, a sawtooth voltage generator $e(t)$ is connected over the condenser C' by means of switch S_2 in position *2*. When R' and C' are correctly chosen, this sawtooth voltage $e(t)$ gives periodic impulses to L_1 with a frequency depending on the frequency of the sawtooth voltage. For the circuit in which this voltage works the following holds:

$$e(t) = V_{C'} + V_{R'},$$

in which $V_{C'}$ is the voltage on the condenser C' and $V_{R'}$ that on R'.

Upon substituting $V_{C'} = \int \frac{i}{C'}\, dt$ where i is the prevailing current and $i = V_{R'}/R'$, we may write:

$$e(t) = \int \frac{V_{R'}}{R'C'}\, dt + V_{R'}, \quad \text{or, differentiating,}$$

$$\frac{d\, e(t)}{dt} = \frac{V_{R'}}{R'C'} + \frac{d}{dt} V_{R'}.$$

If $C'R'$ is sufficiently small, the first term of the right hand member of the equation dominates, and we may safely write the voltage derivation in approximation:

$$\frac{d\, e(t)}{dt} = \frac{V_{R'}}{R'C'}$$

The voltage on R' and therefore supplied to the grid of L_1 is proportional to the derivative of the voltage $e(t)$ with respect to time.

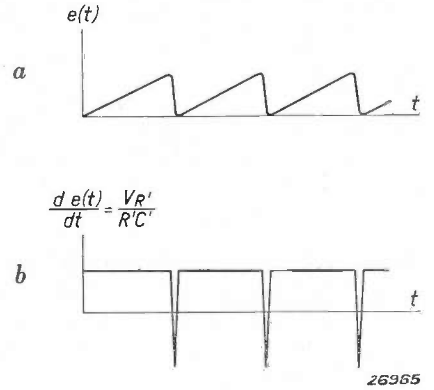


Fig. 4. Form of the sawtooth voltage $e(t)$ and of its derivation de/dt as a function of the time.

The form of the voltage $e(t)$ as a function of the time is indicated in *fig. 4a*, while in fig. 4*b* the variation of $V_{R'}$ as a function of t is drawn for sufficiently small values of $C'R'$. The latter figure shows plainly the occurrence of periodic impulses. Each of these impulses causes the spiral to be described once. In this way a permanent image appears on the screen of the tube. If the time between two impulses is made shorter than that necessary for the almost complete fading of the oscillations of the circuit, the light spot can no longer reach its point of rest. It then does not form a point in the center, but ends at an arbitrary point on one of the windings. An interesting effect is obtained when the frequency of the impulses is continously increased. The spiral is then seen to be erased from the centre outwards.

In recording a spiral the method is as follows. S_1 is set in position *1*, S_2 in position *2* and the camera is focussed on the spiral on the screen. After the lens has been closed the plate or film is introduced. S_1 is then set on *2* and S_2 on *1*, the shutter is opened and S_1 switched back to *1*, which causes the spiral to be described once on the screen. Then S_1 is returned to position *2*. *Fig. 5* shows a photograph taken in the way described with a Philips cathode ray tube DG 16 with a green-fluorescing screen. The scanning velocity reached amounts to (formula (5)): $v = \varrho \cdot 2\,\pi f = 3.5 \times 2\,\pi \times 1\,000 = 22\,000$ cm/s $= 220$ m/s. If the photograph is taken smaller than the true size of the image on the screen, for instance six times reduced, then, *ceteris paribus*, the maximum scanning speed which may be photographed is about three times as great as in a photograph in natural size.

In measurements carried out on this principle the maximum scanning speed can be determined as a function of the anode voltage and the electron current with all types of tubes. Moreover, this method presents a means on the one hand of choosing the most sensitive photographic material in connection with a given fluorescent substance on the screen of the cathode ray tube, and on the other hand of comparing the quality of different fluorescing substances.

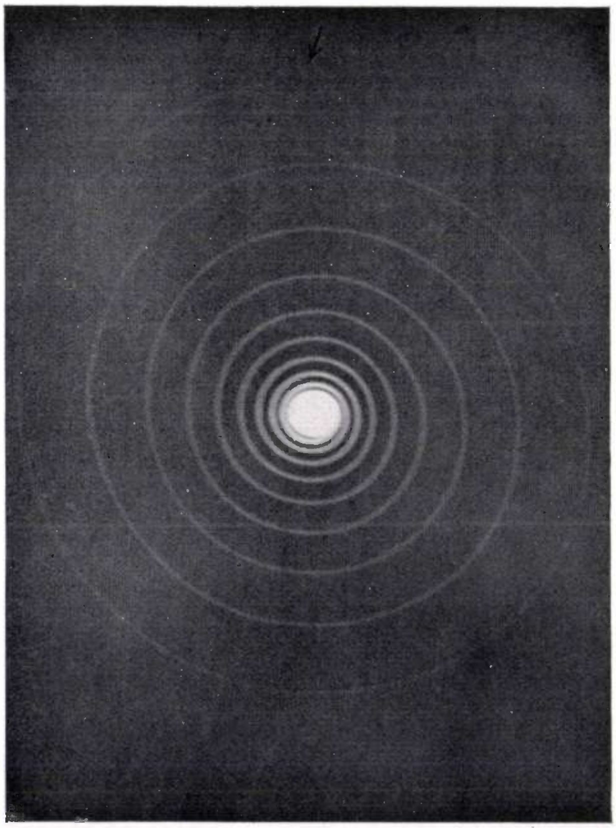


Fig. 5. Photograph for the determination of the scanning speed of a cathode ray tube DG 16. Electron current 50 μA, anode voltage 2 000 volts, ratio of size of photograph to object 1 : 1, aperture of lens 1 : 6.3, Agfa Isochrom plate. $\omega = 2\,\pi f = 2\,\pi\, 1\,000$. The spiral is still just visible for $\varrho = 3.5$ cm. Maximum scanning speed 220 m/s. The enlargement of the reproduction is 1 : 3.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS GLOEILAMPENFABRIEKEN

1286*: H. C. Hamaker: Potential curves as a basis for the discussion of colloidal phenomena (Chem. Weekbl. 35, 47 - 57, Jan. 1938 [1]).

For the contents of this lecture the reader may be referred to the original publications **1154**, **1174** and **1211**. In particular this article gives a more general formulation of the problem treated in **1174**.

1287*: E. J. W. Verwey: Double layer and stability of lyophobic colloids (Chem. Weekbl. **35**, 70 - 79, Jan. 1938 [1]).

On the basis of observations on silver iodine sol the stability of colloids is described as due to the electrical double layer present on these colloids.

1288: N. W. Timoféeff-Ressovsky, K. G. Zimmer and F. A. Heyn: Auslösung von Mutationen an Drosophila melanogaster durch schnelle Li + D-Neutronen (Naturwissenschaften **26**, 108 - 109, Feb. 1938).

A lithium plate bombarded by rapid deutons emits a large number of rapid neutrons. These are used to caused mutations in flies.

1289: W. G. Burgers and J. J. A. Ploos van Amstel: Electron-optical observation of transition of α- into β-zirconium (Nature **141**, 330, Feb. 1938).

Electron-microscopic observations or zirconium are found to give results in agreement with the conclusion drawn from X-ray photographs, that the transition between α and cubic β-zirconium takes place by a homogeneous slipping and stretching (or compression) of connected lattice regions.

1290: W. G. Burgers and J. J. A. Ploos van Amstel: Lecture-room demonstration of electron-optical crystal patterns (Nature **141**, 370, Feb. 1938).

A description is given of the way in which an electron microscope can be constructed which is so powerful that the picture of the crystal structure of the metal surface produced on the fluorescent screen can be projected optically on a screen 50 by 50 cm, upon which a picture is then obtained which is suitable for demonstration in a lecture room.

1291: M. J. O. Strutt: Die charakteristischen Admittanzen von Mischröhren für Frequenzen bis 70 Megahertz (El. Nachr. Techn. **15**, 10 -17, Jan. 1938).

The most important properties to be measured on frequency changing valves (octodes and hexodes) are discussed, followed by a treatment of a circuit and a measuring arrangement which can be used to about 1.5 megacycles/sec.

Input and output admittance and conversion slope are determined for octodes type AK 2 and hexodes type ACH 1, as well as the admittance between first and fourth grids of the octode. Special attention is given to the shielding of the measuring arrangement for short waves. Measuring results to 50 megacycles/sec. are given. Finally measurements and calculations on the induction effect in octodes are discussed.

1292: A. A. Kruithof and F. M. Penning: Contribution of the photoelectric effect to the sparking mechanism in the noble gases at high pressures (Physica **5**, 203 - 204, Mar. 1938).

The ratio of the number of excitations to the number of ionizations caused by electrons in a pure gas increases sharply when the quotient of the electrical field strength E divided by the pressure p decreases. At values of E/p which are considerably lower than that of the minimum breakdown voltage, an increase in the number of electrons freed from the cathode per positive ion was found for various rare gases. This increase becomes understandable if in the breakdown mechanism account is taken of the freeing of electrons from the cathode by radiation.

1293: J. L. H. Jonker and A. J. W. M. van Overbeek: The application of secondary emission in amplifying valves (Wirel. Eng. **15**, 150 - 156, Mar. 1938).

The use of secondary emission for amplification in radio valves is treated in detail. The same subject was discussed more briefly in the May number of this periodical (Philips techn. Rev. **3**, 137, 1938).

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*) An adequate number of reprints for the purpose of distribution is not available of those publications marked with an asterisk. Reprints of other publications may be obtained on application to the Natuurkundig Laboratorium, N.V. Philips' Gloeilampenfabrieken, Eindhoven (Holland), Kastanjelaan.

[1]) 1286* and 1287* are lectures given at a symposium on the dynamics of hydrophobe suspensions and emulsions which was held on Nov. 5th and 6th 1937 in Utrecht. All the lectures given in this symposium, as reprints from the Chem. Weekblad, have been collected into a booklet which is published by Centen in Amsterdam at a price of f 4.— ($ 2.25).

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SOUND AMPLIFICATION AND PUBLIC ADDRESS

by J. DE BOER.

534.846.4

In both the open air and auditoria with defective acoustics the intelligibility can be improved by means of electrical amplifying equipment. The location of loudspeakers and microphone must be given careful thought in order to avoid undesirable conditions arising in addition to the intensification of the sound, such as an amplification of disturbing noises and an increase in the reverberation time. In many cases the acoustics can only be improved by means of an electrical amplifying equipment using microphones and loudspeakers with pronounced directional characteristics.

The introduction of suitable electrical amplifying equipment during the last few years has made it possible for the spoken word and music to be clearly heard at remote points where previously sound was inaudible owing to insufficient sound intensity. An electrical equipment for amplifying the sound waves and thus increasing the range of audibility and intelligibility consists essentially of three components:

1) The microphone, which converts the sound vibrations into electrical alternating voltages, embracing exactly the same frequencies as the sound vibrations themselves and each with an amplitude proportional to the amplitude of the initial atmospheric vibration;
2) the amplifier, which amplifies the alternating voltages in the requisite ratio, and
3) the loudspeaker which again converts the amplified alternating voltages into sound vibrations.

The applications of sound-amplifying equipment can be conveniently subdivided into two main groups:

A) For use where the speaker and audience are in the same auditorium, but the speaker's voice is not sufficiently intelligible to the whole audience without electrical amplification. In this case the electrical equipment serves in some measure to improve the acoustics of the auditorium. Sound amplification in the open air also comes under this group.
B) For use where the speaker is not in the same room as his audience, so that the electrical equipment acts as the direct source of sound in the auditorium.

In the present article we shall discuss mainly the first group of applications, *viz.*, the improvement of acoustic conditions and particularly the intelligibility of speech by means of loudspeakers.

When is sound amplification required?

An improvement of the acoustics by electrical amplifying equipment is required when satisfactory acoustic conditions cannot be achieved by other means, such as a suitable acoustic design of the auditorium or a correct choice of the wall and floor coverings, or where these measures cannot be taken for practical reasons, *e.g.* in an auditorium of too great a volume or in the open air. The problem can be conveniently discussed by a more detailed consideration of the following aspects:

1) The acoustics of an auditorium are satisfactory although its volume is too large, with the result that the natural sound intensity is insufficient for clear intelligibility or the source of sound is too feeble.
2) The geometrical shape of the auditorium does not give a uniform distribution of sound.
3) The auditorium is planned on the correct acoustic lines, but its time of reverberation is too high.
4) The audience is in the open air.

Sound amplification in an auditorium with good acoustics

If the auditorium is of the correct geometrical form so that the sound is uniformly distributed, and it has also a reasonable time of reverberation, the intelligibility can still be unsatisfactory if, owing to the size of the auditorium or disturbing extraneous noises, the mean sound intensity is too low compared with the extraneous noise. By increasing the emission of sound from the source, satisfactory and adequate intelligibility would be obtained in this case.

We must examine how far it is possible to produce a more powerful source of sound by means of electrical amplifying equipment, assuming that the microphone is located in the centre of the auditorium in which the sound is uniformly distributed, and that the loudspeaker is placed close to the speaker.

It is first necessary to decide what effect an amplifying equipment of this type has on the acoustic characteristics of the auditorium. If the loudspeaker has no pronounced directional radiation, the sound distribution is not altered by the provision of an amplifying equipment, which indeed is quite unnecessary if the auditorium has been planned on efficient acoustic lines. But if an amplifying equipment is installed in an auditorium of this type the reverberation time will be automatically altered, and it will indeed be increased in all cases, as may be gathered from the following considerations: The sound intensity in an auditorium is, according to Sabine's theory, independent of the location of the sound source, and in the case in point with an auditorium in which the sound is uniformly distributed this theory naturally applies. The acoustics of the auditorium will therefore not be altered if the amplifying equipment is replaced by a surface which absorbs as much sound as impinges on the microphone and also radiates as much sound as the loudspeaker. This surface will thus have a negative absorption factor. The magnitude of this coefficient depends on the degree of sound amplification, the conditions here being exactly the same as when sound is amplified by reflection. As the reflected sound increases in intensity, the time of reverberation also increases. In fact, with an amplifying equipment the total absorption of sound by the walls of the auditorium can be reduced to zero. The sound waves will then no longer decay; this phenomenon is termed acoustic reaction.

If, as we have assumed, the sound intensity is still inadequate when the correct time of reverberation has been obtained, it is apparent from the above arguments that the installation of amplifying equipment will not result in any improvement. It is fortunate that in practice conditions are not as unfavourable as this. We have assumed that the sound intensity at the place of the microphone is equal to the mean intensity of sound in the auditorium. If the microphone is placed sufficiently close to the speaker, the intensity of sound reaching the microphone can be made much greater than the average sound intensity in the auditorium, with the result that the gain of the amplifier and hence also the negative absorption factor of the equipment can be reduced. With the same mean sound intensity the time of reverberation can again be slightly reduced by adopting this method, while at the same time the extraneous noises in the auditorium are less amplified than the speaker's voice.

A further improvement can be achieved by using a microphone with a directional characteristic, *i.e.* a microphone with a powerful response to sound waves travelling from the direction of the speaker, while its average response (the mean response calculated for all directions) is much smaller. The average response is, however, a measure of the negative absorption coefficient, as reverberant sound impinges on the microphone uniformly from all directions.

Directional Characteristics of Microphones

A microphone, which responds to the acoustic pressure, has the same response for all sound waves impinging on it irrespective of direction, if its dimensions are smaller than or of the same order of magnitude as the wavelength of the sound pulses. A pressure-gradient microphone possesses different response characteristics and reacts to the periodic difference in pressure between two adjoining points, for instance on the two sides of a diaphragm. A sound pulse whose direction of propagation is perpendicular to the line joining these two points produces no difference in pressure between them, and the microphone will therefore not respond to this sound pulse. If the direction of propagation makes an angle φ with this line the microphone will respond only to the component of the differential pressure along this line, the induced potential being thus proportional to $\cos \varphi$. In a directional diagram (*fig. 1*), the response of a microphone with these directional characteristics is represented by two circles making contact at the origin.

When using a combination of a pressure and a pressure gradient microphone, both having the same maximum response, the induced potential is proportional to: $^1/_2\,(1 + \cos \varphi)$. This directional effect is represented by a cardioid, which is also shown in fig. 1.

According to definition, the power response of a microphone is proportional to the square of the potential induced in it by a given acoustic field. If an average is struck of the response from all directions in space, it is found that the average response values in the three directional diagrams discussed above are in a ratio of $1 : {}^1/_3 : {}^1/_3$ [1]), while the maximum response (at $\cos \varphi = 0$) is the same in all three cases. With the two last-mentioned types of microphone the negative absorption coefficient is thus three times smaller than in the pressure microphone, if the sound amplification is the same in all cases.

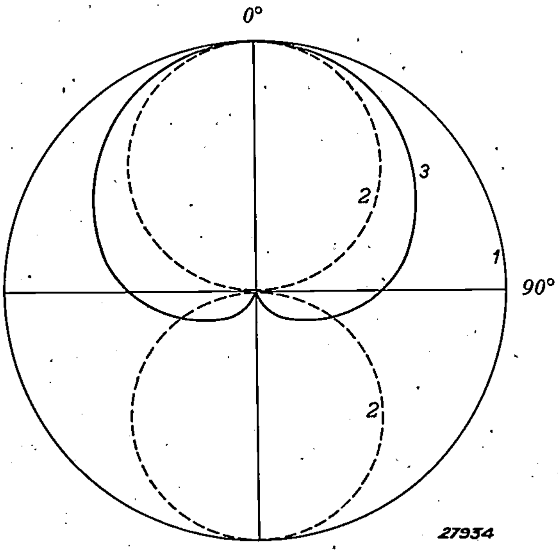

Fig. 1. Directional diagram of microphone characteristics, with the potential at the microphone at a given pressure amplitude plotted as a function of the direction of the sound pulse. *1.* Pressure microphone; *2.* pressure-gradient microphone; *3.* combined microphone (${}^1/_2$ pressure microphone + ${}^1/_2$ pressure-gradient microphone).

Position of loudspeakers

If the auditorium is designed such that the sound is uniformly distributed over it, the loudspeakers can be located at any convenient points. But since the ear can always detect the source of sound in spite of the equalisation of sound distribution produced by reflection, it is desirable to place the sound projectors close to the speaker. The perception of the direction from which a sound pulse is radiated is mainly due to the position of the ear relative to the direction from which the sound waves first impinge on it, *i.e.* the direction of propagation of the direct waves. If the wavelength of the sound pulses is of the same order of magnitude or smaller than the dimensions of the head (*i.e.* in the case of notes with a frequency above about 300 c/s), the sound intensity impressed on the two ears is not the same. The difference in intensity which depends on the direction of the incident sound is a measure of the directional impression. The time interval elapsing between the impression of sound on the two ears also affects the directional sensitivity, especially with low notes for which refraction is negligible.

Differences in sound intensity and in the time of arrival of the sound waves occur only when the source of sound is displaced horizontally out of the plane of symmetry containing the two ears, but not when it is displaced in a direction perpendicular to this plane. The latter displacement is hence much less perceptible, a point which must be borne in mind when deciding on where to place the loudspeakers. The latter should as far as possible be placed above the speaker and not to his side. Also, the distances between the individual projectors and between them and the speaker should not be too great, as otherwise they will be heard independently of each other instead of simultaneously. Their distances from the audience should not differ by more than about 60 ft., for no time difference in the arrival of sound should exceed about ${}^1/_{17}$ sec.

In the majority of auditoria the sound intensity is not entirely independent of the positions of the loudspeakers, for the audience at the back of long halls will usually receive a greater volume of sound the higher the loudspeakers are placed. (A sound wave which travels directly above the heads of an audience is considerably enfeebled by acoustic absorption). A compromise must therefore be made here between the need for placing the loudspeakers as close as possible to the speaker and the need for obtaining satisfactory intelligibility in the rearmost rows of seats.

Sound intensity

Intelligibility is determined by the ratio between the sound intensity and the noise level or reverberation; this relationship has been shown diagrammatically in Philips techn. Rev., 3, 139, 1938, fig. 2. If the sound intensity is 15 decibels higher than the noise level, the intelligibility for nonsense syllables is still 70 per cent and simple sentences can be clearly understood. In the absence of disturbing noises the sound intensity must not drop below about 30 phons, otherwise the intelligibility immediately drops considerably.

The relationship between the loudness level L in decibels above the threshold value, for which a value of 10^{-10} micro-watts per sq. cm. is assumed, and the output N of the source of sound is given by the expression:

$$N = \frac{V}{T}\, 10^{\frac{L-73.8}{10}},$$

[1]) If $R(\varphi)$ is the potential induced at an angle of incidence, $d\Omega$ the solid angle between φ and $\varphi + d\varphi$, then the mean response apart from a factor of proportionality is:

$$\frac{1}{4\pi}\int R^2(\varphi)\, d\Omega = {}^1/_2 \int_0^{\pi} R^2(\varphi) \sin\varphi\, d\varphi.$$

For the pressure microphone, the pressure-gradient microphone and the combined microphone $R(\varphi) = 1$, or $\cos\varphi$, or ${}^1/_2\,(1 + \cos\varphi)$ respectively.

where V is the volume of the auditorium in cub. m.,
T is the time of reverberation in sec.,
W is the output of the source of sound in microwatts,
L is the loudness level above 10^{-10} microwatts per sq. cm. This formula can be readily derived from the relationship between the output of a sound source and the energy density given in an earlier article (A. Th. van Urk, Philips techn. Rev., 3, 65, 1938, *cf.* p. 70). A nomogram is given in *fig. 2* for the relationship between N and L, which indicates that with the normal sound intensity of the human voice (N = 10 microwatts) a loudness level of 45 phons (above the threshold of audibility) can be obtained in an auditorium having a volume of 10000 cub. m and a reverberation time of 1.5 sec. This is just sufficient to understand every syllable the speaker utters, provided the interference level does not exceed 30 phons. With a higher noise level electrical amplifying equipment must be employed.

The required output of the electrical amplifying equipment can be calculated by adding 15 decibels to the noise level of the auditorium which gives the

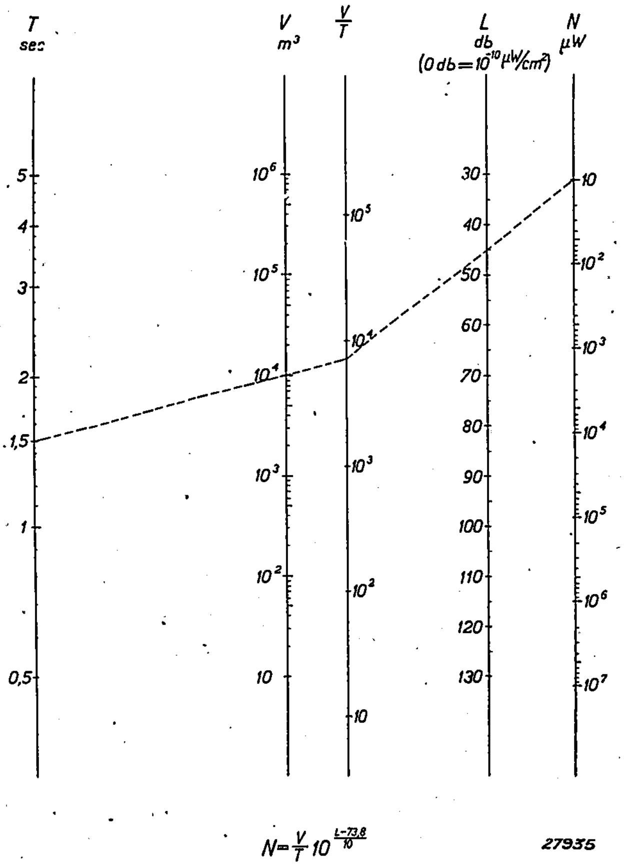

Fig. 2. Nomogram $N = \frac{V}{T}\, 10^{\frac{L - 73.8}{10}}$.

L = Sound intensity in decibels above 10^{-10} microwatt per sq. cm.
N = Output of sound source in microwatts.
T = Reverberation time in seconds.
V = Volume of auditorium in cub. m.

sound-intensity level required; the output which the loudspeakers together with the speaker must furnish is then given by fig. 2. If the efficiency of the projectors is known the average power output of the amplifier can be deduced therefrom. The maximum undistorted output of the amplifier must be about 15 decibels higher [2]) than the mean value calculated in this way.

Sound amplification in buildings with defective acoustics

The production of echo and the focussing of sound waves in halls as a result of irrational planning and unsuitable dimensions have already been discussed in a previous article [3]). If the auditorium is too large so that an amplifying equipment must be used in any case owing to the inadequate sound intensity, these deficiencies in the design of the auditorium can to some extent be rectified by selecting suitable positions for the loudspeakers. In this way the amount of sound directed against walls responsible for undesirable reflection can be reduced to a minimum. Yet a really effective improvement in these cases is generally only possible by altering the plan of the auditorium or by covering the troublesome reflecting surfaces with sound-absorbent materials.

Another troublesome feature frequently found in churches, which is also due to the geometrical design of the auditorium and which can also be considerably remedied by means of an amplifying equipment, is the presence of many thick pillars in old churches which seriously obstruct the propagation of the high notes of direct sound waves. If the wavelength of the sound pulses is large compared to the dimensions of the pillars (lowest notes), the latter will not interfere with the transmission of sound. But if this ratio is small, as with high notes, the sound waves will be refracted, and in fact with very short waves a zone of silence may even be obtained behind the pillars. To illustrate how a zone of silence of this character is produced, the distribution of sound pressure about a spherical obstruction is shown in *fig. 3*. The ratio of the radius a to the wavelength λ is a measure of the refraction. The appearance of zones of silence of this character can be avoided by placing additional loudspeakers behind the pillars.

Another factor adversely affecting audibility, also frequently found in churches, is that the hearers

[2]) F. Trendelenberg, Klänge und Geräusche, J. Springer, Berlin, 1935. p. 92.

[3]) R. Vermeulen. Auditorium acoustics and intelligibility, Philips techn. Rev., 3, 139, 1938.

are located both in front and to both sides of the preacher. The intensity of the sound waves is not the same in all directions owing to the refraction of sound waves round the speaker's head; the differences in intensity in different directions are greater for the higher frequencies than for the lower ones. Measurements of the directional distribution of the mean sound pressure of the human voice are reproduced in *fig. 4*, which shows that the intensity behind the speaker is about 14 decibels lower than straight ahead. Furthermore, the preacher frequently occupies a pulpit closed at the rear which intensifies the screening effect in this direction. An improvement in audibility can also be obtained here by placing a loudspeaker behind the pulpit.

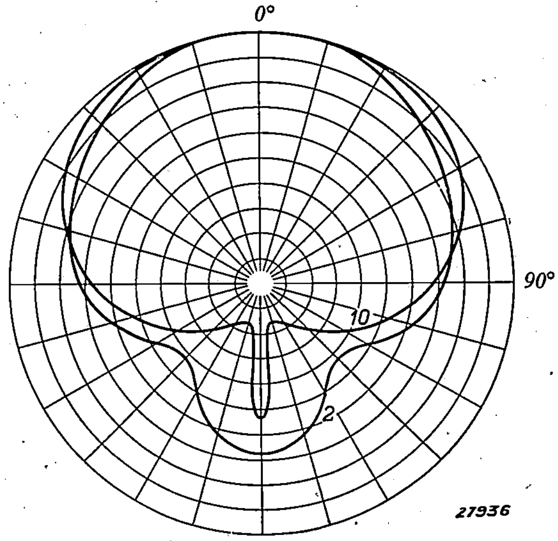

Fig. 3. Zone of silence behind an obstruction. Distribution of the sound pressure produced by a plane sound wave of wavelength λ at the surface of a sphere. The figures against the different curves are the values of the product ka, where $k = 2\pi/\lambda$ and a is the radius of the sphere.

The screening effect of the head can be calculated by replacing it by a sphere with a point source of sound located at its surface. Instead of calculating the sound pressure at a considerable distance in front of the sphere as a function of the direction, the pressure in a specific direction can be calculated as a function of the position of the source of sound on the surface of the sphere. But this relationship is identical with the distribution of the sound pressure at the surface of the sphere when the source of sound is far removed; this distribution of sound pressure has already been given in fig. 3. The reciprocal theorem employed here states in general terms that the sound pressure remains the same when the source of sound and the point at which the pressure is measured are interchanged.

Sound amplification in auditoriums with an excessive reverberation time

A discussion of the dependance of intelligibility on the ratio of the "useful" sound (*i.e.* the sound which reaches the hearers within an interval of $^1/_{17}$ of a sec.) to the reverberation has already been given in a previous article (Vermeulen, *loc. cit.*). If this ratio is too small the speaker's utterances become unintelligible. Knudsen has analysed the connection between intelligibility and

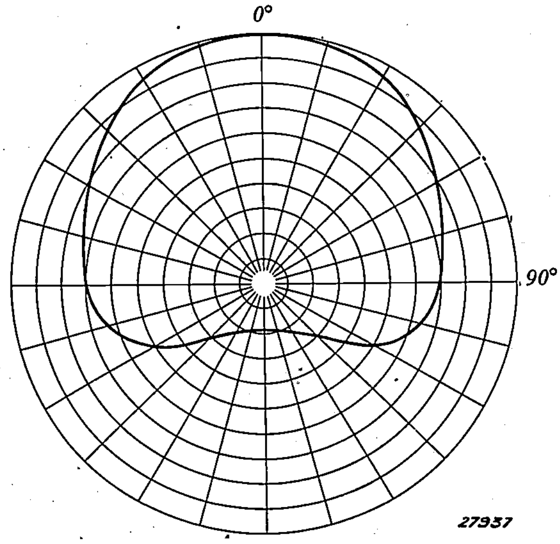

Fig. 4. Directional distribution of the mean sound intensity of the human voice according to O. Zwikker (Ingenieur, 44 E, 39, 1929). The curve shows the effective acoustic pressure in different directions. The acoustic pressure is about 14 decibels lower behind the speaker's head than directly in front.

the reverberation time, his results being reproduced in *fig. 5*. If the reverberation time exceeds 3 secs, the speakers's voice is as a rule unintelligible, and to obtain an improvement in intelligibility the acoustic pressure of the direct sound pulses must be increased but without a comparable increase in the reverberation. This may be realised by using loudspeakers with pronounced directional characteristics and which direct the total sound energy in the direction of the audience. The reflection of

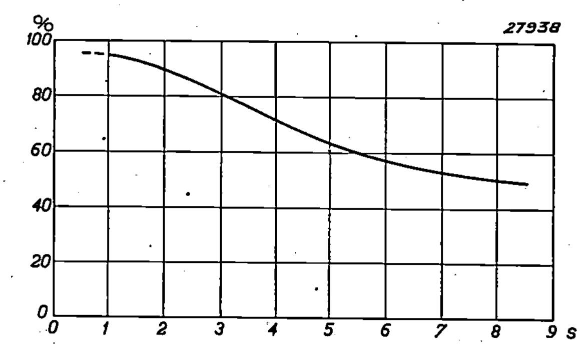

Fig. 5. Intelligibility as a function of the reverberation time according to Knudsen.

sound at a surface occupied by an audience is almost zero, so that the reflected sound waves are not amplified and hence reverberation does not become more marked.

Directional characteristics of loudspeakers

Whether the sound waves are propagated in a specific direction or not depends on the ratio of the dimensions of the radiating surface to the wavelength of the radiated sound pulses, as well as on the shape of the diaphragm and its method of mounting. If the wavelength is large compared to the dimensions of the radiating surface, a spherical sound wave is emitted which has the same intensity in all directions. On reducing the wavelength and making it more of the order of magnitude of the dimensions of the diaphragm, a pronounced directive effect may be obtained; the sound is radiated principally in a direction perpendicular to the radiating surface, interference between sound waves radiated from different points of the diaphragm taking place in other directions.

The distribution of sound can be accurately calculated for a plane, circular radiating surface which moves as a whole and which is mounted in an aperture in an infinitely large baffle, being given by the expression:

$$\frac{p_a}{p_0} = 2\,\frac{J_1\,(ka\,\sin\,\alpha)}{ka\,\sin\,\alpha} \quad . \quad . \quad . \quad . \quad . \quad (1)$$

where $k = 2\pi/\lambda$; λ = wave length of the sound waves;

a = radius of the diaphragm;

p_a = sound pressure in a direction making an angle α with the normal to the surface of the diaphragm;

p_0 = sound pressure in the direction of the normal;

J_1 = first order Bessel function.

The directional distribution expressed by equation (1) has been plotted in *fig. 6* for various values of the parameter *ka*. This diagram shows that a satisfactory directional effect for a pulse of *e.g.* 500 c/s ($k = 0.1$ cm^{-1}) can only be obtained when a diaphragm 100 cm in diameter ($ka = 5$) is used. As a rule the diameter of a loudspeaker cone is only of the order of 20 cm and it is not practicable to make the diameter any larger. A larger diaphragm, which must also be sufficiently rigid to vibrate as a whole, would be very heavy, and this would detract from the efficiency of the loudspeaker as a whole.

The directional characteristics of a loudspeaker can however be improved in still another way, *viz.*, by placing the cone in a horn, the diameter of the radiating surface of the horn being larger than that of the cone so that a more pronounced directive action is realised.

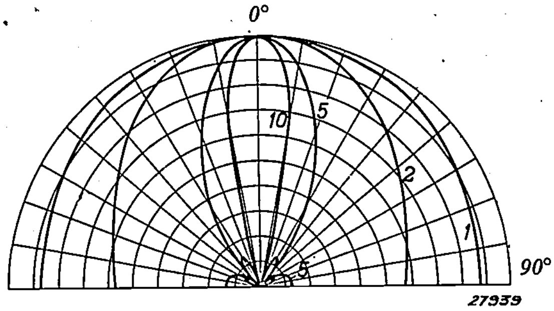

Fig. 6. Directional distribution of the sound intensity with a circular vibrating plate in an infinitely large baffle. The figures against the curves give the values of the parameter *ka* which determines the distribution.

If an attempt is made to use this method for diverting also the lowest audio-frequencies in a specific direction, the requisite dimensions of the horn would be so large that a practical design would prove entirely out of the question. If those frequencies which are not radiated in a specific direction by standard horns now in use are removed by an electrical filter in front of the loudspeaker input, satisfactory directional radiation can then be obtained for the whole sound output of the loudspeaker. The absence of the low frequencies will only slightly affect intelligibility (*cf.* Vermeulen, *loc. cit.*); if, for instance, all frequencies below 500 c/s are removed the intelligibility will still be of the order of 95 per cent. Yet, on the other hand, the filtering out of the low-pitched sound must not be taken too far, which would be obtained with hornless loudspeakers, as then the timbre would be altered too much and the audience become aware of a difference in timbre between the speaker and the sound radiated by the loudspeakers.

Fig. 7 shows the measured directional distribution of the radiated sound energy for a loudspeaker with a cone diameter of 22.5 cm, firstly with a baffle of 100 cm in diameter and, secondly, mounted in a horn with an aperture of 62 cm. It is seen that in the latter case the energy at frequencies above 500 c/s is radiated mainly within a cone of solid angle 60 deg. From the distance of the loudspaekers from the audience, the number of loudspeakers required for satisfactory diffusion of sound in a large auditorium can be immediately deduced. If the sound intensity level required is also taken into consideration, the acoustic energy which the loudspeakers must radiate can also be deduced.

The microphone should be placed as far as possible outside the beam of waves radiated from the loudspeakers. The use of a pressure-gradient microphone is an advantage here and it should be so placed that the lines joining the microphone and the loudspeakers lie outside the response zone of the microphone.

Another method of intensifying the direct sound pulses in an auditorium relative to the reverberation consists in using a large number of small loudspeakers which are located at various points in the audience. Owing to the proximity between the loudspeakers and the audience the sound radiated by the projectors is entirely absorbed by the audience without any of the loudspeakers requiring to have a pronounced directional action. When using this method, care must be taken that the hearers do not receive any sound pulses from loudspeakers farther than 60 ft away, as otherwise the time difference in the arrival of sound pulses coming from the speaker direct and that radiated by the loudspeakers will begin to be disturbing.

Sound amplification in the open air

In the open air no reverberation likely to reduce the intelligibility will occur, although the level of other disturbing noises is usually higher than in an auditorium (sound from stands and public enclosures, the audience as a rule being far noisier than in an auditorium); the sound intensity required in the open air for satisfactory intelligibility is thus generally greater than that needed in an auditorium. In many cases amplification of sound will therefore prove imperative. One way of doing this is by mounting a screen or sound reflector behind the speaker, although owing to the small dimensions of the reflectors used this arrangement is much less effective than reflection at the walls and particularly at the ceiling of an auditorium, and a really satisfactory improvement cannot be obtained by this means.

The intelligibility can be improved in this case by means of an amplifying equipment. If only a few large loudspeakers are to be used, their directional characteristics must be very good so that the available energy is utilised to the best advantage. When deciding the direction of the loudspeakers, reflection of sound at the walls of the dais or stand should be carefully excluded.

The wind plays an important part in sound amplification in the open air. The wind intensity increases in a direction perpendicular to the ground, so that sound pulses which are propagated in the same direction as the wind are deflected towards the ground. But if the waves are propagated in the opposite direction. *i.e.* against the wind, they are deflected upwards away from the ground. In the first case, the range of the sound waves is greater and in the second case smaller than in still air.

With a wind of normal strength, the distance from the source of sound at which the spoken word is still intelligible may be double in the direction of the wind than in still air and only half as great when the waves travel against the wind.

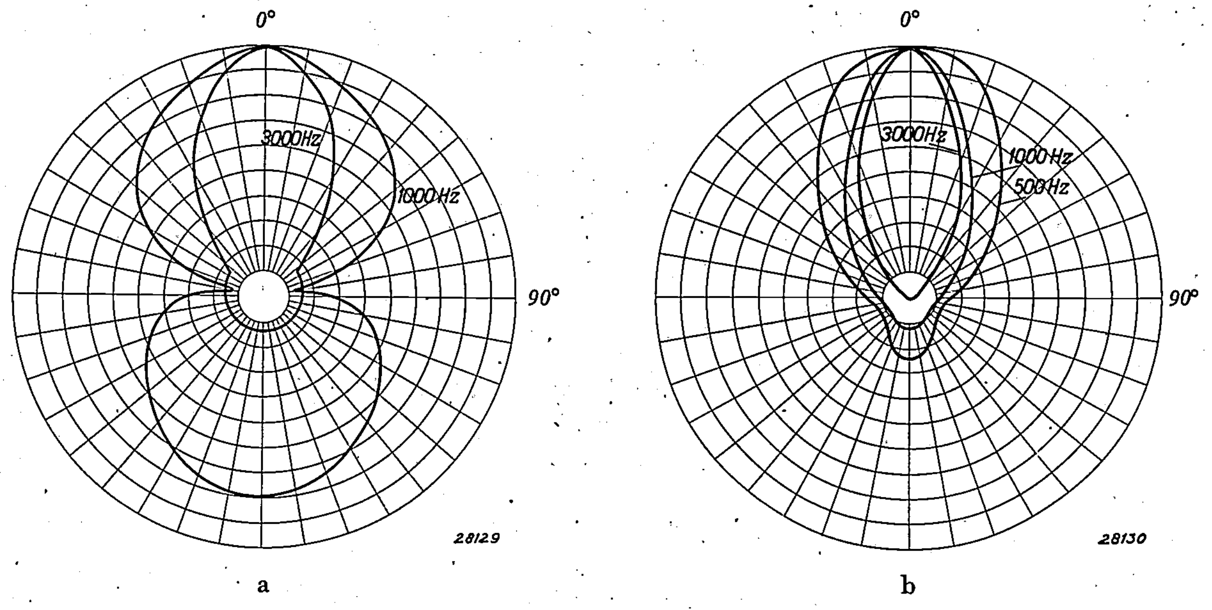

Fig. 7. Measured directional distribution of the sound pressure for a loudspeaker: a) when mounted in a baffle, and b) when mounted in a horn.

AUTOMATIC MACROSCOPIC EXAMINATION OF MATERIALS WITH X-RAYS

by J. E. DE GRAAF and J. H. VAN DER TUUK. 539.26 : 537.562

A method is described for the automatic detection of flaws, *e.g.* cavities and blowholes, in castings, etc., using ionisation chambers and an amplifying arrangement.

Introduction; General arrangement of ionisation chambers

The disadvantage of the method of radiographic examination of macrostructure previously discussed in a series of articles in this review [1]) is that the cost per object examined is comparatively high; it is sometimes even greater than the value of the object itself in the case of cheap mass produced components, owing to the high cost of the films used and the length of time occupied for each examination. Instead of making a permanent radiograph, it is also possible to study the screened image visually on a fluorescent screen, this method of examination being frequently used for testing aircraft components, which usually have thin walls or are made of light metals, such as aluminium and magnesium. The method is convenient to apply and is quick, but has various disadvantages for it is not objective, does not give a permanent record, and it necessitates the observers being frequently relieved as the work is very fatiguing.

A method of examination is described in the present article which has been developed in our Laboratory and serves for the automatic indication of faults and flaws, such as cavities in castings. It should be emphasised at the outset that the apparatus described has not reached a stage of development fitting it for general adoption in industry, for its primary purpose up to the present has been the investigation of the fundamental possibilities of the method.

As a cavity in a metal object represents a centre of reduced absorption, the intensity of the X-ray beam which passes through a cavity will be greater than that of the same beam after traversing a solid path throughout, so that a measurement of the difference in intensity serves to indicate the presence and location of a cavity. The intensity of an X-ray beam can be measured with an ionisation chamber, the current flowing through this chamber as a result of ionisation being a measure of the beam intensity. But if only one ionisation chamber is used small fluctuations in mains voltage will also be indicated, since the radiation transmitted by a thick metal wall is proportional to a very high power (5 or 10) of the voltage applied to the X-ray tube, *i.e.* of the mains voltage. The effect of these fluctuations can be eliminated by using the compensation circuit shown in *fig. 1.* As long as the intensities of the two X-ray beams are equal the currents flowing through the two ionisation chambers a and b will also be equal, the current through the high resistance R will remain zero and the p.d. between c and d will also be zero. If the intensities of the beams are unequal, as for instance when one beam passes through a flaw in an object under examination, a p.d. is obtained between c and d. This p.d. can be employed for actuating an alarm arrangement connected to terminals e and f.

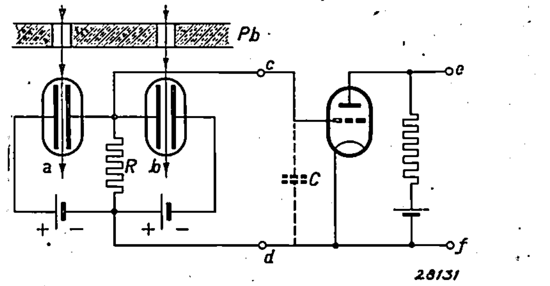

Fig. 1. Arrangement of two ionisation chambers a and b in a compensation circuit, enabling measurements to be made independent of simultaneous and equal fluctuations in the intensity of the two X-ray beams (fluctuations in mains voltage). A lead screen with two apertures for the passage of the beams is placed in front of the chambers.

The procedure of a test is as follows: The object under examination is moved between the ionisation chambers and the X-ray tube. In the case of a tubular object the best method is to rotate it about its axis and at the same time displace it along its axis. The X-ray beam thus scans the object along a helical path and any flaws present pass in front of the ionisation chamber at speed equal to the speed of revolution of the object.

This method can of course only be employed for detecting the presence of a flaw, a radiographic picture still being necessary to diagnose the cause, since only a photograph can give complete information regarding the shape, position and interrelationship of the flaws.

Sensitivity of the method

The difference in the currents through the two ionisation chambers and hence the p.d. between c and d (fig. 1) produced by a flaw in front of one of the ionisation chambers can be conveniently

[1]) Philips techn. Rev. 2, 314, 350, 377, 1937; 3, 92, 186, 1938.

calculated approximately when the X-rays are monochromatic. With an initial intensity I_0 and a thickness of material d or $d—\Delta d$ the intensities behind the object examined are (*fig. 2*):

$$I_1 = I_0\, e^{-\mu d}, \text{ resp. } I_2 = I_0\, e^{-\mu(d-\Delta d)},$$

where μ is the coefficient of X-ray absorption of the material. The ionisation currents produced in the ionisation chambers of length D by these X-ray intensities are:

$$i_1 = p\, I_1\, (1 - e^{-\mu' D}) \text{ resp. } i_2 = p\, I_2\, (1 - e^{-\mu' D}),$$

where μ' is the coefficient of absorption for X-rays of the gas in the ionisation chambers. The factor of proportionality p is partly determined by the wave-length λ of the X-rays, since with decreasing wave-length a diminishing portion of the energy primarily absorbed for the generation of a photo-electron is utilised for generating secondary electrons (ionisation). The ionisation chambers employed are in fact usually smaller than the range of the photo-electrons, produced by the absorption of the X-ray quanta, in the gas filling of the tube; these electrons thus reach the wall of the chamber before the whole of their energy has been consumed for ionisation and the unused fraction will be greater the greater the speed of the photo-electrons, *i.e.* the shorter the wavelength of the absorbed X-rays.

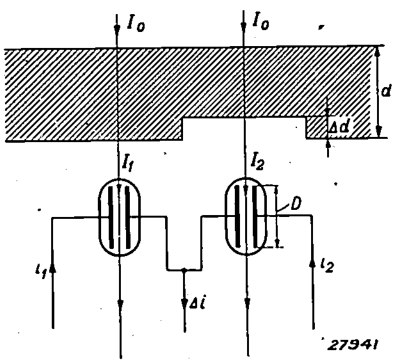

Fig. 2. Illustration of the calculation of the differential current. Δi due to a difference in thickness Δd in the object under test.

We thus get for the differential current:

$$\Delta i = i_2 - i_1 = i_1 \left(\frac{i_2}{i_1} - 1\right) =$$

$$= i_1 \left(\frac{I_2}{I_1} - 1\right) = i_1\, (e^{\mu \Delta d} - 1).$$

As $\mu\Delta d \ll 1$ (since Δd is very small), we have:

$$\Delta i = i_1\, \mu\, \Delta d \quad \cdots\cdots\cdots \quad (1)$$

The absorption in the thin-walled glass of the ionisation chambers and hence also the ionisation due to the photo-electrons emitted by the glass have been neglected here, as well as the effect of the radiation scattered by the object under examination; this is permissible as the X-ray beam used had a very small diameter.

The experimental relationship found between the strength of the signal S and the depth Δd of the flaw is shown in *fig. 3* for a test object with a wall-

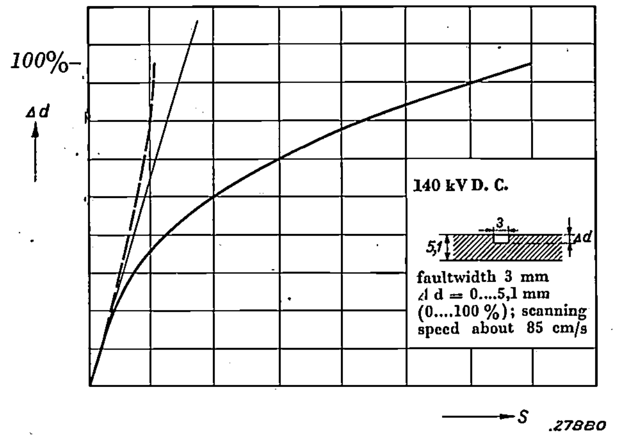

Fig. 3. Relationship between the depth Δd of the flaw expressed as a percentage of the wall thickness tested, and the signal strength S, in arbitrary units, with a flaw of constant width and a constant scanning speed.

thickness of 5.1 mm, the direct beam from an X-ray tube being used. According to equations (1) and their derivation, the curve would be expected to consist of an initial straight portion followed by a slower increase in the signal strength S with the depth of the flaw Δd, as for instance shown by the broken line in fig. 3. But in actual fact, S is found to increase rapidly with Δd, the increase commencing at a depth of about 30 per cent and approximating to the heavy line. This is probably due to the pronounced softening of the radiation transmitted by the flaw as Δd increases; towards the long waves this effect is less marked than with the rays transmitted through the overall wall thickness. The softer the X-rays, the greater will be their absorption by the gas filling in the tube and hence also the sensitivity.

There is a minimum current difference Δi which can still just be detected. At high amplifications the detection of flaws is limited by the statistical processes in the ionisation chambers, signifying that $\Delta i_{\min}$ is proportional to $\sqrt{i}$, e.g. $\Delta i_{\min} = q\sqrt{i}$. The smallest difference in thickness which can still be just detected in given circumstances is therefore:

$$\Delta d_{\min} = \frac{\Delta i_{\min}}{i_1\, \mu} = \frac{q}{\mu \sqrt{i_1}} \quad \cdots\cdots \quad (2)$$

The smallest relative difference in thickness is then:

$$\left(\frac{\Delta d}{d}\right)_{\min} = \frac{q}{\mu\, d \sqrt{i_1}} = \frac{q}{2\sqrt{p\, I_0\, \mu'\, D}}\, \frac{e^{\mu d/2}}{\mu\, d/2} \cdot \quad (3)$$

The quality of the X-rays (wavelength λ) is important in determining the smallest relative flaws which can be detected. If this relationship is investigated in the first place with a constant primary dosage ($p\, I_0 z \mu'\, D$ = const.) and variable wave-length, *i.e.* variable absorption coefficient, a minimum at $\mu\, d/2 = 1$ is found, in other words when about 12 per cent ($= e^{-2}$) of the initial intensity is transmitted. This optimum value is situated at a longer wave-length than that at which the thickness under examination is equal to the half-value thickness $d_{1/2}$, *i.e.* the thickness of the material at which 50 per cent of the initial intensity is transmitted. Since for this we have:

$$e^{-\mu d_{1/2}} = 1/2, \text{ or } \mu = \frac{\ln 2}{d_{1/2}} = \frac{0{,}7}{d_{1/2}}\,.$$

it can therefore be concluded that with a constant wavelength (and constant primary dosage) the optimum value is obtained at a thickness which is several times greater than the half-value thickness for the radiation in question.

With a non-homogeneous beam of X-rays, equations (2) and (3) naturally apply only for each wave-length individually (with the corresponding values of μ and μ'), and in addition the primary dosage at constant tube current increases roughly with the square of the tube voltage. As a result of this increase the optimum tube voltage is displaced at constant thickness of material towards higher values than that at which 12 per cent is transmitted. Conversely, with a constant tension the optimum thickness approximates more and more closely to the half-value thickness, although always remaining above it. This indicates, *inter alia*, that the highest practicable tube tensions should be employed, because as most objects tested are 10 mm or more thick the half-value thickness after moderate filtering is only 6 mm. of steel when using a beam obtained with a load of about 400 kilovolts on the tube. The optimum value can thus be obtained only with the thinnest and lightest objects, and in any case higher tube voltages must be used than in radiography, where about 100 kilovolts are required for screening through 10 mm of steel *e.g.* in 3 minutes when using intensifying screens. The X-ray energy transmitted is then about 1 per cent of the primary dosage, instead of something over 12 per cent ($\mu d \sim 5$ instead of ~ 1).

The curve in *fig. 4* was obtained in an investigation of the validity of equations (2) and (3), in which the relative flaw detectible is seen to have a minimum in terms of the wall thickness d, when the ratio between the strength of the signal and the average of the statistical variations is always

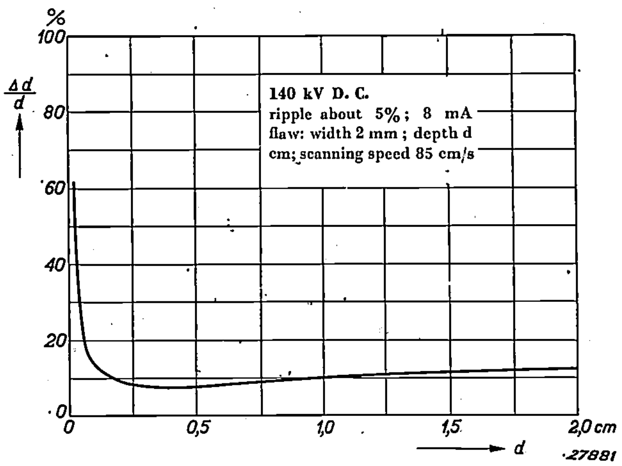

Fig. 4. The relative flaw $\Delta d/d$ detectible as a function of the wall thickness d at a constant tube voltage (140 kilovolts constant direct voltage; 8 milliamps), constant width of flaw (0.2 cm) and constant scanning speed (85 cm per sec.).

taken as equal to 2. The thickness d = 0.4 cm at the minimum, is, as can be expected, somewhat greater than the half-value thickness at the particular tube voltage applied (approximately 0.3 cm steel at 140 kilo-volts with the rays filtered as in the present case). The smallest detectible relative flaw is thus $7^1/_2$ per cent, in which connection it should be noted that, firstly, no steps were taken here to keep the statistical variations small, and, secondly, that a value of about 2 for the ratio of the signal required to the average statistical variation is on the small side for technical purposes. Double this value would be more acceptable.

Construction of the ionisation chambers

The construction of the ionisation chambers is shown in *fig. 5*, their principal features being exceptionally good insulation of the electrodes and rigidity, the latter being essential to prevent mutual vibration between the various parts of the

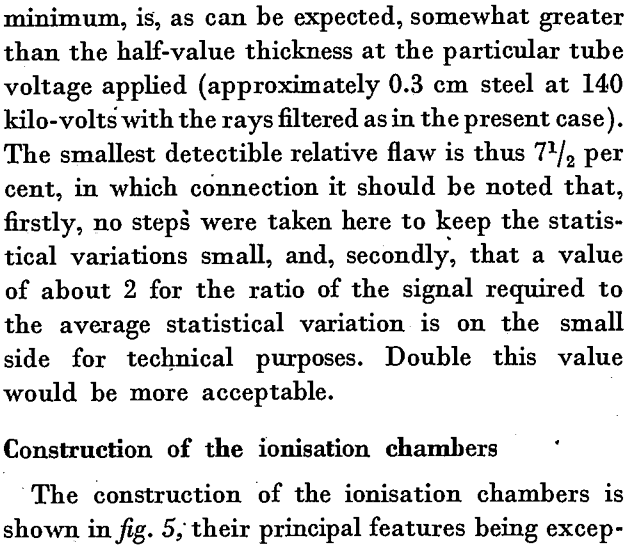

Fig. 5. Twin ionisation chambers for the automatic examinations of materials. A - earthed ring; B - beam of X-rays; the wall C is of a type of glass with very high insulating properties; D - exceptionally well insulated central electrode; E - electrodes connected directly to batteries. The wall F is made of a thin, very permeable type of glass.

apparatus (microphonic effect). The requisite insulation was realised by using a glass with a very low conductivity and by mounting an earthed ring A between the electrodes. All electrodes were made of chromium-steel so that they could be sealed through the glass. The middle electrode is a close fit in the ionisation chamber, but does not touch the glass so that the insulation is not in any way effected. The gas pressure in the two ionisation chambers is thus the same; they are naturally also of the same length.

The absorption of the radiation by the gas, and hence the magnitude of the ionisation current, increases with the atomic number of the gas and its pressure, xenon at a pressure of about 1 atmos. was used in the majority of the experiments, as well as in those to which fig. 4 applies. The pressure in a xenon ionisation chamber can be raised above 1 atmos. merely by allowing a known quantity of the pure gas to condense in the evacuated chamber, when the latter is immersed in liquid air. After sealing, the chamber is again raised to atmospheric temperature when the xenon evaporates. In this way a xenon pressure of 10 atmos. can readily be built up. The measurements described here with a load of 150 kilovolts applied to the tube showed that the number of ions produced in the xenon is roughly 50 times greater than in air at the same pressure.

The length of the ionisation chamber (D in fig. 2) should be made as large as possible, although it is shown below that no advantage accrues from increasing D with a specific cross-section if the capacity of the chamber is already large as compared with the residual capacity at the grid of the first amplifying valve. The diameter of the X-ray beam is made larger than the diameter a of the smallest blowhole to be detected, for these blowholes must give a just detectible signal, requiring that their whole area lie within the beam. Thus to examine a hollow cylindrical object, it is scanned by the beam along a helical path as already indicated above, the object being rotated about its axis and the chamber at the same time being moved parallel to this axis. If the pitch of the scanning helix is s, the optimum diameter of the beam will be $a + s$, since with a larger-diameter beam the signal given by the smallest flaw will not be any greater, although there will be an increase in the ionisation current in the rest condition and hence in the statistical fluctuations which limit the sensitivity. For the same reason the pitch should not be made too great and should be $s = {}^1/_3\, a$.

A diaphragm with two apertures should be placed between the object under test and the chamber to prevent scattered radiation from one beam reaching the other chamber, as this would reduce the differential current.

Statistical variations in the ionisation currents

The fact that the charged particles reach the electrodes at irregular intervals determines on the whole the magnitude of the minimum differential current detectible, since this current must be, for instance, three to five times greater than the average of the variations in the ionisation current. It is apparent that the statistical fluctuations of an ionisation current are considerably greater than those of an equivalent electron current emitted from an incandescent filament, since the fluctuations in current depend on the magnitude of the elementary charge. In the case of a radiating filament this is an electron, while in an ionisation chamber a very rapid primary electron, which generates a very large number (about 1000) of secondary electrons on collision with the gas atoms, is produced per absorbed X-ray quantum (in the intervals of statistical variation). On the average these collisions take place at equal intervals and in this way the whole yield of a primary electron more or less plays the part of the elementary charge in the statistical analysis of the current fluctuations. These fluctuations may in fact be 30 times ($\sim \sqrt{1000}$) greater than with a current of similar intensity emitted from an incandescent filament.

The average value of the fluctuations is proportional to the square root of the ionisation current [2]), so that as the ionisation current increases the smallest detectible flaw diminishes, when by taking the necessary measures for obtaining a high ionisation current (enlarging the chamber or raising the pressure) the differential pressure also increases proportionally (equation (3)). For a specific flaw the ratio of the signal to the fluctuations is then improved in proportion to the square root of the gas pressure in the chamber; increasing the current through the X-ray tubes gives the same result.

Differences in the velocities of the ions

The velocity of the carriers of the positive charges in the gas contained in the chambers, the xenon ions, is about 10 to 100 times smaller than that of the carriers of the negative charges in this gas. The gas used is in fact so pure that the electrons liberated cannot form negative ions with a much lower velocity by linking up with electro-negative

[2]) Philips techn. Rev. 2, 136, 329, 1937; 3, 189, 1938.

atoms. The mean time of travel required by the positive ions to reach an electrode is about 3.10^{-3} sec., while that of the electrons is negligibly small. On a sudden change in the intensity of the X-ray beam, the associated change in the electron current in one of the chambers will occur about 3.10^{-3} sec. earlier at the middle electrode than the change in the ionic current in the other chamber. As a result a ripple in the tension applied to the X-ray tube, owing for instance to disturbances at the rectifier or insufficient smoothing of the tube voltage, will still give a signal in spite of using the compensation circuit in fig. 1. (Naturally this signal is much weaker than would be obtained from a single ionisation chamber with the ripple in question).

The high-tension unit

For two reasons the high-tension unit must furnish as constant a voltage as possible:

a) Because the detectible flaw depends on the voltage (I_0 and μ in equation (3)), and
b) because the ripple produces a signal.

A circuit must therefore be used which gives a constant direct voltage and which is provided with sufficiently large smoothing condensers. It is known from dosage measurements that the difference in the intensities through 10 and 7 mm. of steel increases with the fourth power of the tube voltage, when the latter lies between 100 and 200 kilovolts. An alteration in voltage of 5 per cent will thus result in an alteration in signal strength of about 20 per cent, indicating that the ripple in the tube tension must not exceed 1 to 2 per cent. A highly pulsating voltage is in any case quite useless, because time intervals then occur during which a flaw cannot be detected at all as no X-rays are passed through the object under examination.

Detectible flaws

According to equations (2) and (3) the detectible flaws are also determined by the thickness of the wall through which the rays are passed: the ratio of the relative flaws which can still be detected in 10 and 8 mm of steel is equal to 1.3 : 1, so that the ratio of the absolute flaws is 1.6 : 1. It was actually found in one case that flaws of 3 and 2 mm in depth could be located with equal facility in steel 10 and 8 mm thick respectively, which applies for a constant adjustment of Δi_{min}. The alteration in sensitivity with the wall thickness can be easily avoided by moving a template, which is also placed in the path of the rays, in unison with the object under examination. If the template has a suitable range of thicknesses the total thickness of material can be kept constant. A further advantage of this arrangement is that the ionisation currents, and hence also their fluctuations and the sensitivity of the apparatus (Δi_{min}), are kept constant irrespective of the thickness of the material. With a constant primary dosage ($p\, I_0\, \mu\, D$ = constant) the relative flaw would remain constant with a variable d, if the wavelength of the radiation is varied in such a way that the product of the coefficient of absorption and the thickness of the material d remains constant, in other words d is proportional to $1/\mu$. As for the half-value thickness $d_{1/2} = \log_e 2/\mu$ the quotient of thickness of the material and the half-value thickness $d/d_{1/2}$ would also be constant with a constant relative flaw. The curve of the half-wave thicknesses plotted as a function of wavelength, when drawn on a certain scale, will also represent the wall thickness with a constant value of the smallest detectible relative flaw as a function of the wavelength (with constant primary dosage).

First amplifying stage; form of signal

If a flaw, such as a blowhole in a casting, moves past one of the ionisation chambers a differential current with a time function as shown in *fig. 6* will

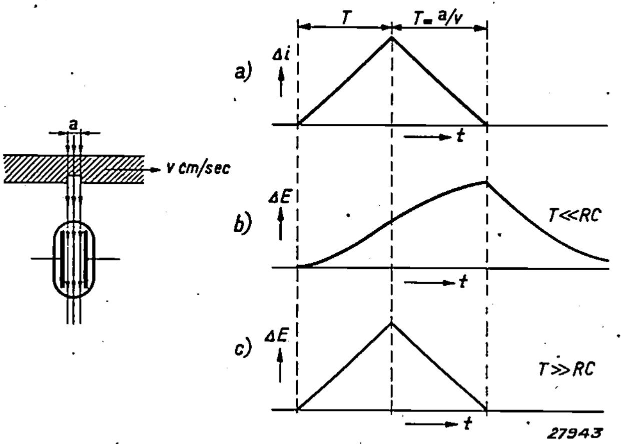

Fig. 6. Variation of current $\Delta i(a)$ through the grid resistance and the voltage ΔE (b and c) at the grid when a flaw of size a travels past one of the ionisation chambers with a velocity v.

be obtained, provided the diameter of the cavity is the same as that of the beam. Two limiting cases occur according to the circuit used, *viz.*, that the time T by the flaw to pass in front of the chamber is either very short or very long as compared with the relaxation period RC, where R is the grid resistance in the compensation circuit and C the total capacity at the grid (fig. 1). The variation of the grid voltage of the first valve in these two limiting cases is shown in fig. 6b and 6c respectively. The maximum resulting variations in grid voltage can be easily calculated. For $T \ll RC$, we have:

$$\Delta E_{\max} = \frac{1}{C}\int_0^T \Delta i\, dt = \frac{1}{C}\, T\, \Delta i_{\max} \quad \cdot \quad (4)$$

and for $T \gg RC$

$$\Delta E_{\max} = \Delta i_{\max}\, R \cdot \cdot \cdot \cdot \cdot \cdot \quad (5)$$

It may therefore be expected that at high scanning speeds v, to which equation (4) applies, the maximum strength of signal will be proportional to T (and inversely proportional to v). With very low scanning speeds the signal strength may be expected to be independent of the velocity v in accordance with equation (5). This is confirmed by *fig.* 7 for a particular case where $\Delta i_{\max} \sim 10^{-10}$

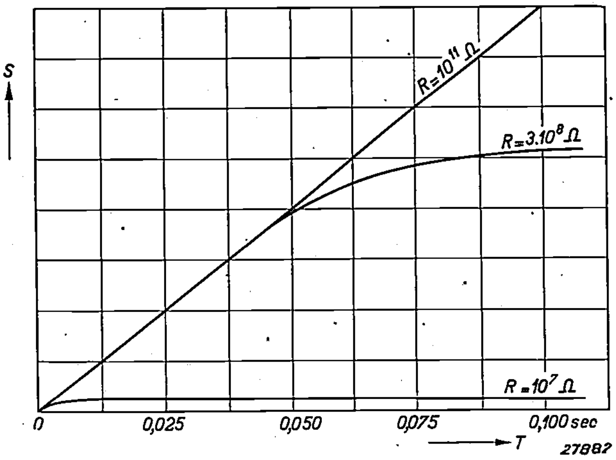

Fig. 7. Variation of the maximum signal strength S in arbitrary units with variable scanning speed v. The time T which the flaw takes to pass the chamber is plotted along the abscissa (see fig. 6); T is the quotient of the diameter of the flaw a and the scanning speed v.

amps. and $C \sim 10^{-10}$ farads. The greater the resistance R, the longer will the condition $T > RC$ be satisfied, and the longer will the curve remain a straight line and hence the higher will be the signal strength at which the curve bends to become horizontal.

For given values of $\Delta i_{\max}$, C and T the signal will become stronger with increasing R, until the condition in equation (4) is reached. A further increase in R is then useless, and in fact undesirable, as with a higher resistance there is a greater probability of interference occurring. The smaller C is made, the greater will naturally be the alteration in grid voltage. On making the ionisation chamber longer, $i_{\max}$ will increase in proportion to the length D (fig. 2), although at the same time the capacity of the chamber will increase, again in proportion to D. Under the conditions represented in equation (4) increasing the length of the chamber is useless, for as soon as the specific capacity of the ionisation chambers is made large as compared with the remaining grid capacity, then both Δi and C increase in proportion to D, and ΔE remains constant. In the case represented by equation (5) an increase in length will always give an improvement, although in general it will not be advisable to work with the conditions in this equation, since then ΔE is much smaller than with equation (4). An attempt could be made in the latter case to deduce the shape of the flaw from the shape of the ΔE curve, but an amplifier which will not distort arbitrary non-periodic voltage impulses is then required. In many cases, tests according to equation (5) offer an advantage, for instance in the examination of objects in which the various parts pass in front of the chambers at different velocities, *e.g.* tubes with different diameters which are rotated with a constant angular velocity; in these cases the smallest detectible flaw must not be determined by the velocity of motion of the object as required by equation (5). A better method in these cases is to adapt the speed of revolution of the tube to the diameter of the section passing in front of the chambers at each particular moment. It should also be pointed out that the maximum variation in voltage when $T \ll RC$ is a measure for the volume V of the cavity in the path of the rays, since:

$$\Delta E_{\max} = \frac{1}{C}\int_0^{2T} \Delta i \cdot dt \,(:)\int_0^{2T} \Delta d \cdot dt\,(:)\int_0^{2a} \Delta d\,\frac{ds}{v}\,(:)\int_0^{V} \frac{dV}{v},$$

where s is the distance traversed by the flaw. A necessary condition for this relationship to obtain is that the flaw is displaced at a constant speed v. When $T \gg RC$ the maximum variation in voltage is a measure for the maximum size of the flaw in the direction of the rays. These characteristics of the limiting cases (4) and (5) will also determine which conditions should be selected according to the circumstances ruling. These rules also apply approximately after further amplification of the signal.

When the flaw passes the two chambers in succession where both lie along the line of the scanning motion, a signal will be obtained first in one sense, *e.g.* an increase in grid voltage, and then in the opposite sense. The first signal must actuate the alarm device, which may be achieved by a reduction in a given voltage. To detect cavities in an object only one direction of motion is correct with a given amplifier, while to detect increases in thickness the opposite direction of motion is usually required. Difficulties can naturally arise with this arrangement, when the flaw is located in front of one of the chambers for so short a time that the times of transit of the ions are no longer small compared with the rate of displacement of the flaw itself.

An electrometer triode, Type No. 4060, was used as the input valve, since this type of valve allows a very much greater grid resistance than other types. The voltage amplification was approximately 1 : 1 and the current amplification about 10^6. To protect the valve adequately against mechanical and electrical disturbances, it was enclosed in a heavy block of iron, itself mounted on resilient rubber feet and surrounded by a copper enclosure with walls 2 mm thick. The sensitivity of the valve itself is limited by the current fluctuations in the grid circuit in the absence of ionisation current. Alfvèn [3]) has calculated and measured a maximum charging sensitivity of approximately $2.5 \cdot 10^{-17}$ coulomb for an electrometer triode of this type. Measurements of the variations and an estimation of the grid capacity of the valve (approximately 10^{-11} farad) gave nearly the same value for the apparatus described here. Yet these variations are not a measure of the sensitivity for the present combination of ionisation chambers and electrometer valve, for the sensitivity is on the other hand determined by the variations in the ionisation currents flowing in the rest condition.

Further amplification and alarm device

For detecting small blowholes in castings which during the test must be moved rapidly in front of the chambers, the apparatus must be capable of indicating short voltage impulses; this allowed an ordinary alternating-voltage amplifier with condenser-resistance coupling to be used as a further amplifying stage. A single-valued relationship was found between the strength of the signal at the output of the amplifier and the size of the blowhole. The amplification was about 10^4.

The alarm arrangement consisted of a relay valve which is normally blocked by an appropriately high negative grid bias. A voltage impulse which reduces this grid bias friggers the valve, so that the anode current then switches on a siren or a lamp through a relay. By a suitable choice of the negative grid bias of the relay valve adjustment can be made to the minimum size of flaw detectible. The valve does not then allow current to pass with flaws smaller than this limiting value and no alarm is given. The alarm only ceases when cut off by the attendant (interrupting the anode current), so that no flaw can pass undetected owing to inattention on the part of the operators.

[3]) H. Alfvèn, Z. Physik, 99, 24, 1936.

Temporary design of apparatus and potential improvements

A temporary apparatus was set up for detecting blowholes not less than 3 mm in diameter in cast pipes with walls 10 mm thick. The ionisation chambers were placed inside these pipes, which were then scanned along a helical path (*fig. 8*). A special

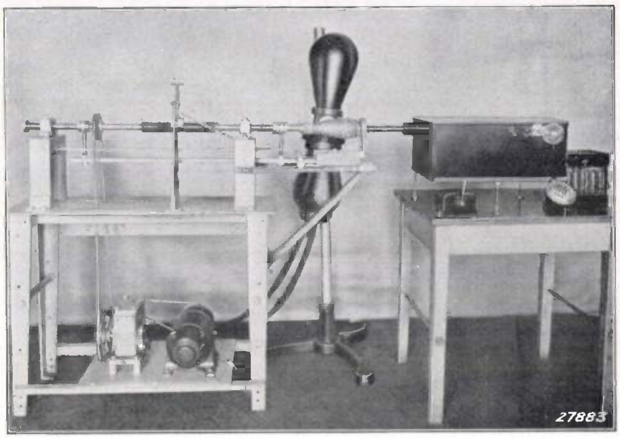

Fig. 8. The apparatus is adjusted for detecting blowholes 3 mm in diameter in cast-iron pipes with walls 10 mm thick.

machine was used for rotating the pipes and for moving them parallel to the axis. The speed of revolution was 200 r.p.m., and the pitch of the scanning motion 1 mm. The diameter of the scanning beam was 4 mm. A pipe length of 20 cm. was tested per min, the peripheral speed of the pipe being about 1 m. per sec. The walls of the pipes were 8 and 10 mm thick. If the apparatus is so adjusted that flaws of 3 mm are still just detectable in the 10 mm wall, an alarm is obtained with the 8 mm wall when the flaw measures only 2 mm. This undesirable alteration in sensitivity was corrected by a template of variable thickness which was moved together with the pipe, so that the total thickness of metal radiographed was maintained at 10 mm. Alternatively the negative grid bias of the relay tube could have been varied in direct relation to the position of the object holder. A variable diameter in the object under test can be allowed for by automatically altering the speed of revolution.

The apparatus is enclosed in a sheet-copper housing, which acts as a protection against external disturbances. The batteries for the ionisation chambers and the input amplifying valve are also accommodated in the same housing.

The characteristics of the apparatus described were adapted to the requirements specified, which were not very severe. Various features can still be considerably improved:

a) The ionisation chambers were about 15 cm from the tube focus, but by using a modern small-diameter X-ray tube this distance can certainly be reduced to one half, thus increasing the intensity by a factor 4.

b) The current rating of the tube used can be at least doubled.

c) The gas pressure in the ionisation chamber can be increased to about 30 atmospheres, by means of which it can be safely assumed the ionisation current will be increased tenfold.

d) In very many cases the ionisation chambers can be made much longer, *e.g.* 60 mm instead of the 12 mm used above, *i.e.* five times greater. By using the X-ray tube referred to above, in which the focus is at the end of a narrow tube, the ionisation chambers can in most cases be placed outside the objects under examination, and the tube itself inserted within the object. The length of the chambers is then no longer determined by the internal diameter or bore of the object under test.

It thus appears possible to increase the intensity of the ionisation current about $4 \times 2 \times 10 \times 5 = 400$ times, allowing the thickness of wall of the object under test to be increased to possibly 30 mm. These improvements do not take into consideration any increase in the tube voltage, which as indicated above can be taken very high.

RADIO INTERFERENCE

by L. BLOK.

621.396.823

This article discusses how radio interference is produced by electrical apparatus and how the interference is impressed on a radio receiving set. The cause of radio interference is always a more or less discontinuous variation of voltages and currents in an electrical apparatus or appliance. Interference is transmitted to the receiving set mainly through capacitive coupling of the interference-carrying mains with the aerial or the aerial lead-in or by propagation of the interference through the mains to the mains unit of the receiver. In both cases the way in which the interference is transmitted from the producer to the mains is an important consideration and is discussed in detail. A distinction must be drawn between symmetrical and asymmetrical interference producers, the latter in particular being responsible for serious interference with radio reception.

Introduction

The increasingly severe demands imposed on the quality of radio reception have led to more and more attention being devoted to those factors which adversely affect and detract from the efficiency of reception of broadcast radio transmissions.

Distortion has been reduced to a minimum by progressive improvements in the construction of transmitters and receiving sets and it is, therefore, not surprising that as a result increasing attention has been given to the problem of interference with radio reception, whether due to atmospheric causes or to any kind of electrical machinery or appliance. This desire for purer reception has of course resulted in extraneous noises which were formerly accepted without complaint now being regarded as a serious disturbance. In consequence international efforts have been made to prohibit by legislation the use of electrical apparatus and appliances liable to cause interference with radio reception. The present article discusses the following aspects of this subject:

1) The production of interference by certain electrical apparatus:
2) The manner in which a disturbance is propagated from its source to a radio receiver; it is important in this connection to distinguish between symmetrical and asymmetrical components.
3) How a disturbance enters the receiving set.

Production of radio interference

Interference with radio reception is always due to more or less discontinuous variations in current and voltage. The clicking sound produced in a loudspeaker by the operation of a switch in a receiving set irrespective of the wavelength to which the receiver is tuned, is a well-known example of this. The cause of this noise is that a current impulse is composed of a frequency band in accordance with the Fourier integral theorem, which may contain all frequencies from zero to a value which is higher the more abrupt the impulse. As the receiver responds to all waves in the frequency band to which it is tuned, the clicking sound is produced.

If the current impulses occur periodically, as the tube focus, but by using a modern small-diameter X-ray tube this distance can certainly be reduced to one half, thus increasing the intensity by a factor 4.

b) The current rating of the tube used can be at least doubled.

c) The gas pressure in the ionisation chamber can be increased to about 30 atmospheres, by means of which it can be safely assumed the ionisation current will be increased tenfold.

d) In very many cases the ionisation chambers can be made much longer, *e.g.* 60 mm instead of the 12 mm used above, *i.e.* five times greater. By using the X-ray tube referred to above, in which the focus is at the end of a narrow tube, the ionisation chambers can in most cases be placed outside the objects under examination, and the tube itself inserted within the object. The length of the chambers is then no longer determined by the internal diameter or bore of the object under test.

It thus appears possible to increase the intensity of the ionisation current about $4 \times 2 \times 10 \times 5 = 400$ times, allowing the thickness of wall of the object under test to be increased to possibly 30 mm. These improvements do not take into consideration any increase in the tube voltage, which as indicated above can be taken very high.

RADIO INTERFERENCE

by L. BLOK. **621.396.823**

This article discusses how radio interference is produced by electrical apparatus and how the interference is impressed on a radio receiving set. The cause of radio interference is always a more or less discontinuous variation of voltages and currents in an electrical apparatus or appliance. Interference is transmitted to the receiving set mainly through capacitive coupling of the interference-carrying mains with the aerial or the aerial lead-in or by propagation of the interference through the mains to the mains unit of the receiver. In both cases the way in which the interference is transmitted from the producer to the mains is an important consideration and is discussed in detail. A distinction must be drawn between symmetrical and asymmetrical interference producers, the latter in particular being responsible for serious interference with radio reception.

Introduction

The increasingly severe demands imposed on the quality of radio reception have led to more and more attention being devoted to those factors which adversely affect and detract from the efficiency of reception of broadcast radio transmissions.

Distortion has been reduced to a minimum by progressive improvements in the construction of transmitters and receiving sets and it is, therefore, not surprising that as a result increasing attention has been given to the problem of interference with radio reception, whether due to atmospheric causes or to any kind of electrical machinery or appliance. This desire for purer reception has of course resulted in extraneous noises which were formerly accepted without complaint now being regarded as a serious disturbance. In consequence international efforts have been made to prohibit by legislation the use of electrical apparatus and appliances liable to cause interference with radio reception. The present article discusses the following aspects of this subject:

1) The production of interference by certain electrical apparatus:
2) The manner in which a disturbance is propagated from its source to a radio receiver; it is important in this connection to distinguish between symmetrical and asymmetrical components.
3) How a disturbance enters the receiving set.

Production of radio interference

Interference with radio reception is always due to more or less discontinuous variations in current and voltage. The clicking sound produced in a loudspeaker by the operation of a switch in a receiving set irrespective of the wavelength to which the receiver is tuned, is a well-known example of this. The cause of this noise is that a current impulse is composed of a frequency band in accordance with the Fourier integral theorem, which may contain all frequencies from zero to a value which is higher the more abrupt the impulse. As the receiver responds to all waves in the frequency band to which it is tuned, the clicking sound is produced.

If the current impulses occur periodically, as those originating in electrical motors, rectifiers, electro-medical apparatus, etc. the disturbance will also be repeated periodically and will then become audible as the crackling sounds familiar to all listeners.

In many commutator motors disturbances are caused by voltages produced as the brush moves from one commutator bar to the next. If the brush remains in contact with two consecutive bars that part of the armature winding between these two bars is short-circuited. The voltage between the bars and the brush is then zero, but as soon as the brush breaks contact with one of the bars the short circuit over part of the armature winding is suddenly removed, so that the short-circuit current rapidly drops to zero. This action is accompanied by a rapid increase in the p.d. between the bar and the brush, which may give rise to more or less intensive sparking. When the brush comes in contact with the next bar another section of the armature winding is shorted and a short-circuit current is again rapidly built up, and so on. These sudden variations in current induce voltages over a wide frequency band in the whole of the armature winding, and these pulsating voltages may be impressed on the mains to which the motor is connected and in this way cause interference with radio apparatus fed from the same mains or situated close to a conductor in the mains network.

It should be noted that sparking is not essential for interference to be produced. If a motor does not spark, this does not justify the conclusion that the motor cannot interfere with radio reception.

Other pieces of apparatus which owing to poor design may interfere with radio reception are rectifiers. Abrupt changes in voltage occur in gas-filled rectifiers owing to the difference between the re-ignition voltage V_D required in each half-wave and the running voltage V_B. The greater this difference V_D-V_B, the more powerful will be the disturbances caused by the apparatus. Exceptionally high values of V_D-V_B may occur in controllable rectifiers with relay valves in which V_D is dependent on a variable grid voltage [1]).

Other electrical apparatus liable to act as sources of disturbance include discharge tubes used for advertisement signs and electro-medical apparatus, which generate high-frequency electrical fields.

[1]) The grid voltage varies with the same frequency as the voltage at the anode of the rectifying valve. The D.C. can be controlled by carrying the mutual phase displacement of these two voltages. *Cf.* the detailed discussion by J. W. G. Mulder and H. L. van der Horst: A controllable rectifying unit for 20 000 volts and 18 amps. Philips techn. Rev., **1**, 161, 1936.

Propagation of interference

The disturbance generated can be impressed on radio receivers in various ways:

a) by direct radiation or by capacitive coupling of the source of disturbance itself with the aerial or the aerial lead-in of the receiver;
b) by radiation or by capacitive coupling of lighting mains propagating a disturbance, with the aerial or the aerial lead-in;
c) by transmission of the disturbance through the mains direct to the receiving set;
d) by a combination of the above.

A disturbance rarely reaches a radio set by the means enumerated under a), in fact only when the aerial is located close to the source of the disturbance. In practice radiation is as a rule quite small. Screening of the source of disturbance is the only method of suppression which can be adopted with this type of interference, if the cause of the disturbance itself cannot be eliminated directly at the source.

Far more frequently, and in fact nearly always, interference with radio reception is due to one or another of the methods enumerated under b) and c). To obtain a closer insight into the conditions ruling in these cases we shall investigate how a disturbing voltage can be impressed on the low-voltage mains and how it reaches a radio receiving set through these mains. An equivalent layout has been evolved below in which the circuits affected by the interfering voltage are indicated.

Equivalent circuit of a disturbance source

The interfering voltage of an electrical apparatus can in general be represented as a varying electromotive force E with an internal impedance Z_i in series with it (see *fig. 1*). If the apparatus is connected to the mains, both terminal *1* and terminal *2* of the interfering voltage are connected through certain impedances with the mains terminals *3* and *4*. In addition a certain impedance will exist between terminals *1* and *2* and between terminals *3* and *4*, so that a generalised quadripole (V_1) is produced (see component *I* in fig. 1). Frequently certain parts of the electrical apparatus are earthed or possess an appreciable capacity to earth. As a result terminals *1* and *2* are connected to earth through certain impedances Z_{1a} and Z_{2a}. All these impedances are of course, usually complex and dependent on the frequency.

Having determined which components of the circuit transmit the interfering voltage to the mains terminals *3* and *4*, we shall discuss the component marked *II* in fig. 1 which connects the mains terminals of the interference source to the mains terminals *5* and *6* of the receiving set. If consideration is limited to terminals *3*, *4*, *5*, *6* and to the earth A of the whole complex network — the importance of the earthed pole will appear later on in this discussion — the mains circuit together with the connected apparatus (earthed or not) can be replaced by a generalised quadripole N and four impedances Z_3 to Z_6, which link the four connections of the quadripole with earth [2]).

Finally, component *III* of fig. 1 gives an equivalent circuit for the receiver which is composed of an impedance between the terminals *5* and *6* and two earthing impedances Z_{5a} and Z_{6a} between these terminals and earth.

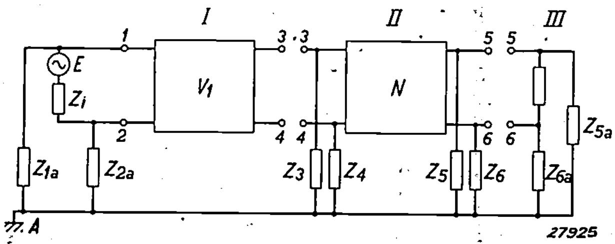

Fig. 1. Equivalent circuit of an interference source. The interfering voltage E with the internal impedance Z_i is located between terminals *1* and *2*. Terminals 5 and 6 are the inputs of the receiver; A is earth. V_1 and N are generalised quadripoles; the first connects the interfering voltage to the mains terminals of the interference producer, and the second represents the network between the interference source and the terminals of the receiver.

Symmetrical and asymmetrical interference

Before referring in detail to the various components of the circuit outlined above, our diagram must be further simplified in order to show that a disturbance can be transmitted in two fundamentally different ways from terminals *1* and *2* to terminals *5* and *6*. We here make use of the fact that an e.m.f. E with an internal impedance Z_i in series with it is electrically equivalent to a current source of intensity $I = E/Z_i$ and an impedance Z_i in parallel with it. This is shown diagrammatically in *figs. 2a* and *b*. The three impedances in fig. 2*a* form a triangle, which can be electrically replaced by a star circuit as shown in fig. 2*c*.

The simplification referred to consists essentially in substituting for the whole circuit between terminals *1*, *2* and *5*, *6* a single quadripole whose terminals are earthed through given impedances. These impedances are connected in parallel to the impedances Z_{1a}, Z_{2a}, Z_{5a}, Z_{6a} shown in fig. 1 and to both sides of the quadripole form a triangle of impedances between the points *1*, *2*, A and *5*, *6*, A respectively. On transforming this triangle into a star circuit according to fig. 2 and replacing, also as in fig. 2, the interfering voltage by an interfering current, a new equivalent circuit is obtained which is shown in *fig. 3*.

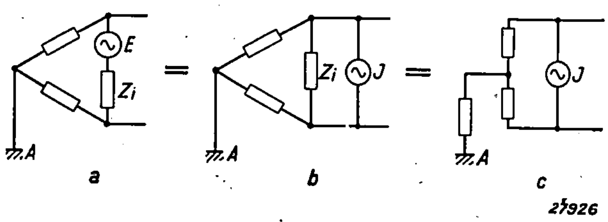

Fig. 2. Three equivalent circuits for an interference source. The voltage source can be replaced by a current source, and the impedance triangle can be transformed into a star circuit.

The interfering current I will flow partly through the impedances Z_1 and Z_2 and partly through the junction between terminals *1* and *2* in the mains circuit V (for instance, through the capacity between the mains conductors) and finally through impedances Z_5 and Z_6 in the receiver. In general the midpoints P_1 and P_2 of the star circuits will not be at the same voltage, so that part of the interfering current will flow through the earth connection. The impedance $Z_1 + Z_2$ can, however, always be subdivided into two impedances Z_1' and Z_2' (see *fig. 4a*), such that point P_0 is at the same voltage as the star point P_2 in fig. 3. A potential difference will exist between P_0 and P_1 which will be in a constant (complex) relationship to the interfering current I. We shall term this potential difference the asymmetrical interfering voltage and denote it by the symbol E_a. If we impress an e.m.f. equal to the asymmetrical interfering voltage $E_a = V_{P_1} - V_{P_2}$ as shown in fig. 4*b*, no p.d. will exist between the terminals P_0 and P_0' in this diagram. An opposing voltage source will be connected in series with this e.m.f., so that 4*b* really coincides with 4*a*. By the principle of superposition 4*b* can be resolved into two circuits as shown in fig. 4*c*. The voltages and currents at the receiver are equal to the sum of the voltages and currents

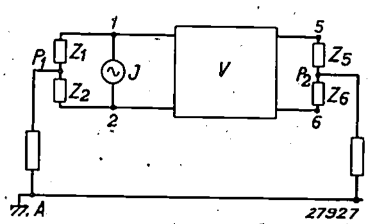

Fig. 3. Equivalent circuit of an interference source as in fig. 1, after simplification by introducing the transformations shown in fig. 2.

[2]) Regarding this point and other general characteristics of linear networks *cf.* the article by Bath. van der Pol and Th. J. Weyers, Electrical filters, Philips techn. Rev. 1, 240, 270, 298, 327, 363, 1936.

produced by the interference sources *I* and *II* in fig. 4*c*.

This resolution has the following advantage: It is seen that although impedances are present in component *II* of the interference source along the earth connection between P_0 and P_2, no potential differences are yet obtained. The current in the earth conductor is therefore zero. A voltage E generates a current which produces the same potential difference along both mains conductors (and thus has the same direction and roughly the same magnitude in both conductors), and which flows back through earth. The voltage between the mains conductors is usually subject to only a slight change as a result of this, the principal effect being voltage fluctuations between the

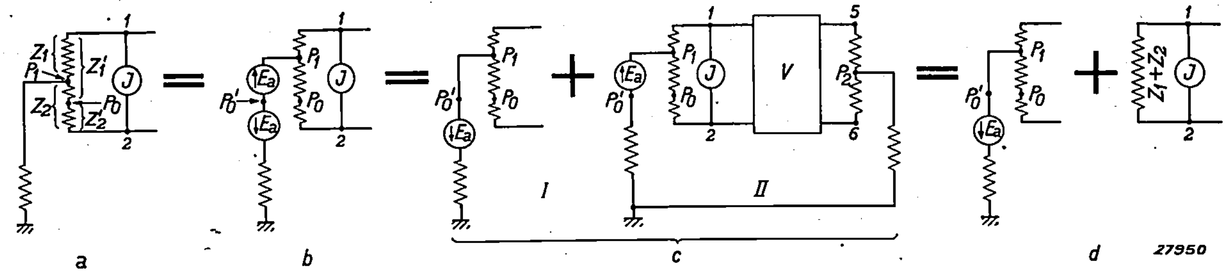

Fig. 4. Four similar equivalent circuits of the input in fig. 3. By introducing an asymmetrical e.m.f. E_a in such a way that the potential of P_0' is made equal to that of P_0, the source of interference can be subdivided into two components *I* and *II* (fig. *c*) of which the symmetrical component *II* produces no flow of current to earth.

part of the aggregate interference is thus isolated which produces currents of equal intensity but of opposite polarity in the two mains conductors, so that no currents flow to earth. This component we shall term the symmetrical source of interference. The earth connections need not be considered for a closer investigation of the effect of the symmetrical interference component, so that the equiualent circuit can be further simplified to give that shown in

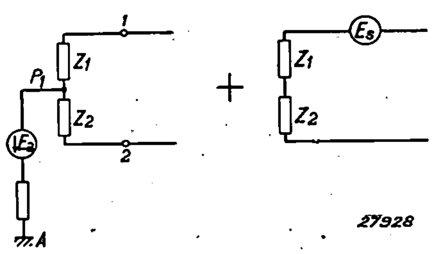

Fig. 5. Simplified layout of the symmetrical interference component E_s and the asymmetrical interference component E_a, which together furnish the interference voltage as shown in fig. 1.

fig. 4*d*. Finally, by eliminating the earth connection the current source with the impedance $Z_1 + Z_2$ in parallel can be again replaced by a voltage source $E_s = I\ (Z_1 + Z_2)$ with the impedance $Z_1 + Z_2$ in series. The new equivalent circuit for the complete disturbance is shown in *fig. 5*.

The symmetrical interfering voltage E_s produces opposing currents in the two mains conductors, no potential differences appearing between the two star points and earth. The asymmetrical interfering conductors and earth. These two interfering voltages can now be usefully discussed in further detail.

Symmetrical interfering Voltage

The equivalent circuit for the symmetrical interfering voltage is shown in *fig. 6* in somewhat more detail than in the previous figure. Z_i is the internal impedance of E_s; the characteristics of the mains are for the sake of simplicity repressented by three impedances [3]). The magnitude of Z_1 (which can be both inductive or capacitive) is dependent on the frequency.

The mains impedances, Z_s and Z_p, are also functions of the frequency, being governed by the length and layout of the network. Measurements have shown that the two mains conductors usually have a very high capacity with respect to each other so that the impedance Z_p is small, and as a

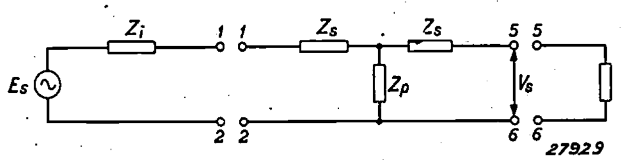

Fig. 6. A detailed scheme of the symmetrical interference component. The interfering voltage V_s is generally small compared with E_s, because the parallel impedance Z_p between the mains conductors is usually very small owing to their mutual capacity.

[3]) This is possible as the earth connections have been eliminated so that there is no external connection between terminals *1* and *2* and terminals *3* and *4*.
Cf. the article by Balth. van der Pol and Th. J. Weyers, *loc cit.*

result the interfering voltage V_s at the receiver is usually very small as compared with E_s, provided Z_i does not fall below a certain lower limit. A symmetrical interfering voltage in the mains will thus as a rule cause little interference in radio receivers connected to these mains.

As the two mains conductors are generally situated close together, the radiation from the mains is also small where a symmetrical interfering voltage exists.

Asymmetrical interfering voltage

As already indicated above, a circuit with an asymmetrical interfering voltage is made up of the two mains conductors connected in parallel and the earth connections, according to the arrangement shown in *fig. 7*.

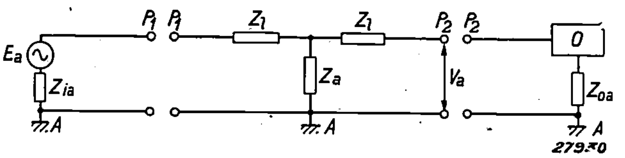

Fig. 7. Layout of the asymmetrical interference component. The interfering voltage V_a may be much greater then V_s (fig. 6), particularly when the earthing resistance Z_a of the mains is not too small.

The interfering voltage E_a is connected to earth through the impedance Z_{ia}. The series impedances of the mains conductors in parallel are represented by Z_l, and the capacitive impedance between mains and earth is denoted by Z_a.

In the case of cables the capacity of the cores to earth is usually fairly high; not only has a cable section buried in wet soil a high capacity to earth, but the unburied sections also exhibit a pronounced capacity between the cores and the lead sheath, the latter being more or less well earthed. Thus Z_a will usually be small. But in the conductors of house installations Z_a may have a much greater value. Although the tubing or lead sheath surrounding these conductors is frequently earthed at one point, this is not sufficient. Close to the earthed point the tubing or the lead sheath may be regarded as equivalent to earth; at some distance from this point the resistance of the tubing cannot be neglected, especially as it increases with the frequency as a result of skin effect. Owing to the considerable distance between the conductors in a house installation and earth, which in this case acts as a return, radiation from the conductors with an asymmetrical interference component is much greater than with a symmetrical disturbance, which accounts for the fact that in by far the majority of cases interference with radio reception can be ascribed to the asymmetrical components of the interfering voltage, when the source of interference is not situated in the immediate neighbourhood.

A feasible method for suppressing asymmetrical interference appears to be a modification of the receiver circuits so that the star point P_1 in fig. 3 is at the same potential as P_2. This is not altogether simple since the p.d. between P_2 and P_1 depends on the nature of the quadripole V between terminals *1* and *2* of the interfering source and terminals *5* and *6* of the receiver. But this quadripole contains all side branches of the network, and in these branches the nature or magnitude of the load may vary, which will result of course in a change of the characteristics of the quadripole V.

Yet if the earthing resistance Z_{ia} of the interfering source is made sufficiently large the asymmetrical current can be kept within specified limits. Z_{ia} will then be determined not only by the earthing resistance of the interference producer, but also by the earthing impedances Z_3 and Z_4 in fig. 1. This is apparent from the method by which fig. 3 has been derived from fig. 1. These impedances are usually small and as a result Z_{ia} is also reduced. The effect of the impedances Z_3 and Z_4 in fig. 1 on the earthing resistances Z_{ia} can however be reduced to negligible dimensions by making the series impedance of V_1 (fig. 1) high on the side of terminals *3* and *4*. In certain cases it is then possible to give Z_{ia} a high value.

Another method of suppressing interference consists in making Z_1 and Z_2 (fig. 3) very small, so that the interfering current I is shorted in the interference producer itself and no longer flows through the asymmetrical circuit.

Which of these methods of suppression will give the most practical solution in each particular case will be discussed in a subsequent article.

How interference reaches the Radio Receiver

We have still to investigate how interference is produced in a radio receiver itself. The principal cause is the capacitive coupling between the aerial or the aerial lead-in and mains carrying an interference characteristic. It may also be possible for the mains transformer, whose primary is connected to the disturbing mains, to transmit the interference to the input circuit of the high-frequency component.

The fundamental circuit of that part of the receiver responsible for interference is shown in *fig. 8*, with an equivalent circuit at the side. C_t is the capacity between the primary and secondary windings of the mains transformer, Z_1 is the impedance of the tuned circuit, Z_{0a} the resistance of the earth connection, C_A the earthing capacity of the aerial and C_k the capacity between the aerial and the mains conductors.

The interfering voltage V_a is responsible for the amplifying valve as an interfering voltage.

If the capacity C_t between the primary and secondary windings of the transformer is not very small, the interfering voltage can also give rise to currents in two other circuits, *viz.*, those flowing:

3) through C_t, Z_{0a}, A_1 to A_0; and
4) through C_t, Z_1, C_A, A_2 to A_0.

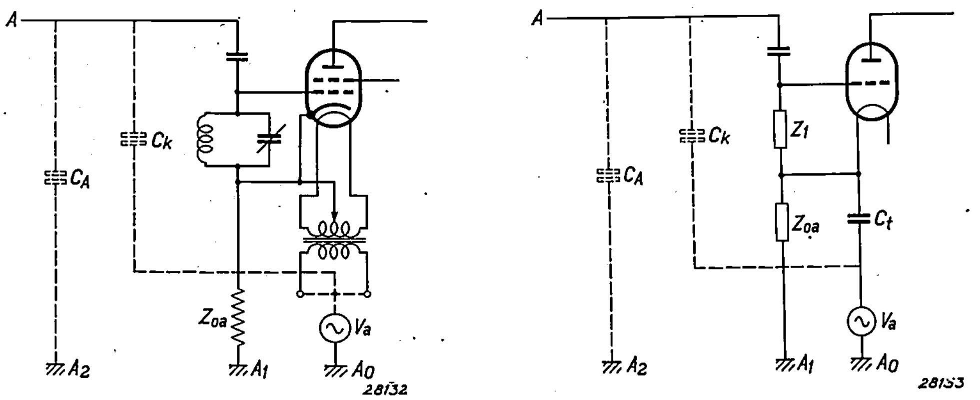

Fig. 8. Circuit of the input component of a receiver, with associated equivalent circuit. It is shown here how the asymmetrical interfering voltage V_a produces an alteration in voltage at the control grid of the high-frequency amplifying valve. This alteration in voltage is produced by currents which flow either through the capacity C_k between the aerial and the mains conductors or through the capacity C_t between the primary and secondary windings of the mains transformer.

appearance of a current in various circuits. Assuming that the capacity C_t is very small, as in modern receiving sets, then current will flow:

1) through C_k, C_A, A_2 to A_0; and
2) through C_k, Z_1, Z_{0a}, A_1 to A_0.

In the second case a voltage drop will appear at Z_1 and will reach the grid of the high-frequency valve. The latter current will again produce an interfering voltage at the grid of the high-frequency amplifying valve. This form of interference would not occur if Z_{0a} were zero. The importance of a good earth for the receiver, in addition to a low value of C_t, in order to prevent interference reaching the set through the mains transformer is thus self-evident.

FLUORESCENCE AND PHOSPHORESCENCE

by J. H. GISOLF and W. DE GROOT.

535.37

To supplement a previous article, the phenomena of fluorescence and phosphorescence (photoluminescence) of a series of different compounds are discussed here, *viz.*, lines and bands in fluorescence spectra of organic salts and organic compounds in solid and liquid solutions, and fluorescence and phosphorescence due to impurities, such as the luminescence of alkali-halide luminophors, ruby, sulphide luminophors, silicates, and fluorite. Tentative explanations regarding the mechanism of luminous emission in different cases are advanced on the basis of investigations of the afterglow.

Introduction

In a previous article [1]) the phenomenon of fluorescence was discussed with the aid of a few simple examples (resonance in monatomic and diatomic gases), some reference also being made to the fluorescence of solid substances (uranyl compounds) and liquids (eosin and fluorescein solutions), and to the so called luminophors. The difference between fluorescence and phosphorescence rests on the difference in the ability to re-emit selectively absorbed radiation, usually of a spectral composition different from that absorbed, in some cases only during the period of irradiation, *i.e.* fluorescence, and in other cases the persistence of an appreciable emission of light after irradiation has ceased, *i.e.* phosphorescence. Closer investigation has, however, shown that all bodies which can be rendered luminous by irradiation exhibit a finite afterglow period which may vary between a thousand-millionth of a second (10^{-9} sec.) and several months. These two phenomena are therefore frequently described as luminescence or more correctly as photo-luminescence to distinguish them from the emission of light under the action of cathode rays, X-rays, heat, mechanical deformation, chemical reactions, etc.

The phenomenon of photo-luminescence is quite common, in fact so widespread that with the exception of perfectly pure metal surfaces nearly all bodies can be assumed to luminesce when irradiated with light of suitable wave-length. A. van Wijk [2]) has already called attention to this in an article on the application of ultra-violet radiators for the investigation of luminescence. In lighting technology strongly luminescent bodies are naturally those to which the greatest interest attaches.

Luminescence may be a specific property of a certain atom or a certain atomic group or of a specific body in a pure and rarefied state, such as sodium vapour and iodine vapour discussed as examples in the previous paper. In many cases of this type the luminescence is retained in concentrated solutions or in the solid state; examples of this are salts of the rare earths, benzene, and uranyl salts. In other cases the luminescence of certain compounds is found only in the solid state, as in the platino-cyanides. But frequently, as seen in discussing fluorescence, a high concentration reacts adversely on luminescence, while dilute solutions exhibit a marked fluorescence, both in the liquid state (alcohol, water) and in the solid state (solutions in alcohol at —180 deg. C and in boric acid) and in the adsorbed state. Many bodies of organic origin probably owe their luminescence to that of impurities present in low concentrations. Inorganic luminophors whose luminescence is due to traces of certain admixtures, occur in Nature (ruby, fluorite) and can also be prepared artifically.

Although it is readily possible to co-ordinate the very diverse types of photo-luminescence in a rational classification, this does not signify that the mechanism of the phenomenon must be the same in all cases. On the other hand, a closer examination of the variation of the intensity of the afterglow with time has shown that *e.g.* the luminescence of a solution of rhoduline orange in sugar solution is due to an entirely different mechanism to that of a zinc-sulphide/silver luminophor. The various types of luminescence and the nature of the afterglow are discussed in further detail below.

Line and band fluorescence of crystals

Contrary to the behaviour of salts of other metals (alkalis, alkaline earths, copper, silver, etc.) the salts of the rare earths, both in the crystalline form as in solution, give small absorption bands in the visible and infra-red portions of the spectrum. In the solid state and at a low temperature (—180 deg. C.) these bands are seen to consist of a series of well-defined absorption lines with a definition comparable to that of the spectral lines obtained with gases at

1) Philips techn. Rev., 3, 125, 1938.

2) Philips techn. Rev., 3, 5, 1938.

low pressures. Irradiation with light having the wave-length of the absorption bands induces a fluorescence containing not only the same bands but also those with a longer wave-length. These absorption and emission bands have been classified in an energy-level scheme, an example of which is shown in *fig. 1*, *viz.*, for the trivalent terbium

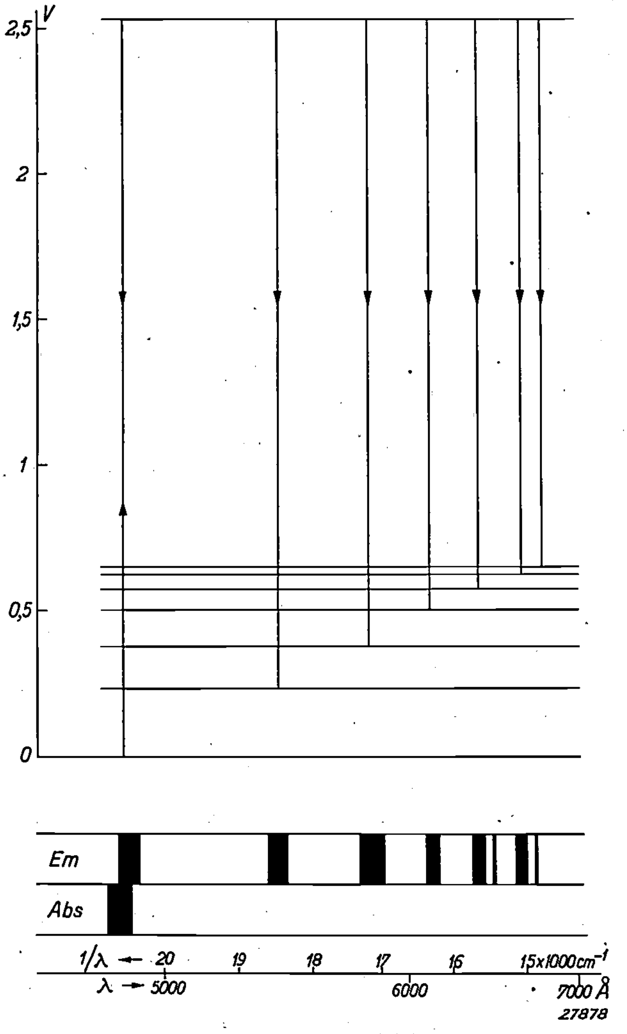

Fig. 1. Energy scheme of the ion Tb^{+++} and absorption and emission spectra of $Tb_2(SO_4)_3$, 8 H_2O according to Gobrecht (1937). Each of the transitions shown corresponds to a group of fine lines, each group being shown in the figure enclosed by a black rectangle. Particularly at low temperatures are these lines distinctly separated.

ion Tb^{+++}. The corresponding emission and absorption spectra are shown below, and the individual groups of lines are represented in the scheme as black rectangles. The resolution of each group into a number of fine lines (fine structure) is due to the action of electrical fields from the neighbouring ions in the crystal lattice. The most striking feature is that the width of each group is comparatively small (about 0.05 eV), *i.e.* much smaller than would be expected by analogy of the effect of electric fields on other atoms. Bohr's theory of the atom explains this phenomenon quite naturally, as in this theory corresponding energy-levels are ascribed to electronic orbits which lie at such a depth in the atom that they are only slightly affected by external electric fields, and which only in the case of the atoms of the alkaline earths occupy such positions that emission and absorption of light can be associated with them.

The uranyl salts also exhibit absorption and emission lines in solution and in the crystalline state, these lines being well defined in the solid state and at low temperatures. It has already been shown (Philips techn. Rev., 3, 130 and 131, 1938) that here the oscillation of the oxygen atom with respect to the uranium atom in the radicle UO_2^{++} produces a band spectrum and that the important characteristics of this spectrum can be explained on simple lines.

Similar to the two cases just discussed, still a third type of band fluorescence can very probably be ascribed to a specific atomic grouping, *viz.*, that observed in the double cyanides of platinum with other metals. Photoluminescence has been observed in the platinocyanides of barium, magnesium, lithium, sodium, potassium and rubidium; thus the use of barium platinocyanide for screens in radiography is well known, although of course this application naturally does not rest on photoluminescence in the narrower sense. Similar to the absorption spectrum, the emission spectrum consists of very broad ill-defined bands occupying very different positions with the various compounds. This indicates that if the platinocyanide group $(Pt(CN)_6^{--})$ is the focus of luminescence, it is nevertheless strongly influenced by its surroundings.

It is extremely difficult to make quite certain whether an observed luminescence is due to the pure substance or to admixtures present in such very small quantities that they are not capable of analytical determination. In the case of the platinocyanides, it is not very probable, although perhaps not quite impossible, that the effect is due to impurities.

Other compounds are also known which luminesce only in the crystalline state and whose absorption and emission bands are regarded as most likely due to the pure substances; these include a number of tungstates and molybdates. Of these calcium tungstate, which on irradiation with ultra-violet light below 2600 Å emits an intensive blue light, has found practical application. In this case the fluorescence is observed to increase in intensity with the purity of the substance, which is in favour of the assumption that luminescence is due to the pure substance, although it should be remembered in this connection that many "Lenard luminophors" exhibit their optimum luminescence at one exact low value of impurity concentration, while arbitrary quantities of impurities reduce the luminescence. The emission and absorption spectra of the tungstates and molybdates are again composed of very diffuse bands. This feature (not confined to these substances) makes the collection of quantitative data difficult, without which a deeper insight into the mechanism of luminescence of a solid is not feasible.

Luminescence of organic compounds

It can be generally stated that the majority of pure organic compounds, with the exception of the aromatic compounds, exhibit no or only very slight photoluminescence in the gaseous state.

The absorption spectra of many organic compounds, such as those of the aliphatic series, are continuous and accompanied by a number of diffuse bands towards the long-wave side. These bands indicate that there are certain "energised" (excited or activated) states in the molecules. If the energy of excitation or activation of such a state is greater than the binding energy of any bond in the molecule, the energy of the activated state may be utilised for the spontaneous separation of this bond (pre-dissociation). This transfer of energy within a molecule may in certain circumstances be promoted by electric fields, such as are due, for instance, to neighbouring molecules. This phenomenon of pre-dissociation which will occur the more readily the more complex the molecule, prevents the absorbed energy being re-emitted as fluorescent radiation. The absorbed energy can also be converted to thermal motion without any dissociation resulting (de-activation).

The most marked fluorescence among the aliphatic compounds is exhibited by the aldehydes and ketones (the carbonyl group $\overset{|}{\underset{|}{C}} = O$ appears to play an important part in this reaction).

Acetone vapour at atmospheric pressure consumes for pre-dissociation 17 per cent of the absorbed light quanta which are irradiated in the wavelength region of the discrete absorption bands, and only 3 per cent is emitted again as fluorescence. The remaining 80 per cent of the irradiated quanta are directly converted into heat.

Chemical reactions, also, can be induced by irradiation and this will also reduce the yield of fluorescence; thus with the aldehydes polymerisation accompanies fluorescence.

The number of aromatic compounds exhibiting fluorescence is very large. The benzene ring, the nucleus of all aromatic compounds, is very resistant to external influences, with the result that benzene and its derivatives possess a luminescence not only in the gaseous state but also in the liquid and solid states, as well as in a whole host of solvents.

The structure of the absorption and emission spectra of benzene closely resembles that of the diatomic iodine molecule already discussed. The absorption spectrum of benzene vapour consists of an extensive system of bands in the ultra-violet. This system of bands has been completely resolved and can be interpreted as the total sum of the series of oscillatory levels in the benzene ring belonging to different electronic states. The fluorescence spectrum, which is similarly located wholly in the ultra-violet, also contains a large number of bands and forms a continuation of the absorption spectrum in that the longest-wave absorption bands coincide with the shortest-wave fluorescence bands, as has already been seen in the case of the uranyl bands. No less than 450 small bands have been counted in the fluorescence spectrum of benzene.

As already stated above, luminescence is also exhibited by benzene when in solution, the bands in both the absorption and emission spectra being more or less indistinct depending on the solvent and the concentration, and being displaced slightly towards longer wave-lengths. The luminosity is a maximum at a certain concentration. Pure benzene in the liquid state has a weak luminosity, so that the strongest bands only are barely and very indistinctly visible. On cooling the benzene until crystallisation occurs, the bands become sharper again and the luminosity considerably increases. At the temperature of liquid air (—180 deg. C.) the definition again becomes comparable to that in the gaseous state (*fig. 2*).

Solid benzene exhibits a distinct afterglow, contrary to the same substance in the liquid and gaseous states in which the afterglow has a duration of only 10^{-7} to 10^{-8} sec. We shall return to this point again later.

The absorption and emission spectra of the simple benzene derivatives resemble those of benzene itself. The benzene ring retains its independence in behaviour, and the substituted radicles merely exercise a disturbing influence which is shown by a displacement of the absorption and emission spectra towards longer wave-lengths, although the general composition of the spectrum remains unaltered. At the same time the bands become less sharp and the whole spectrum gradually merges into a continuous one. The luminescence of the benzene group is thus closely comparable to that of the uranyl group.

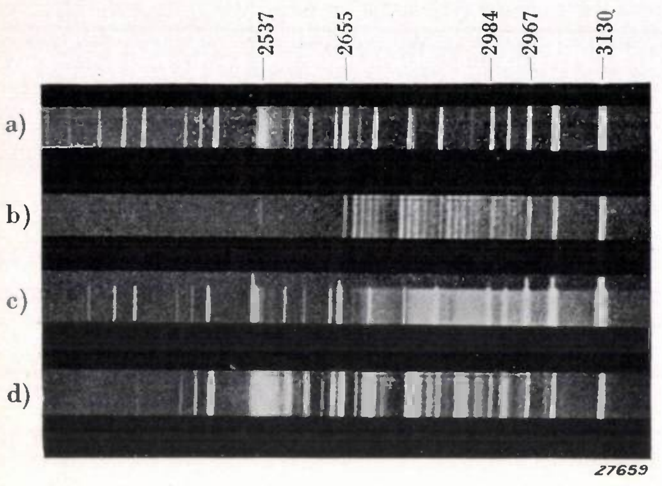

Fig. 2. Fluorescence of benzine (according to Pringsheim).
a) Mercury spectrum.
b) Fluorescence of benzene vapour.
c. Solid benzene at 0 deg. C.
d) Solid benzene at —180 deg. C.

Although very important from a theoretical point of view, the benzene derivatives are not the most striking examples of the photo-luminescence or organic compounds, since their emission spectra are situated mainly in the ultra-violet. Compounds with a visible emission spectrum have been known for many years and have been closely investigated, as for instance the highly-fluorescent solutions of certain aromatic dyes, such as fluorescein, eosin, rhodamine, etc., as well as quinine sulphate and also chlorophyll, which is the green colouring matter of plants of paramount importance in plant physiology. For these substances, neither the absorption spectrum nor the emission spectrum reveals any generalised characteristics, for both spectra depend largely on the solvent and consist of a number of usually very indistinct bands. These substances are, therefore, little suitable for an investigation and systematic co-ordination of the spectra.

Afterglow and quenching of fluorescence

The light yield of fluorescing aromatic dyes is usually a maximum at a very low concentration of the solution; as the concentration increases the yield diminishes very rapidly, which may be due to various causes. Nevertheless it has been found that in all cases this effect is the less marked the lower the mobility of the molecules of the solution, *i.e.* the more viscous the solvent. Many substances, which exhibit no or only a slight photoluminescence in liquid solvents, acquire this property to a pronounced degree in solid, vitreous solutions.

Vitreous luminescent solutions are easy to prepare. If an alcoholic solution is cooled, it becomes viscous at —125 deg. C. and at —130 deg. C. changes to a hard vitreous mass, which with a certain amount of care can be further cooled to —190 deg. C. without crystallisation occurring. Many solid solutions of this type when irradiated with a suitable wave-length are found to exhibit an intense luminescence which persists for several seconds after irradiation ceases.

This is found not only when alcohol is the solvent, but also with many other solvents. Low temperature is not a *sine qua non* for this behaviour. Solutions of many aromatic compounds in molten boric acid, which solidifies to a vitreous mass already at room temperature, also show an intense phosphorescence. In many cases the absorption and emission spectra of a substance dissolved in solid alcohol are the same as that obtained on solution in boric acid.

Various gels, such as gelatine, silica gel, albumen, cellophane, all kinds of fibrins, such as cotton wool, blotting paper, etc., can be used as adsorbents. This phenomenon is obviously again dependent on a fixation of the molecules, whereby they are protected against disturbing influences. Many substances which exhibit no luminescence whatsoever in liquid solutions fluorescence and phosphoresce strongly when they are adsorbed.

If, for instance potassium iodide is added to a solution of quinine sulphate, in sufficient quantity entirely to inhibit fluorescence when irradiated with ultra-violet light [3]), and a grain of silica gel is then dropped into the solution, the grain will show a strong fluorescence immediately it passes through an incident beam of light. By the strongly negative electrical charge on the surface of the gel, the ions $J^{-}(H_2O)$, which adversely react on the fluorescence, are kept away from the molecules of the quinine sulphate which are adsorbed by the gel.

Adsorption is probably associated with a spatial orientation of the molecules resulting in a polarisation of the light emitted. If the molecules of a dye are adsorbed by a powerful anisotropic base, such as cellophane, and if this is irradiated with non-polarised light, polarisation of the fluorescent light will occur. This orientated adsorption can probably be regarded as a transition case between the vitreous solutions and the luminophors which are produced when aromatic molecules

[3]) *v.* Philips techn. Rev., **3**, 136, 1938.

are introduced into foreign crystal lattices.

Many crystal lattices are apparently very suitable for taking up usually very small quantities of foreign molecules; examples of this are carbohydrates, bezoin acid and phthalic acid. Anthracene is also a well-known example, pure anthracene exhibiting a blue fluorescence, but ordinary commercial preparations always showing a very intense luminescence in the green (*fig. 3*), this being caused by traces of impurities, *viz.*, naphthacene, present in the lattice. Naphthacene is also an excellent example of the polarisation of light emitted by molecules occluded in crystal lattices.

In certain cases the afterglow has been closely investigated as a function of the time. In the case of dilute solutions, *e.g.* of fluorescein the duration of the afterglow is very short, and it is found that the time taken for the intensity to be reduced to one half (the half-life period) is of the order of 10^{-9} sec. Organic substances in the solid of viscous state, *e.g.* rhoduline orange in sugar solution, have an afterglow lasting several seconds. The intensity in this case can be represented with great accuracy by the exponential law:

$$I = I_0\, e^{-t/\tau}, \qquad (1)$$

where the constant τ is determined by the temperature, the viscosity and the concentration of the organic substance. It follows from this that we are dealing with a spontaneous phenomenon, or expressed chemically, with a mono-molecular reaction; the excited molecules return to their initial state according to a simple probability law, at the same time radiating a light quantum. The fact that the fluorescence of ordinary solutions is of such short duration and that of solid and viscous solutions is so much longer has been explained by the assumption that the excitated molecule directly after absorption passes over into a metastable condition with a very low transformation probability. This problem has not been entirely cleared up.

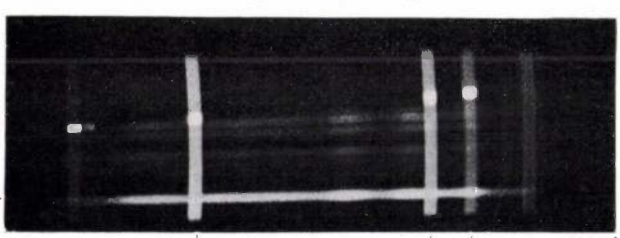

Fig. 3. Fluorescence of anthracene containing an impurity. A plate coated with anthracene was irradiated with the light from a mercury lamp decomposed into its spectrum, the long wave lengths above and the short wave lengths below. The fluorescent light emitted was resolved into a horizontal spectrum by means of a second spectrograph. Starting from the bottom the fluorescence is shown under the action of the mercury lines 2537, (2967), 3650, 4047, (4358) Å, the bracketed lines being comparatively weak. In the latter case only the bluish-green (*c*, *d*) bands are visible, and in the other cases also the blue and violet bands (*a*, *b*). The lines 5461 and 5770 Å give no fluorescence.

Inorganic "impurity" luminophors

Closely analogous to the organic "impurity" luminophors described above, are the many inorganic crystalline luminophors, which owe their luminescence to certain admixtures; a few examples of these are discussed below.

If a trace of a similar halogen compound of another metal, such as thallium, lead copper or silver, is added to an molten alkali halide, *e.g.* KCl, and the molten mass is allowed to crystallise out, a change is observed to have taken place in the absorption spectrum. The absorption spectrum of the alkali halide is composed of a number of absorption bands lying for the most part in the ultra-violet. If one of the above-mentioned metals has been added to the lattice of the alkali halide, new absorption bands make their appearance, which are characteristic for both the added metal as for the halogen. A strong luminescence is observed if light of the wave-length of these absorption bands is used for irradiation. After the termination of the period of irradiation the intensity of the afterglow decreases purely exponentially with a half-life period of about a minute.

The source of the luminescence is here a complex ion of the added metal with the halogen (*e.g.* $TlCl_4^{---}$, $PbCl_4^{--}$). which follows from the fact that the same complex ions exhibit the same fluorescence in solution. If, for instance a concentrated solution of a thallium halide is mixed with a concentrated solution of a corresponding alkali halide, the absorption spectrum of the solution is found to contain exactly the same absorption bands as the phorphorescent crystal of the alkali halide, while the solution also fluoresces.

The emission spectrum of ruby (aluminium oxide containing traces of chromium of the order of 10^{-3} per cent) consists of a large number of lines in the red, some of which occur also as absorption lines and exhibit an afterglow period of 10^{-2} to 10^{-3} sec. These lines are found to be transitions between the known energy levels of the trivalent chromium ion. In the free ion this transition would be prohibited, yet under the action of the electric fields of the neighbouring ions a finite transition probability is obtained, an interpretation which is confirmed by an investigation of the magnetic resolution of the lines.

Since the luminescence as well as the absorption of ruby is determined by the angle between the direction of the incident ray and the axes of the crystal, it is very probable that the chromium ion here forms a part of the crystal lattice and is not located haphazardly within the parent material.

Zinc sulphide exhibits a weak luminescence already in the pure state, so that with some reservation this substance can be included among those luminescent pure bodies of the type discussed above. The very intense luminescence of zinc-sulphide luminophors is however due to certain metals which are admixed with the zinc sulphide, frequently in extremely small proportions. (A concentration of 10^{-6} parts of copper is sufficient to produce a marked luminescence).

With these luminophors various phenomena have been observed which are encountered in none of the substances discussed above. In all examples dealt with up to this point the luminescence of an impurity is only observed when the wavelength of the incident light lies within the absorption spectrum characteristic of the impurity in question. But in the case of the sinc-sulphide luminophors, light which the zinc sulphide crystal itself absorbs is converted largely into a radiation which is characteristic for the sporadically-occurring impurity.

This remarkable behaviour to which the pronounced luminosity of these luminophors is due is accompanied by a second phonomenon, which has also not been observed with the substances discussed above. The zinc-sulphide luminophors are perfect insulators in the dark, but when irradiated with light which they can absorb, they become electrical conductors (photo-conductivity).

The connection between these two properties can be correlated on the following lines: During absorption of a light quantum by the zinc sulphide lattice, which is built up of a divalent positive ion Zn^{++} and a divalent negative ion S^{--}, an electron is removed from an S^{--} ion leaving an S^{-} ion; this electron attaches itself to a Zn^{++} ion which is thus converted into a Zn^{+} ion. The electron does not, however, always remain attached to the same zinc ion, but may migrate from one ion to another. Similarly a sulphur ion which has lost an electron may take up another electron from one of the neighbouring sulphur ions, which in its turn is converted to an S^{-} ion. The gap formed by the absence of an electron will thus traverse the lattice in the same way as the electron itself. These two factors are responsible for the electrical conductivity.

When by recombination an electron again fills a gap, the absorbed energy is again liberated and it is found that this energy is converted into radiation with a high quantum yield, which is characteristic of the activating metal sporadically present. The exact mechanism of this transformation of energy is still unknown, but tentatively the best explanation which can be advanced is that the metal impurity acts as a catalyst in the recombination of the electron to fill the ionic gap, as a result of which a part of the energy of recombination is transmitted as activation energy to the metallic impurity and again emitted as radiation by this component.

This explanation receives strong support from a third phenomenon, which again has not been observed with any of the luminophors previously discussed. The intensity of the afterglow when plotted as a function of the time (*fig. 4*) is not an

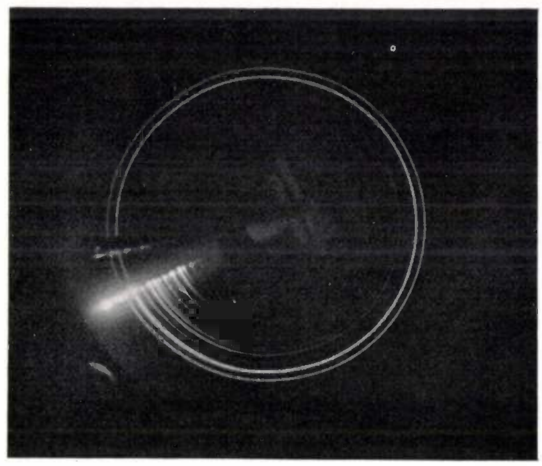

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Fig. 4. Photograph of a revolving disc coated with an ZnS-Cu luminophor. A mercury spectrum is projected on one radius of the disc, the longer wave lengths giving complete circles (long afterglow), and the shorter waves incomplete circles (short afterglow); the rotation time of the disc was of the order of several seconds.

exponential curve, at least during the first few seconds, but approximates more to a "hyperbolic" law:

$$I = \frac{I_0}{(1 + t/\vartheta)^2} \cdot \quad \cdots\cdots \quad (2)$$

In physical chemistry, this behaviour is recognised as a typical time function of a bimolecular reaction. If two types of atoms, A and B, are present in the same concentration, and a reaction:

$$A + B \rightarrow AB$$

takes place, then the reduction in concentration n with time is given by the expression:

$$-\frac{dn}{dt} = bn^2. \quad \ldots\ldots (3)$$

The solution of this differential equation is:

$$n = \frac{n_0}{1 + bn_0\, t}. \quad \ldots\ldots (4)$$

If A is an electron and B a gap due to the absence of an electron, and if the filling of the gap by A causes the emission of a light quantum, then if this combination takes place according to the law of bimolecular reactions, the luminous intensity is given by:

$$I = -\frac{dn}{dt} = \frac{b\, n_0^2}{(1 + bn_0 t)^2}. \quad \ldots (5)$$

This expression is identical with (2) on substituting:

$$I_0 = bn_0^2, \qquad \vartheta = 1/bn_0 = 1/\sqrt{bI_0}.$$

The very long afterglow observed with many zinc sulphides on irradiation with light of suitable wave length is very probably due to a more complex mechanism.

In conclusion, reference must be made to a few other cases in which the luminescence is similarly due to traces of foreign metals introduced into a crystal lattice. Thus calcium tungstate, which has already been referred to as a luminescing substance in the pure state, can be caused to give a luminescence by adding *e.g.* traces of samarium, which contains in addition to the (blue) tungsten spectrum also the (red) samarium spectrum.

Also many silicates, which are not luminescent in the pure state, luminesce to a marked degree when traces of manganese are added. Both the activated tungstates and the luminescent silicates have many technical applications.

Another example of a line and band fluorescence due to the presence of traces of impurities is the mineral fluorite, which in the pure state (CaF_2) is not luminescent. Traces of the rare earths, *e.g.* europium, produce highly-fluorescent crystals which are widely found in Nature and which can also be prepared in the laboratory (*fig.* 5). Whether

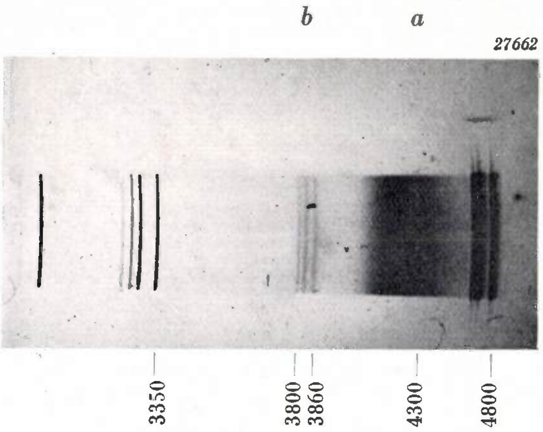


Fig. 5. Fluorescence spectrogram of a blue-fluorescing fluorite on irradiation with light from a zinc lamp (2138 Å). The sharp spectral lines are other zinc lines which also are present, as the light used for irradiation was not sufficiently monochromatic. *a* - blue bands (Eu), *b* - line fluorescence (Tb?).

this luminescence is due to the alkali halides or to the zinc sulphides has not yet been definitely established.

APPLICATIONS OF CATHODE-RAY TUBES. III.

by H. VAN SUCHTELEN. 621.317.755: 621.385.832

Investigation of high-frequency phenomena

In the previous article of this series, reference has already been made to the registration of a high-frequency oscillation of 470 kc/s [1]) with the oscillograph. The cathode-ray oscillograph is, of course, the most suitable apparatus for the direct investigation of high-frequency phenomena, and a number of further examples of its application in this direction are described below.

Measurement of the Depth of Modulation

The shape of the curve of the individual period of a high-frequency oscillation is usually not as interesting as with a low-frequency oscillation. High-frequency oscillations are generated and handled primarily in resonance circuits, so that a pure sinusoidal form is obtained in nearly all cases. In radio technology, however, considerable interest attaches to the behaviour of the oscillation over an interval of time which is large as compared with the natural period of the oscillation. The high-frequency oscillations encountered in practice are nearly always more or less modulated with respect to period, and the form of the modulated oscillation can be rendered directly visible with the cathode-ray oscillograph.

In general the frequency of modulation is very much smaller than the frequency of the oscillation itself, so that to investigate a complete modulation period the image studied must contain a very large number of periods of the high-frequency oscillation.

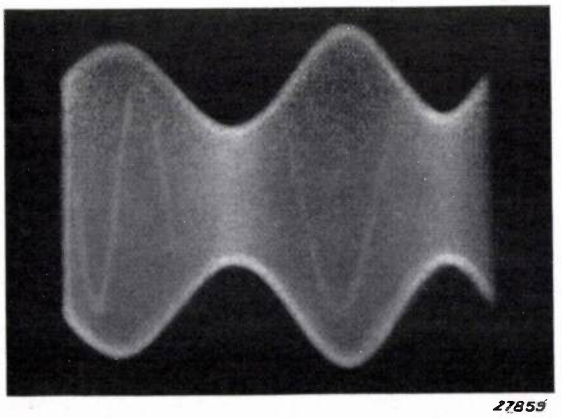


Fig. 1. Oscillogram of a 300-kc/s carrier wave modulated sinusoidally with a 400-c/s frequency. The modulation depth is about 35 per cent. With the sawtooth frequency of 200 c/s used here, the waves of the high-frequency oscillation can no longer be distinguished separately, although owing to the very rapid flyback a few high-frequency oscillations can be clearly picked out.

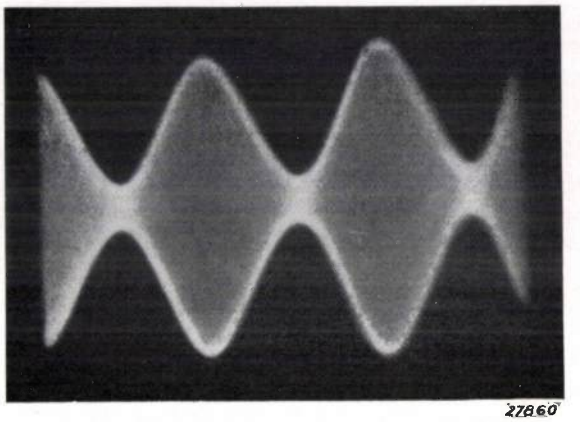


Fig. 2. Oscillogram of the same carrier wave as in fig. 1, but with a modulation depth of 75 per cent.

As a rule this number is so large that the individual curves can no longer be distinguished separately and a uniformly illuminated surface only is obtained, as shown in the registration of a sinusoidally modulated high-frequency reproduced in *fig. 1.* That the high-frequency oscillation cannot itself be differentiated is, in fact, not a disadvantage in most cases since only the envelope of the trace is of interest. A very important magnitude, the depth of modulation, can thus be read off very accurately from fig. 1, and in the case under consideration was 35 per cent, which signifies that the variation in the high-frequency amplitude was 35 per cent of the mean amplitude. In *fig. 2* an oscillogram is reproduced with a depth of modulation of 75 per cent.

In both cases the horizontal motion of the light spot was produced by a sawtooth voltage varying linearly with time [2]). To obtain a steady image the period of this time-base voltage must be equal to an integral multiple of the modulation period, and in figs. 1 and 2 this multiple was two and three respectively.

Not only is the depth of modulation obtained in this way, but also the complete geometrical form of modulation, showing, for instance, whether it is

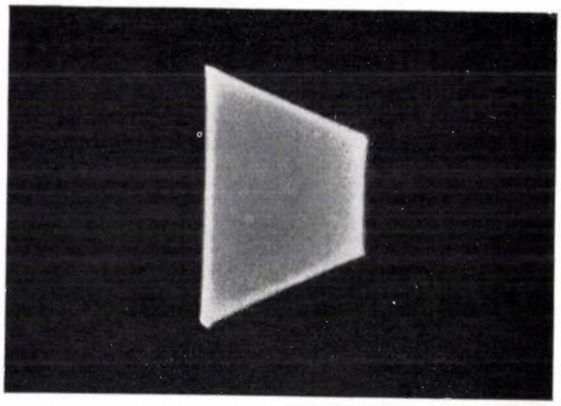
[1]) Philips techn. Rev., **3,** 150, fig. 6, 1938. Another direct oscillogram of high-frequency oscillations is included in an article by L. Blok dealing with high-frequency oscillations in sodium lamps, Philips techn. Rev., **1,** 87, 1936. Fig. 2.

[2]) Regarding the generation of this voltage in oscillographs, *cf. e.g.* Philips techn. Rev., **1,** 147, 1936.

sinusoidal or not. It is, of course, assumed that in all cases modulation is with a sound of constant intensity, as is frequently the case in laboratory measurements.

Other conditions exist during measurements when working on radio transmitters, for the high-frequency oscillation is then usually modulated by speech or music. If an attempt is made to obtain an oscillogram of the type shown in figs. 1 or 2, the image will be found subject to such continuous and rapid changes that it can no longer be steadied by a suitable choice of the time-base frequency. Nevertheless, it is still possible to obtain an indication of the instantaneous depths of modulations in this case also, *viz.*, by eliminating the modulation frequency; this is done by employing for the horizontal deflecting voltage the low-frequency alternating voltage used for modulating the high-frequency oscillation in the transmitter, instead of a sawtooth voltage.

a



b

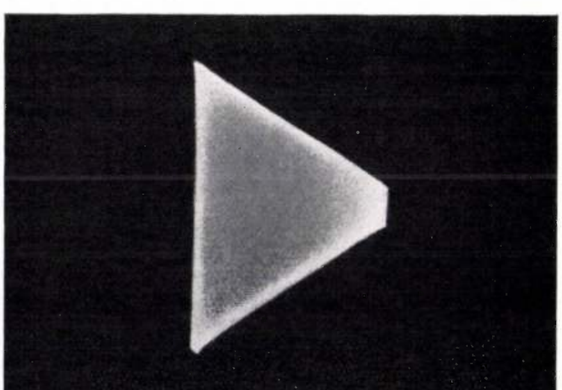


Fig. 3. Oscillogram of the same modulated high-frequency signal as in figs. 1 and 2, but plotted as a function of the instantaneous value of the modulation voltage instead of time. *a*) 35 per cent modulation, as in fig. 1; *b*) 75 per cent modulation, as in fig. 2.

The high-frequency amplitude is then no longer reproduced in the oscillogram as a function of the time, but as a function of the instantaneous value of the modulated voltage. A diagram obtained by this method is shown in *figs. 3a* and *3b*, which correspond to figs. 1 and 2. In these oscillograms the depth of modulation may be deduced from the lengths of the two vertical bounding lines, while the two sloping lines indicate if any distortion occurs in the modulating stage of the transmitter. A pure linear relationship between the high-frequency amplitude and the modulation voltage is naturally desired.

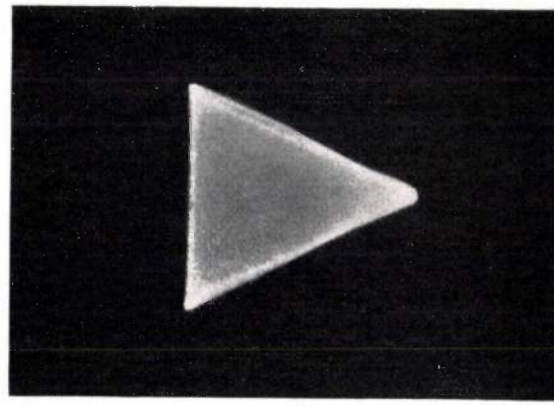


Fig. 4. Oscillogram of the same carrier wave as in figs. 3*a* and *b*, but with a depth of modulation of 95 per cent.

If modulation is not of constant depth as in figs. 3a and 3b, but is produced by speech and music, fig. 3 will pass over into *fig. 4* or *vice versa*, the change being in fact extremely rapid. The appearance of the diagram is determined by the boundaries, which correspond to the maximum depth of modulation obtained. The sloping sides of the various trapezia will always form a single straight line or a curve, while the vertical sides move rapidly to and fro. A triangle is obtained at the moment when the depth of modulation reaches the theoretical maximum value of 100 per cent (fig. 4).

The depth of modulation can also be indirectly measured in various ways and read off on an indicator, but owing to the inertia of the instruments used it is not possible to determine the very short-period maxima which may occur if the depth of modulation is not limited. The method described above, employing a cathode-ray oscillograph, can be used in this case with excellent results. The maximum depth of modulation occurring during a part of the sending interval can in fact be quite easily recorded photographically by this means, and such records may prove useful in the operation of radio transmitters.

Investigation of frequency modulation

In the above method of amplitude modulation the amplitude of the oscillation is altered with respect to time. The frequency also can be made to vary with regard to time; this method of modulation is termed frequency modulation, and can be used for the transmission of speech and music in the same way as amplitude modulation.

In many cases, frequency modulation appears in the form of an undesirable disturbance as a result of the inadequate smoothing of the feed voltages for the amplifying valves. The modulation frequency is then as a rule equal to the mains frequency or a simple multiple thereof.

It is a comparatively simple matter to measure a frequency modulation with the cathode-ray oscillograph provided it is not too small. If the high-frequency is oscillographed over a period of the modulation frequency, the frequency of the oscillation cannot be obtained by direct computation at every point of the oscillogram as the latter is traced much too compactly, and since the frequency of the carrier wave has an altogether different order of magnitude to that of this mains voltage. However, the heterodyne method, which is extensively used in radio circuits, offers a comparatively simple means for reducing or increasing a given frequency in an arbitrary constant ratio [3]), thus if two alternating voltages are applied simultaneously to the grid of an amplifying valve having a curved characteristic, the anode current will be found to contain, in addition to the original frequencies, also the additive and differential frequencies.

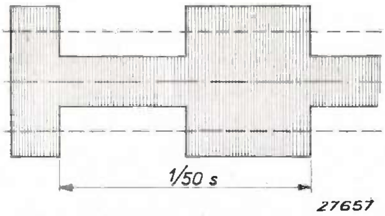


Fig. 5. Oscillogram of a carrier wave of 10^7 c/s with rectangular modulation of 50 c/s. This modulation is particularly suitable for investigating whether the carrier frequency depends on the modulation voltage (see fig. 6).

If, for instance, the frequency modulation of a high-frequency current of 10^7 c/s is to be studied, a mean differential frequency of $\nu = 1000$ c/s can be obtained by superposing an oscillation with a constant frequency of $10^7 + \nu$. This differential frequency will then undergo the same absolute variation as the initial oscillation of 10^7 c/s. If the differential frequency is oscillographed over the interval of a modulation period, $^1/_{50}$ sec., an average of ν : 50 oscillations can be counted over this interval, while if ν is a frequency of 500 to 1000 c/s, these oscillations can quite easily be picked out and the variations during a modulation period can be determined directly.

As an example, an oscillogram is reproduced in *fig. 6* which was obtained in the following way: A high frequency oscillation of 10^7 cycles was modulated by a 50-cycle frequency in such a way that its amplitude had the rectangular form shown in *fig. 5*. Owing to a certain lack of efficiency of the circuit used, a frequency modulation was also obtained.

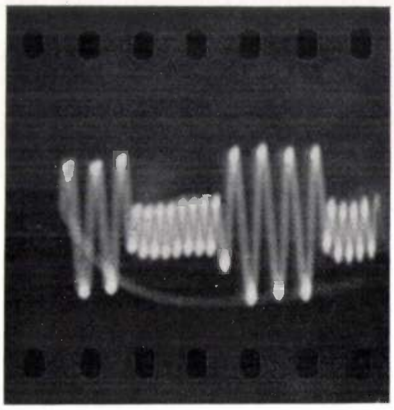


Fig. 6. Oscillogram of the voltage obtained by applying the heterodyne principle to the carrier in fig. 5 and an auxiliary voltage of 10^7 + approximately 500 c/s. The differential frequency increases slightly with diminishing modulation voltage.

The oscillation to be investigated together with the signal of an auxiliary generator was passed to a mixing valve, and the oscillation in the anode circuit passed through a suitable filter to an oscillograph. The time-base voltage of the oscillograph was of the sawtooth type and had a frequency of 25 c/s, which was synchronised with half the frequency of the modulation voltage of 50 c/s. The frequency of the auxiliary generator was then adjusted so that a steady image (fig. 6) of two modulation periods was obtained.

The amplitude modulation is quite distinct in this oscillogram, and at the lower amplitude eight successive high-frequency oscillations and at the high amplitude four can be counted. As these figures were obtained over half a modulation period *i.e.* over $^1/_{100}$ of a second, the frequencies were 800 and 400 c/s respectively. The deviation from the mean value was thus plus or minus 200 c/s. The same deviation was also found in the case of the original oscillation of 10^7 c/s, thus showing that a very small percentage of frequency modulation can be determined in this way. In the case of sinusoidal instead of rectangular modulation a similar

[3]) Regarding the heterodyne principle, *cf.* Philips techn. Rev., **1**, 76, 1936.

alteration is naturally more difficult to count up in the oscillogram, but it is still quite distinct.

"Blocking" of oscillators

Some oscillograms obtained in the investigation of "blocking" of a back-coupled oscillating valve are given as a third example of the application of the cathode-ray oscillograph in high-frequency work. The principle of back-coupling is shown in *fig. 7*. An oscillation can be maintained in the tuning circuit LC by introducing a sufficiently tight coupling between the coils M and L which compensates the losses in the oscillating circuit from the anode circuit. The alternating voltage applied to C produces an alternating voltage in the anode circuit by reacting on the control grid, and at the same time, by rectification in the grid circuit, the condenser C_g receives a charge which is roughly proportional to the amplitude of the oscillation and which makes the grid negative with respect to the cathode.

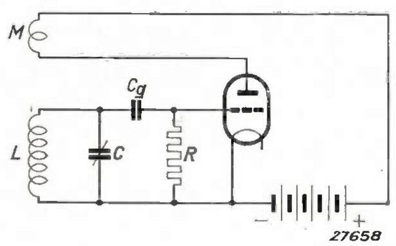


Fig. 7. Principle of an oscillating circuit.

The higher the negative bias on the grid the lower will be the amplification of the valve, and hence the smaller the energy passed to the grid circuit from the anode circuit. In ordinary circumstances an equilibrium will be reached in which the amplitude of the oscillation and the negative bias of the grid are so high that the losses in the tuning circuit are just compensated.

If the coupling between M and L is too tight, the equilibrium may be exceeded and as a result the negative bias of the grid will become so great that oscillation ceases. The negative charge will then be dissipated through R in *fig. 8* and the circuit will again start oscillating. This process is usually repeated with a frequency lying in the audible range and becomes apparent as a crackling in the loudspeaker of the receiver.

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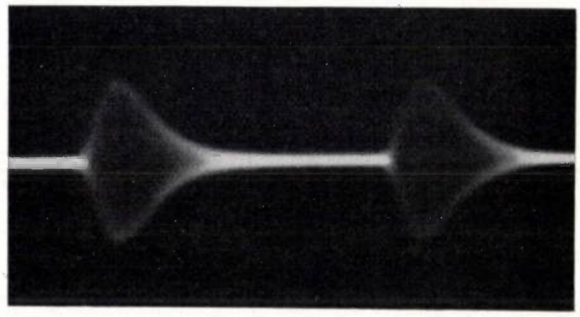

Fig. 8. "Blocking" of an oscillating circuit. If the amplitude of the oscillations exceeds a certain critical value, the valve suddenly ceases to oscillate. The oscillations then start decaying exponentially. Only when the oscillations have nearly become zero does the valve start oscillating again.

An oscillogram is shown in fig. 8 of the voltage in the tuning circuit when these conditions obtain, the tuning circuit being tuned to 1000 kc/s. It is seen that oscillation at first increases very rapidly and then beyond the critical limit decays exponentially.

In certain cases, a second tuning circuit of similar tuning characteristics may be coupled to the oscillating circuit. If the coupling between the two circuits exceeds a certain critical value, part of the oscillation energy will flow to and fro between the two oscillating circuits and as a result the voltage will exhibit fluctuations of the type shown in *fig. 9* for one of the oscillating circuits.

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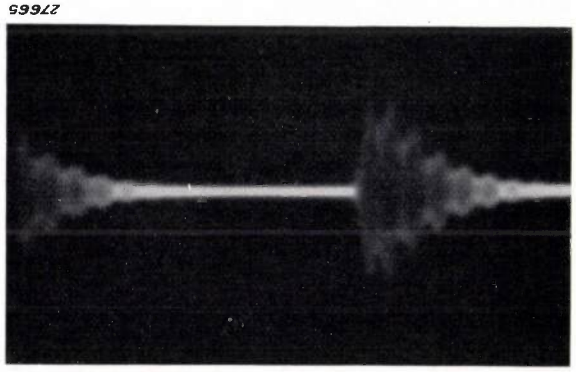

Fig. 9. "Blocking" of an oscillating circuit with a second circuit coupled to it. During the repeated decay of the oscillation voltage, fluctuations occur since a part of the oscillating energy passes periodically from one oscillating circuit to the other and *vice versa*.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS GLOEILAMPENFABRIEKEN

1294*: H. Bruining: On the emission of secondary electrons by solid substances (dissertation, Leiden, April 1938, 119 p.).

The writer contributed an article on this subject in the March number of this periodical (Philips techn. Rev. 3, 80, 1938).

1295: C. J. Dippel and J. H. de Boer: Interlamellare Quellung von Vakuumsublimierten CaF_2-Schichten durch adsorbiertes Caesium (Rec. Trav. chim. Pays Bas **57**, 277 - 290, Mar. 1938).

The ratio is determined between caesium and iodine molecules adsorbed on layers of Ca F_2 which have been deposited by sublimation in a vacuum and then sintered. Such sintered layers have fewer capillaries and interlamellary spaces, because the adsorption of iodine is much less intense than on layers sublimed but not sintered. Caesium however is adsorbed by both surfaces to the same degree, and the adsorption only becomes less when the surface has been heated for several hours at 450° C. The different behaviour of the adsorbing surface with respect to iodine and caesium cannot apparently be explained as a poisoning by adsorption of water molecules. It is shown that a sintered surface upon which caesium has been adsorbed, adsorbs just as much iodine after the caesium has been removed as before sintering. Caesium is therefore able, by means of an interlamellary swelling, to bring the surface back to its original size without reducing the dimensions of the crystals.

1296: J. E. de Graaf: The diagnosis of casting faults with the help of X-rays (Gieterij **12**, 31 - 35, Mar. 1938).

After an introduction on the X-ray examination of macro structures (*cf.* Philips techn. Rev. **2**, 315, 1937), the diagnosis of casting faults is discussed (*cf.* Philips techn. Rev. **2**, 377, 1937). Special attention is given to the chief faults, several characteristic forms of which are treated.

*) An adequate number of reprints for the purpose of distribution is not available of those publications marked with an asterisk. Reprints of other publications may be obtained on application to the Natuurkundig Laboratorium, N.V. Philips' Gloeilampenfabrieken, Eindhoven (Holland), Kastanjelaan.

1297: J. L. Snoek: Kristallorientierung und interkristalline Korrosion (Z. Metallk. **30**, 94, Mar. 1938).

The boundary between two crystals of nickel iron with almost the same orientation oxidizes less easily than is the case with greater differences in orientation. The importance of this phenomenon is pointed out in connection with intercrystalline corrosion.

1298: K. F. Niessen: Über das Feld einer vertikalen Halbwellenantenne in beliebiger Höhe oberhalb einer ebenen Erde beliebiger Konstanten (Ann. Physik **31**, 522 - 539, März 1938).

The Hertz vector is calculated for a vertical half-wave aerial situated above a plane earth with an arbitrary dielectric constant and conductivity. This calculation is carried out to a degree of approximation such that it is permissible to use the formula for a much smaller distance of the transmitter than that to which the reflection formula applies for the case of dipole radiation .The method may be equally well used for an aerial of quite other than half wave length.

1299: J. van Niekerk and F. Franken: Hypervitaminosis-D durch grosse Gaben tierischen, antirachitischen Vitamins an Küken. (Abt. brevia Neerl. **8**, 13 - 15, Febr. 1938).

In this article the influence is discussed of excessive doses of antirachitic vitamine of animal origin on the growth and death rate of chicks.

In June 1938 appeared:

Philips Transmitting News 5, No. 2:

C. G. A. von Lindern and G. de Vries: An ultra-short telephone link between Eindhoven and Tilburg.

Four shortwave broadcast transmitters, type KVFH 10/12a, for British India.

Tj. Douma, Internal inductance of coils and its influence on the temperature coefficient of the coil.

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RELATING TO THE PRODUCTS, PROCESSES AND INVESTIGATIONS OF
N.V. PHILIPS' GLOEILAMPENFABRIEKEN

EDITED BY THE RESEARCH LABORATORY OF N.V. PHILIPS' GLOEILAMPENFABRIEKEN, EINDHOVEN, HOLLAND

RADIO RECEIVERS WITH PUSH-BUTTON TUNING

by A. HOROWITZ and J. A. VAN LAMMEREN. 621.396.62

After a general survey of the various systems of push-button tuning (*i.e.* the reproduction of a number of previously fixed tunings by means of push buttons) a study is made of the system developed by Philips. In this system the rotating condenser is driven by a motor. The motor is started by pushing a button, and at the same time a stop is brought into position on a tuning cylinder corresponding with the button pushed. This stop determines the final position of the rotating condenser. Upon repeated pressing of the button the corresponding tuning is reproduced with a maximum deviation of 0.5 kilocycle per sec.

One of the most striking features in the development of radio receivers has been the attempt to simplify operation. This attempt on the one hand expresses the fact that the radio set, from a somewhat mysterious instrument which impressed the layman because of the technical achievement which it represented, has become an article of daily use. On the other hand the simplification of operation, and particularly of tuning, has become an urgent requirement, because inherent distortion of reproduction is now so much less that incorrect tuning by the listener is more noticable than it used to be. In other words, it is now more difficult for the listener to tune his set in such a way that the optimum effect is obtained. The resonance curve of present sets is smooth, and no wider than strictly necessary, which makes the adjustment to a definite wave length fairly critical. At the same time in sets with automatic volume control the listener can no longer be led by the intensity of the signal received [1]), but must judge the fidelity of reproduction which is greatest at the correct tuning point (lowest noise level). It is obvious that this offers certain difficulties and that such a receiver, especially when tuned by unskilled hands, would often be used at an incorrect tuning point. In order to avoid this, various measures have been applied from time to time. We may mention here visual indication, artificial reduction of sound intensity, resonance and tuning by touch.

An important advance is the completely automatic tuning realized in receivers operated by push buttons. The task of the listener in tuning is here reduced to the proverbial simplicity of "pressing the button". In general however the listeners would find it less enjoyable if they were compelled to listen only to a number of stations chosen by the manufacturer. This requirement of freedom of choice can be satisfied by retaining in the set, in addition to the push-button tuning, the ordinary dial tuning, and providing further that the push-buttons themselves can be set on different stations.

The advantage of the push-button tuning may then be summed up in the following way. The user is able to reproduce a number of tunings, which he himself has previously determined, by simply pressing the proper button. The accuracy of reproduction must of course satisfy high requirements if the advantage of the whole system is not to be illusory.

Various systems of push-button tuning

The various sytems of push-button tuning may be divided into two groups: those with electrical and those with mechanical pre-adjustment. The systems with motors form a special division of the second group.

In the systems with electrical pre-adjustment, the button which is pressed by the listener only serves as a switch. Each button switches in several

1) This is explained in Philips techn. Rev. 1, 264, 1936.

semi-variable capacities or inductances which tune the set on a definite wave length. This method has the advantage that the push button need only be pressed lightly and need not be pushed in deeply. On the other hand there are various disadvantages. The setting of the push buttons is carried out by tuning the semi-variable components by means of screws. This is rather elaborate, and with the better kinds of receivers with more than two tuning circuits, in which therefore more than two parts must be adjusted for every station, the manipulation is impracticable for the ordinary user. Another difficulty of these sets is the alteration of the previously fixed tuning elements due to external influences such as shocks, variations in temperature, moisture, etc. In order to prevent the push buttons becoming completely detuned at times due to these causes, permanent capacities or inductances which are practically insensible to the influences in question are often used in circuit with the semi-variable elements. The latter then need form only a small part of the whole capacity or inductance, as the case may be. Small semi-variable elements may be made much more constant than large ones, and moreover their variations are relatively less important. This method limits the wave length zone over which the pre-tuning of one push button may be varied. If the wave length zones of the available push buttons are distributed over the whole wave length range of the broadcast, the user must then choose stations for the push buttons which are also distributed over this whole wave length range [2]).

The systems with mechanical pre-tuning have the inherent advantage of requiring no change in the electrical part of the receiving set. The push buttons act through a mechanical transmission system directly on the tuning condenser. Each push button moves the condenser to a definite, previously determined position. If one considers the wide angle which the plates of a rotating condenser can describe, it is obvious that a relatively large force must be applied or a great distance must be covered. Against this disadvantage, in addition to the advantage already mentioned, there is the fact that the accuracy of reproduction of the adjustment is determined wholly by the mechanical construction, and is therefore less sensitive to external influence [3]). The problem of the previous adjustment of the push buttons on the desired stations can be solved by means of relatively simple constructions. In connection with the reproducibility the following requirements must be fulfilled. When the button is not pressed in sufficiently far (because of carelessness of the user) it must nevertheless reach its final position or spring back again to its initial position. Furthermore the wear down to which this arrangement, like every other mechanical arrangement, is exposed may affect only the adjustment (which can easily be regulated again) and may not affect the accuracy of adjustment.

The motor-driven systems have the same advantages as the other systems with mechanical pre-adjustment, and in addition the advantage of the systems with electrical pre-adjustment, namely that the main function of the push button is that of a switch for starting the motor, and that a light pressure is therefore sufficient. The fact that the tuning manipulation itself is left to a motor makes it possible to increase the distance covered in that manipulation and this improves the accuracy of the adjustment, as we shall see later.

In the earliest sets with push-button tuning, where the accuracy of adjustment was still unsatisfactory, use was usually made (and sometimes still is) of automatic frequency correction. When incorrect tuning leads to the fact that the carrier wave, passed on the intermediate frequency part, does not have the same frequency as the permanently tuned intermediate frequency circuits, this frequency correcting circuit causes the oscillator frequency to be changed automatically so that the intermediate-frequency signal finally does assume the correct frequency. At the same time, however, the high-frequency circuits are far from correctly tuned, so that the chance of distortion and cross modulation becomes greater. Automatic frequency correction moreover necessitates the use of extra valves. In the push-button tuning system to be described sufficient accuracy is attained to make the application of automatic frequency correction unnecessary.

A motor-driven push-button tuning system

The system developed by Philips which we are

[2]) This presents no difficulties in the United States because the distribution of wave lengths is such that stations which are neighbours geographically are suitably far away from each other in the ether, at least 50 kilocycles/sec. The stations which will naturally be chosen by a listener are therefore automatically fairly uniformly distributed over the whole wave length range. The situation in Europe in this respect is quite different.

[3]) Of course there always remains the slight spontaneous detuning also experienced in ordinary receivers without push-button tuning, which is due to the heating up of the apparatus in use. This can, however, be eliminated by carrying out the pre-tuning of the push-buttons only after the set has become warm.

about to describe makes use of a motor. The principle is the following: a push-button starts the motor and at the same time brings into position a stop which determines a definite final position of the rotating condenser. The motor turns the shaft of the rotating condenser and is switched off automatically when the stop is reached, *i.e.* when the rotating condenser has reached the proper position.

Fig. 1 shows the whole tuning mechanism diagrammatically. The motor S *via* two friction wheels A and B and a toothwheel transmission C, D turns a shaft E, on which are fixed a number of tuning cylinders F_1, F_2, ... The motion of the shaft E is transmitted by the toothwheels G, H, to the rotating condenser in a transmission ratio of 1 : 5. When the condenser moves from minimum to maximum capacity, *i.e.* a half turn, the tuning cylinders make $2^1/_2$ turns. *Fig. 2* is a photograph of a single tuning cylinder. A helical groove of five turns is cut on the surface of the cylinder. Along this groove moves a peg which is fastened to one end of a fairly long strip, T, the other end of which is fastened in such a way that it can rotate about a horizontal and a vertical axis (see photograph *fig. 3*). In this way the peg easily moves up and down as well as from side to side to follow the spiral groove. Exactly in the middle of the cylinder, *i.e.* $2^1/_2$ turns from either end, there is a hole bored in the cylinder into which the peg fits. When the peg is pressed against the bottom of the groove by means of the push button, and when the cylinder turns, the peg will slip into the hole at a given moment and prevent further motion of the tuning cylinder and also of the rotating condenser. The strip to which the peg is fastened due to its rather elastic flexibility takes up the kinetic energy of the whole mechanism (*i.e.* of the rotating condenser, the tuning cylinders and the toothwheels).

Actually the peg is not pressed down by the push button itself, but by a spring which is freed for action by the pressing of the button. In this way it was made possible for the push button to be pushed in immediately to its full depth, and the user need not keep his finger on the button until the peg has slipped into its hole. A simple locking arrangement (see Q in *Fig. 4*) holds the button in the final position and makes it spring back as soon as another button is pressed.

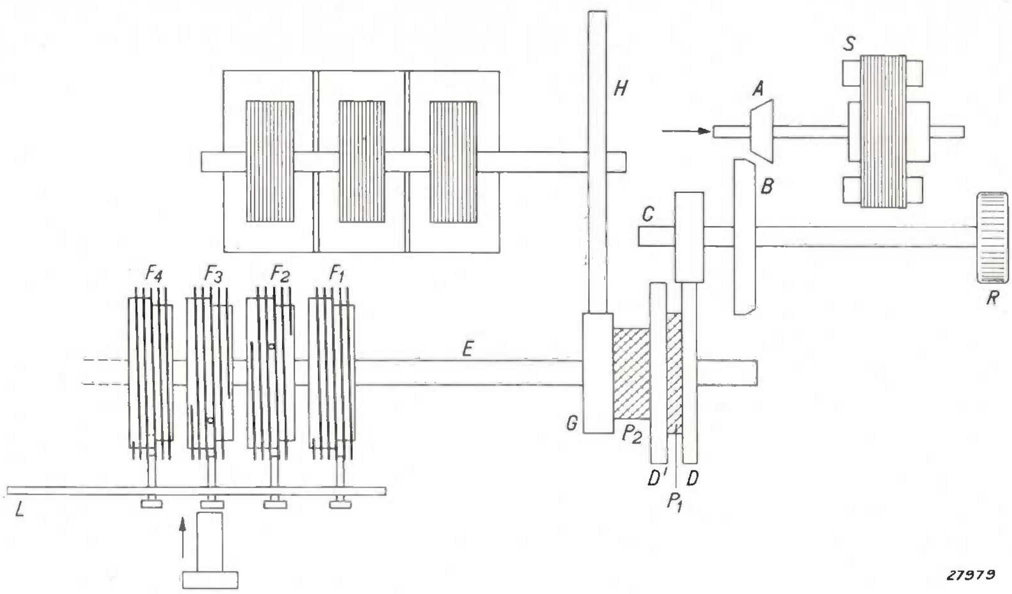


Fig. 1. Diagram of the mechanism of motor-driven push-button tuning. S = motor, A-B = friction drive, C-D and G-H = toothwheel transmissions, F_1, F_2 ... = tuning cylinders P_1 and P_2 = friction couplings; R = knob for hand tuning, L = contact bridge.

When the rotating condenser is in any arbitrary position, it will have to rotate to the right to reach some stations and to the left for other stations. The groove in the tuning cylinder is cut to two different depths: the bottom of the left-hand $2^1/_2$

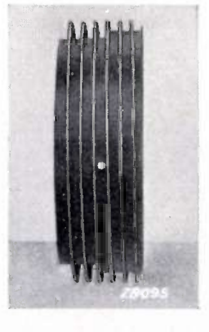


Fig. 2. Photographs of a tuning cylinder. The left-hand windings of the groove are somewhat deeper than the right-hand ones.

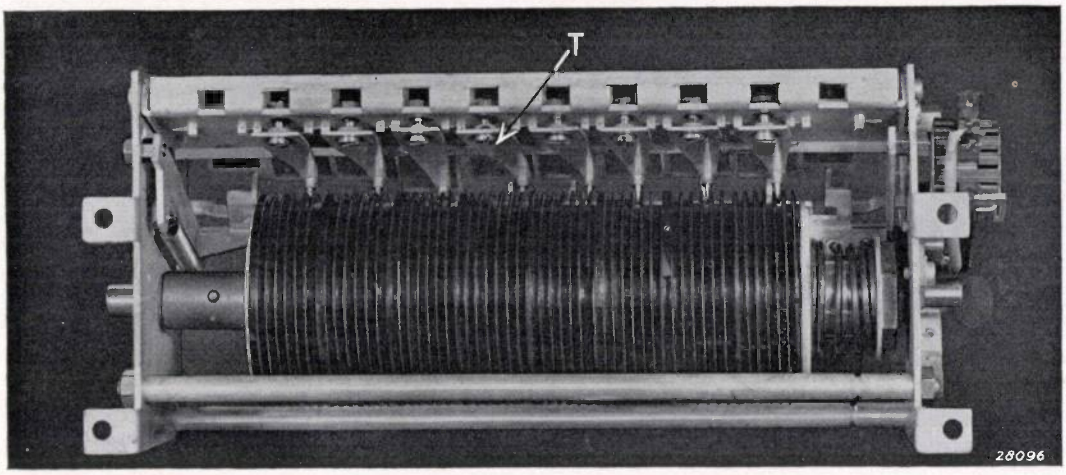


Fig. 3. Photograph of the keyboard, taken from the rear. The long strips carrying the pegs may be seen. To the right are the contacts which serve for starting and reversing the motor. The third from the left is pushed down.

windings is deeper than that of the right-hand windings. When the peg is down it takes with it the bridge L which extends over all the tuning cylinders and moves several contacts at the side of the keyboard (see photograph, fig. 3, right-hand side). If the peg is to the left of the hole, it, and with it the contact bridge, will be somewhat lower when the peg is pushed down than when the peg is to the right of the hole. In the two different positions of the contact bridge two different contacts are made and the motor turns in one direction or the other according to which contact is closed. In this way provision is made that the rotating condenser is always turned automatically in the direction in which the peg must travel to reach the hole, and the condenser is thus rotated over the smallest possible angle to the position corresponding with the station chosen.

We shall now examine the commutation arrangement of the motor. In *fig. 5* the group of contacts I to IV is drawn separately. The contact piece k which is fastened to the above-mentioned bridge L is directly connected with the contact spring II. The bridge may now assume four different po-

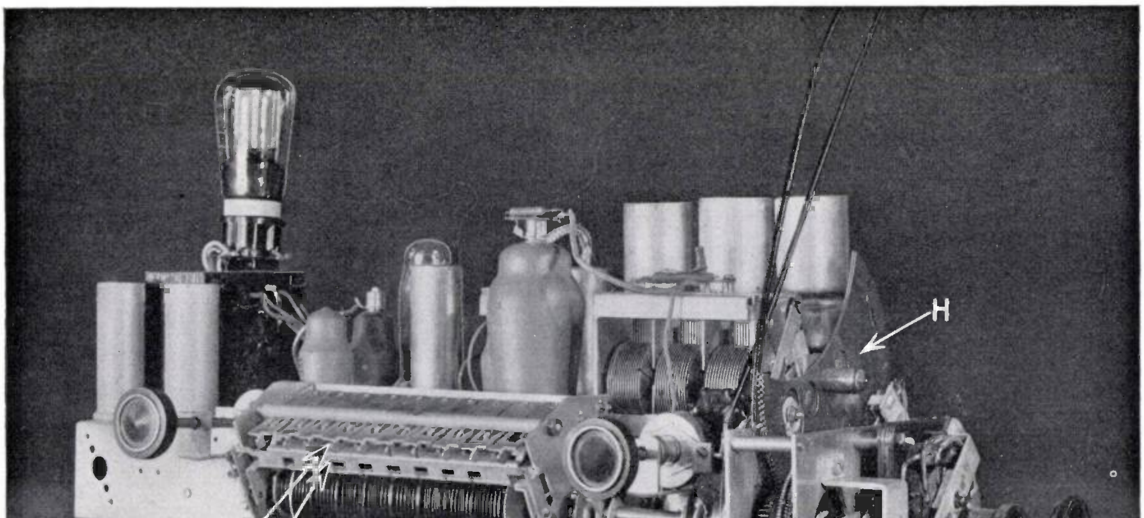
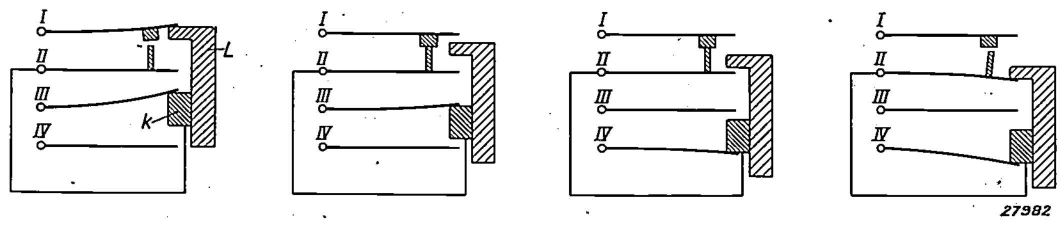


Fig. 4. Photograph of the push-button tuning arrangements mounted in the receiver. In the middle is the keyboard, to the right the motor and transmission. The locking arrangement Q holds down the key pressed: the end of the key n falls into the corresponding opening of the beam Q. When another key is pressed the beam Q is first pushed away against the action of a spring (visible in fig. 3 to the left), and the key previously held is thereby released.

sitions. If none of the buttons is pressed down the bridge is drawn up by a weak spring in position *a* (fig. 5*a*), in which the springs *I* and *II* make no contact with each other and the motor is therefore not in action (see fig. 6*b*). If a button is now pressed down, as we have already explained, a peg and with it the contact bridge are pressed down by a stronger spring. If the peg of the tuning cylinder in question rests on the shallower part of the groove the contact bridge is in position *b* (fig. 5*b*). The contact *I* - *II* closes and the motor starts. In this position there is contact between the terminals *II* and *III*, which corresponds to a certain direction of rotation of the motor. If the peg rests on the deeper part of the groove the bridge assumes position *c* (fig. 5*c*) where *I* and *II* still make contact, but the magnetic field and thus displaced axially against the spring. The friction drive *A*, *B* (fig. 1) thereby begins to work. At the moment when the motor is switched off, the axle of the drum armature springs back and the motor can come to rest freely. The relatively high kinetic energy of the motor is in this way unable to damage the precise mechanism of the tuning.

The axis of the toothwheel *C* in fig. 1 can be turned by hand with the knob *R*. In this way ordinary continuous tuning may be carried out. Considering the high transmission ratio of the toothwheels *C*, *D* and *G*, *H*, the user would have to turn for some time the rotating condenser from one final position to another. In order to facilitate this process two additional keys are added



a b c d

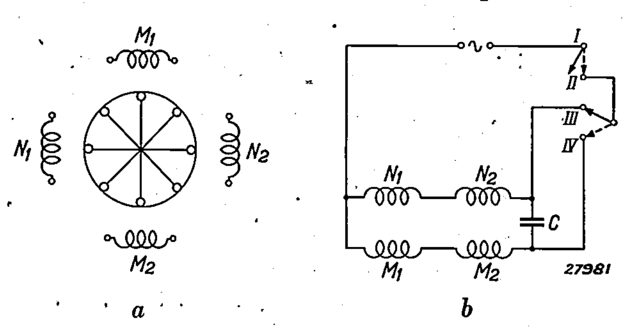
Fig. 5*a* - *d*. Position of the contacts *I* - *IV* in the four different positions of the contact bridge *L*.

where *II* is connected with *IV*, and the motor turns in the other direction. As soon as the peg slips into the hole in the tuning cylinder the contact bridge falls somewhat lower still, into position *d* (fig. 5*d*). The contact *I* - *II* is here open again, the motor is therefore switched off and the receiver is tuned in to the desired station.

Fig. 6a gives a diagram of the motor. A drum armature is situated between four poles. The coils M_1 and M_2, as well as N_1 and N_2 are connected in series. Fig. 6*b* gives the diagram of the circuit. If the contacts *I* - *II* - *III* are connected (position *b* of the contact bridge *L*, fig. 5*b*), the presence of the condenser *C* causes the phase of the current in the pair of coils N_1 - N_2 to lag behind that of the current in the coils M_1 - M_2. Due to the phase shift of the two perpendicular fields, the resultant field is rotating and a couple is produced on the drum armature. If, however, the contacts *I* - *II* - *IV* are connected (position *c* of the contact bridge *L*, fig. 5*c*), the phase of the current in the pair of coils N_1 - N_2 is ahead of that in M_1 - M_2, the field (and therefore also the drum armature) now turns in the opposite direction.

Upon switching on the motor the drum armature, which, due to the action of a spring, is held somewhat outside of the axis of the coils, is drawn into to the keyboard to the right and left of the station keys. These additional keys serve to set the motor in motion in either direction, and may be used by the listener to bring the dial pointer into the neighbourhood of the desired station. The accurate tuning may be then carried out by hand.

If it is desired to change over from push-button to hand tuning, the last button pushed in must be made to spring back to its initial position. However, the mechanism is made fool-proof in this



a b

Fig. 6*a*. Diagram of the motor. A drum armature rotates in the field of four coils M_1, M_2, N_1 and N_2.

Fig. 6*b*. Circuit of the motor. *C* is a condenser which provides for the necessary phase shift in the currents in the pairs of coils M_1-M_2 and N_1-N_2.

[4]) Incidentally it may be noted that with this displacement of the shaft of the armature a contact is also closed which connects a condenser in parallel with the loud speaker. During the process of tuning, therefore, almost no sound is heard.

respect: between the toothwheels D and D' (fig. 1) the shaft E is broken and a rather weak friction coupling P_1 joins the parts. If the user turns the tuning knob R when the whole mechanism is locked by the peg in the tuning cylinder, the coupling P_1 slips and the toothwheel D_1 does not move.

The tuning knob R also serves for pre-adjustment of the push buttons on the desired stations. For this purpose the knob must be pushed in slightly. The toothwheel C then catches in D_1 directly and the friction coupling P_1 is out of action. The tuning cylinders F_1, F_2 ... are not fastened immovably to the shaft E, but are movable by a frictional force stronger than that in the coupling P_1. If a tuning cylinder, F_3 for example, is held by the corresponding peg, and if the slightly depressed tuning knob is then turned, all the other tuning cylinders turn with the shaft E. If one tunes in to a station in this way, the push button of cylinder F_3 in then set on that station. The pre-adjustment of the push buttons thus takes place in the same way and even with the same knob as ordinary hand tuning. The dial which is turned by the shaft E also turns at the same time.

When the rotating condenser is in one of its two end positions and the user tries to turn the tuning knob still further, the friction coupling P_2 between the toothwheels D' and G begins to slip, so that no harm results. For satisfactory functioning of the mechanism the friction in the coupling P_2 must of course be greater than that which causes the cylinders F to turn with the shaft E.

Construction of the motor

The motor is so constructed that it works not only on ordinary alternating current of 50 cycles but also on alternating current supplied by a vibrator [5]). The latter current has a higher frequency, namely 100 - 120 c/s. When one keeps in mind that the inductive resistance of the windings (see fig. 6*b*) and the capacitive resistance of the condenser takes on different values at the higher frequency, it is obvious that the motor has to fulfil different requirements in the latter case. The result is that the starting couple of the drum armature becomes smaller and the number of revolutions with no load becomes higher. By a suitable construction of the drum armature the couple has been made great enough even at the higher frequency. A regulator has been introduced on the axle of the armature which provides that even at the higher alternating current frequency the number of revolutions of the motor is not too high. At the vibrator frequency the axial force on the drum armature becomes smaller and due to this the friction drive A B (fig. 1) may begin to slip. The regulator on the axle of the armature is so constructed that the segments, forced away from each other by centrifugal force, have a tilted motion and in this way exert not only a braking effect but also a force in the axial direction, so that the total pressure is sufficient for the friction drive A, B.

Accuracy of tuning

In spite of the fairly complicated working of the mechanism an excellent degree of reproducibility has been attained. This is due in the first place to the fairly high transmission ratio (1 : 5) between the axle of the tuning cylinders and that of the rotating condenser. In any case the reproducibility will depend upon the amount of play of the peg in the hole in the tuning cylinder. This play will be of less importance the longer the path covered by the peg on its way to the hole. When the condenser is turned from minimum to maximum capacity, *i.e.* half a turn, the tuning cylinder makes $2^1/_2$ turns; with a cylinder diameter of 44 mm therefore the relative path of peg and hole is about 350 mm. If we assume that the frequency of the apparatus varies approximately in proportion to the angle of rotation of the condenser, and therefore also in proportion to the distance between peg and hole, and if the frequency on medium waves increases for instance from 500 to 1500 kc/s upon a half turn of the condenser, then a play of 0.1 mm means a variation in frequency of about $0.1/350\ (1500 - 500) \approx {}^1/_3$ kc/s. The maximum deviation in reproduction of a given push-button tuning is actually found to amount to not more than 0.5 kc/s. Of course it was necessary to suppress any backlash in the toothwheel transmission G, H (fig. 1), between tuning cylinders and rotating condenser. This was done by means of the following method of construction. The toothwheel H (see fig. 4) consists of two exactly similar toothwheels which are placed next to each other on the shaft. One of them is fastened to the shaft, the other is carried along by a spring acting in a tangential direction and fastened to both toothwheels. The driving toothwheel G engages with both toothwheels of H at the same time; by means of the spring one tooth of G is always clamped firmly between a tooth of one and a tooth of the other two toothwheels H, and there can be no play.

[5]) *Cf.* Philips techn. Rev. **2**, 346, 1937, for details of a vibrator for the supply of A.C. sets from D.C. mains.

X-RAY TUBE FOR THE ANALYSIS OF CRYSTAL STRUCTURE

by J. E. DE GRAAF and W. J. OOSTERKAMP. 621.386.1: 548.73

The following special requirements must be made of an X-ray tube which is intended for the analysis of crystal structure: the exposure time required must be short, the radiation must be spectrally pure and the operation of the tube must be easy. A tube with water-cooled anode which satisfies these requirements is described. Special attention is paid to the shape of focus necessary to obtain the greatest possible brightness. The problem of heat conduction connected with the cooling of the anode is dealt with.

There exist in the main two large fields for the application of X-rays: medicine and the testing of materials. In medicine X-rays are used for diagnostic and therapeutic purposes, in the testing of materials for detecting macroscopic defects [1]) and for studying crystal structure by means of diffraction diagrams [2]).

The requirements made of the X-ray tube are different for each of these two applications. To name a single example: in certain therapeutic applications of X-rays, where use is made of their destructive effect upon tissues, it is of foremost importance that the rays possess high penetration and energy. The shape of the focus (*i.e.* the spot where the X-rays originate) is of less importance; it may be fairly large. In the other applications mentioned where it is a question of obtaining an image of internal details or a diffraction diagram, it is very important that the focus be small for the sake of sharpness of the image or diffraction diagram. On the other hand there are also conditions which must be kept in mind in the construction of all kinds of X-ray tubes, among others are the protection of the operator against X-rays other than those of the beam directed on the object, and the safeguarding against the high voltages necessary for the working of the tubes.

In this article we shall deal with a tube which has been developed especially for use in the analysis of crystal structure, not only in scientific and technical laboratories, but also in the workshop. For a detailed description of the procedure in the different branches of structure analysis (with a single crystal according to Laue, with a powder according to Debye and Scherrer, investigation of stresses, etc.) we refer the reader to the series of articles by Burgers [2]). The various possible arrangements all have the common feature that the beam of X-rays coming from the tube is first reduced by means of a diaphragm to a narrow ray; this ray then falls upon the substance to be examined. It is then deflected or reflected in different directions, which are determined by the crystal structure, and a blackening occurs at corresponding points on a photographic film situated at some distance from the substance (*fig. 1*).

In the construction of the X-ray tube which we are about to describe the following points have been especially considered:

1) The required exposure times must be as short as possible. This is important not only for routine work, but also for scientific research, where reasonable exposure times are desirable even in the case of poorly reflecting substances or when a small opening of the diaphragm is used.
2) The radiation must possess a pure spectrum. When this is not the case the undesired spectrum lines themselves give rise to interference images, and in this way make the diagrams much more complicated and their interpretation unnecessarily difficult.
3) The tube must be easily handled and adjusted, especially for industrial purposes.

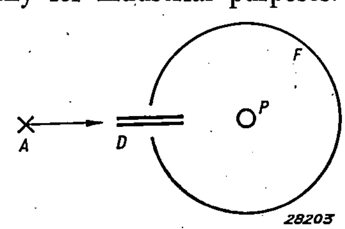


Fig. 1. Diagram of an arrangement for crystal structure analysis. A = focus from which the X-rays are emitted, D = diaphragm, P = substance, F = film.

Construction of the tube

Fig. 2 is a diagram of the construction of the tube. The hot cathode K emits electrons which are accelerated by a potential difference of 20 to 40 kilovolts in the direction of the anode A. At the anode a small part of the kinetic energy of the electrons is transformed into X-radiation, the greater part, however, into heat. The anode is soldered into a chrome-iron can D, to which the glass part G which bears the cathode is fused. This construction makes it possible to cool the anode

1) See the five articles by J. E. de Graaf on this subject in the 2nd and 3rd volumes of this periodical.

2) See the twelve articles by W. G. Burgers on this subject in the 1st and 2nd volumes of this periodical.

with water, so that an efficient dissipation of heat is attained. The cooling water flows in and out at W. At L there are glass windows in the anode can through which the X-rays are emitted to the outside. A metal cylinder B kept at cathode potential surrounds cathode and anode. The wall of the cylinder has only several small openings for the passage of the X-rays. In this way provision has been made against secondary electrons from the anode being able to bombard the window L. A lead jacket P surrounds the anode can and captures all X-rays except the beam passing through the windows.

The anode is earthed, so that it is not necessary to have an insulated pump with a closed water circuit for cooling, but water from the mains may be used. Since the cathode is at high potential (which is supplied *via* the high tension cable H) the heating current transformer must be insulated. This transformer is built in with the high tension transformer. The whole tube is placed in a metal container M which, like the metal covering of the high tension cable, is earthed, so that it is impossible for the user to come into contact with current bearing components. The cathode is mounted in the opening of a plate C which is kept at cathode potential. This plate with its opening acts as an electron lens, *i.e.* it causes such a deflection in the paths of the electrons that they are focussed sharply on the anode. By giving the cathode a suitable form, a focus of given shape and dimensions can be obtained.

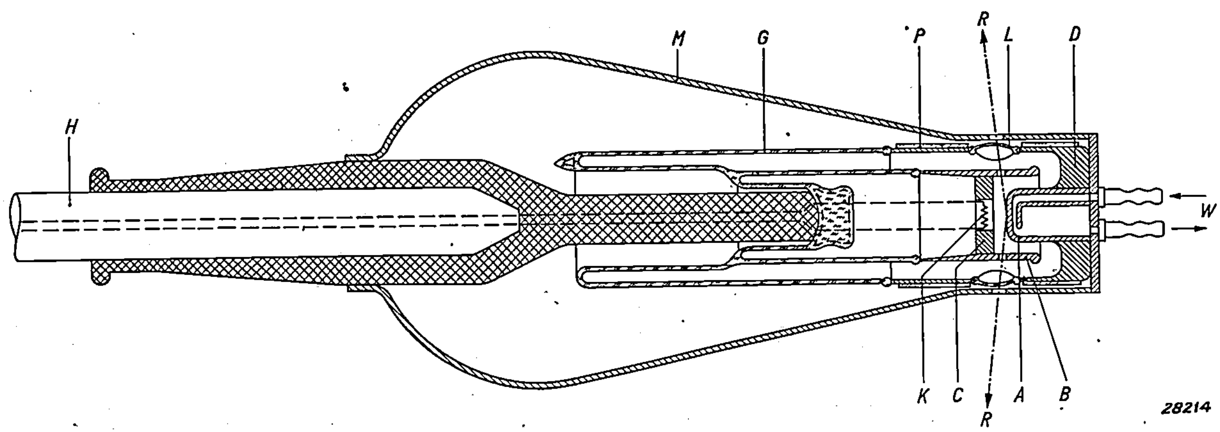


Fig. 2. Construction of the X-ray tube for structure analysis. The anode A is part of the chrome-iron can D, to which is fused the glass part G, which bears the shielding cylinder B with the plate C and the cathode K. The X-rays pass through the windows L in the direction of the arrows R. At W the cooling water flows in and out. The tube is placed in a metal container M which, like the anode and the metal covering of the high tension cable H, is earthed. A lead jacket P surrounds the anode can.

Short exposure times

We shall now discuss the way in which the requirements mentioned at the beginning of this article are fulfilled. There are various important factors in obtaining a short exposure time, for example, the filter of the tube itself. In order to pass from the vacuum in the tube into the outer atmosphere, the X-rays must inevitably pass through a wall by which they are weakened. The weakening is kept as small as possible in the tube described by making the windows L (*fig. 2*) of Lindemann glass, which contains only elements with a low atomic number and therefore absorbs only little of the radiation. Moreover, the windows are very thin, namely 0.12 mm in thickness. This type of window was made possible by the above-mentioned protection of the windows from bombardment by secondary electrons from the anode. The active radiation obtained with a copper anode is weakened only 25 per cent by the windows.

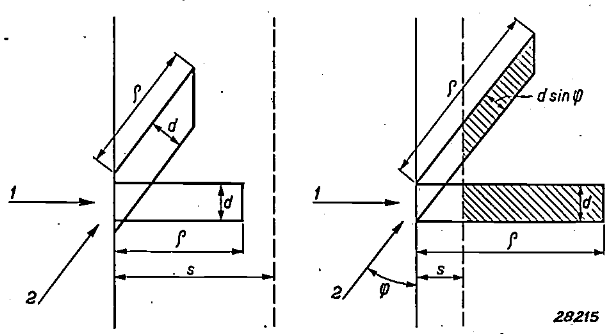
Furthermore it is clear that the exposure times may be chosen shorter the closer the camera can be placed to the focus: the intensity of the radiation is inversely proportional to the square of the distance. The minimum distance is determined by the diameter of the tube. This latter is closely connected with the maximum voltage to be used, since the distance between the shielding cylinder B and the anode necessary to avoid breakdown increases proportionally with the voltage to be used. In order to be able to use the tube for Laue diagrams also, where relatively high voltages are necessary, the maximum permissible voltage was fixed at 60 kilovolts (for alternating as well as for direct voltage). Nevertheless, because of the compact construction of the components in the anode can, the diameter is only 60 mm, so that the camera may be brought up to a distance of about 35 mm from the focus.

The most important factors for obtaining short exposure times are, however, the shape and the brightness of the focus. The first method that would occur to one, to increase the intensity of the X-ray beam, would be to increase the dimensions of the focus and choose correspondingly greater diaphragm openings. This method, however, as was mentioned in the introduction, is limited by the lack of sharpness which would result from too large a source of radiation. In general, for making crystal structure diagrams a diaphragm opening of 1 mm diameter [3]) at the most is used, and it is therefore useless to make the source of radiation much larger than 1 × 1 mm. When the surface to be irradiated is determined, the X-ray energy can only be increased by increasing the brightness of the focus.

Shape and brightness of the focus

The fact that X-radiation does not follow Lambert's law may be utilized in determining the shape of the focus. According to that law the brightness of a radiating surface is the same from all directions, and it is valid in the case of so-called body radiation. If one looks normally at the surface, for instance of a volume of incandescent gas of the thickness s (arrow *1*, *fig. 3a*), the outermost layer as well as the deeper layers contribute to the light observed. The light of the more deeply lying parts will, however, be weakened by absorption, and it will be more weakened the more deeply its source lies in the gas. After a definite depth ϱ practically no more contribution will be made to the light leaving the surface. If one looks obliquely at the surface of the incandescent mass (arrow *2*), a column of the length φ will again contribute to the observation of light, and with the same diameter of column as in the first case, *i.e.* from the same apparent surface, the same light intensity is obtained. This means that the brightness (light intensity/apparent surface) is the same in both directions. From fig. 3*a*, however, it may be seen immediately that the above deductions are only valid for body radiation, *i.e.* when the depth s of the layer in which the light emitting processes take place is greater than the depth ϱ of the layer which contributes to the light leaving the surface. Fig. 3*b* shows the opposite case, where $s < \varrho$, *i.e.* a surface radiation. The cross hatched areas of the column (whose length is determined only by the absorption in the substance of the radiation emitted) now contribute nothing, since no light-emitting processes take place in those areas. When the column has an oblique position that part of it which can cause radiation to be emitted from the surface becomes longer. If one now considers the emission from the same real surface, the same light intensity is obtained in both directions, because the volume of the column which contributes has remained the same (oblique cylinders with equal bases and equal altitudes). Considering, however, that in the oblique direction the apparent radiating surface is smaller, the brightness in the oblique direction will be greater than in the normal direction of observation. *Fig. 4* shows the dependence of brightness on the angle of observation for both cases, body radiation (*I*) and surface radiation (*II*).

In the case of X-rays we are actually concerned with a surface radiation. The absorption of X-rays in the material of the anode is much less than that of the bombarding electrons. The electrons excite X-radiation in only a thin layer (thickness $s = 0.01$ mm, for instance), while the radiation would be



a *b*

Fig. 3*a*) Body radiation. The depth s of the emitting layer is greater than the depth λ of the layer which contributes to the light emitted. Upon normal (*1*) and oblique (*2*) direction of observation the same light intensity is obtained from the same apparent surface (*d*). The brightness is therefore the same in all directions (Lambert's law).
b) Surface radiation. In this case $s < \lambda$. With oblique direction of observation (*2*) the same light intensity is obtained from the apparent surface $d \sin \varphi$, as is obtained with normal observation (*1*) of the surface d. The brightness in the oblique direction is greater ($1/\sin \varphi$ times) than that in the normal direction.

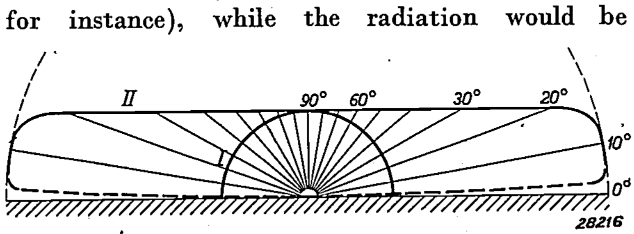


Fig. 4. Polar diagram of the brightness with volume radiation (*I*) and surface radiation (*II*). At small angles the straight line *II* bends toward a circle (Lambert's law, as in *I*). At still smaller angles the irregularities in the surface of the anode become noticeable and the brightness diminishes toward zero (indicated by dotted line).

[3]) For special cases smaller diaphragm openings are also used, 0.5 or 0.25 mm.

able to leave the surface from much greater depths (ϱ) (fig. 3*b*). If the focus is made oblong in shape, for instance a rectangle with the width b and the length $l = b/\sin\varphi$ (line focus *fig. 5*), a radiating square of side b is observed in the direction φ, and its brightness is a factor $1/\sin\varphi$ greater than upon normal observation. This principle was first applied in X-ray tubes by Goetze.

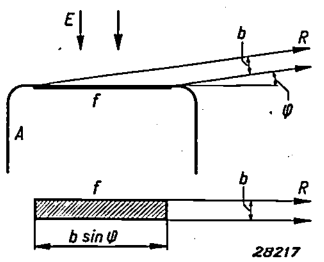


Fig. 5. Shape of the focus on the anode A. The electrons fall perpendicularly upon the anode (E), the X-rays are observed at the angle φ (R). The width of the focus is b, the length $b/\sin\varphi$, so that the focus is observed as a square in the direction φ.

From fig. 3*b* it may be seen that with a very small angle between the direction of observation and the surface the above considerations no longer hold, because the whole column with the length ϱ then falls in the emitting layer of thickness s. In the latter case Lambert's law again holds, and the straight line II (constant light intensity) in fig. 4 bends around toward a circle [4]). After a definite angle therefore, further reduction in that angle produces no further increase in brightness. Moreover, at small angles to the surface slight irregularities in the anode (due to the roughening of the surface, see below) have a disturbing effect. For these reasons in the tube described the angle between the direction of observation and the surface of the anode is chosen at 6°. In order to be able to carry out the focussing easily even at the largest diaphragm openings used (1 mm), the apparent dimensions of 1.2 × 1.2 mm were chosen for the focus; the actual dimensions are therefore 1.2 × 12 mm ($\sin 6° \approx 0.1$). The line focus has two identical directions of observation. For certain applications a direction perpendicular to these may also be used, so that four windows in all have been introduced [5]).

[4]) Actually the deflection only appears at much smaller angles than that drawn in fig. 4. With a given monochromatic X-radiation F. Wisshak (Ann. Physik 5, 507, 1930) found that at $\varphi = 20°$ the brightness was 2.8 times, at 10°, 5.2 times, and at $\varphi = 6°$ 8.1 times as great as with normal observation ($\varphi = 90°$).

[5]) In principle it is possible to apply several such foci to the anode in different directions (by means of several cathodes) and thereby to make more than two normal photographs at the same time.

Specific loading of the focus

With given shape and dimensions of the focus the brightness may still be increased by increasing the specific load (the energy given off by the electrons per square centimeter and per second). The brightness is proportional to the specific load at constant voltage. However, the heat developed also increases proportionally, so that the temperature in the neighbourhood of the focus becomes higher. The permissible temperature of the anode is limited by the occurrence of evaporation of the focus material which causes the focus to become more or less rough. The requirement is made that the erosion of the focus after 1000 or 2000 hours use must not exceed a certain limit which is connected with the above-mentioned decrease in brightness at small angles. This requirement therefore fixes a maximum permissible rate of evaporation, *i.e.* a maximum permissible temperature for every focus material.

As much heat as possible must be developed on the focus (high specific load), but the maximum permissible temperature must not be exceeded. It is therefore a question of conducting the heat away as well as possible. The heat dissipation takes place chiefly by conduction in the anode material toward the rear wall (*fig. 6*) which is cooled with running water, and which has a practically even temperature over the whole surface cooled. This temperature must remain everywhere lower than 100 °C, since otherwise the cooling water would begin to boil with the result that an insulating layer of steam would be formed and finally boiler-scale. This limitation of the temperature of the cooled surface determines the heat flux to be handled per unit of the cooled surface at a given rate of flow of the cooling water. The specific load W_f of the focus which can be reached under the two conditions mentioned: maximum permissible focus

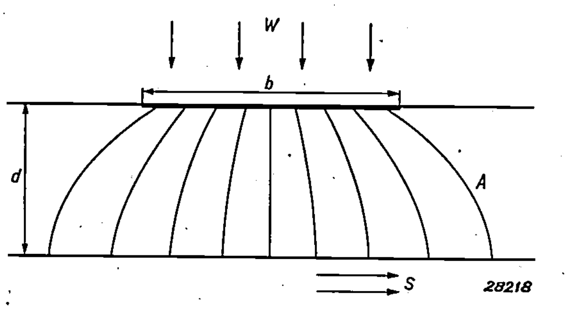


Fig. 6. The problem of heat conduction at the anode. At the focus f (whose longest side must be imagined perpendicular to the plane of the drawing) the heat flux W is developed. The heat is conducted through the body of the anode A to the rear wall where cooling takes place by means of running water s. As is indicated a divergence in the lines of heat flux occurs which is more pronounced the greater the thickness of the anode d.

temperature and maximum permissible density of heat flux on the cooled surface, depends upon the thickness d of the anode. At a certain thickness of the anode the value which may be assumed by W_f reaches a maximum. This may be explained in the following way.

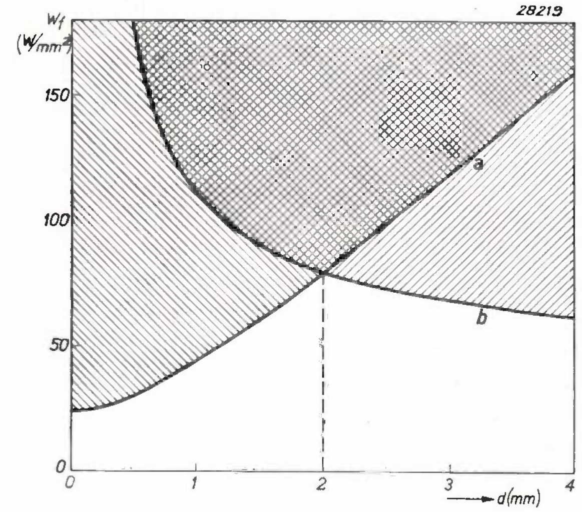


Fig. 7. Dependence of the specific focus loading W_f on the thickness of the anode d, with an infinitely long line focus 1.2 mm wide on a copper anode. In curve a the density W_s of the heat flux at the cooled surface is at its maximum permissible value (24 W/sq.mm). In curve b the focus temperature T_f is at its maximum permissible value (300 °C), while the temperature of the water-cooled surface is fixed at 100 °C (in the case of curve a this condition was automatically fulfilled). The "permitted" region is below both curves. The thickness d which corresponds to the point of intersection of the curves is the optimum thickness. The curves are calculated for a constant loading (direct current); the maximum value of W_f attainable, when $d = 2$ mm, is then 80 W/sq.mm. When alternating current of the same effective value is used the attainable value of W_f becomes smaller, namely 60 W/sq.mm.

In *fig.* 7 W_f is plotted against d. Curve a is calculated for the case where the density of heat flux W_s is always at its maximum permissible value, which is here assumed to be 24 W/sq.mm. The temperature of the focus is disregarded in curve a. With a very thin anode (small d) the lines of heat flux run practically parallel to each other toward the rear wall, *i.e.* the density of heat flux is uniform, and W_f may therefore be no greater than W_s (24 W/sq.mm). If the anode is thicker (larger d) a certain divergence of the heat flux appears in the heat conduction (see the lines of flux in fig. 6), so that the density of the heat flux is smaller at the rear wall than at the focus. The permissible value of W_f therefore becomes greater than W_s; in the first approximation it increases proportionally with d. The region above the curve is "forbidden", because every point in that region represents a combination of W_f and d, where W_s would be greater than 24 W/sq.mm. Curve b is calculated for the case where the focus temperature T_f is always at its maximum permissible value, which is here taken as 300 °C (copper anode). In this curve the density of heat flux at the cooled surface is disregarded but it is assumed that the rate of flow of the cooling water is always high enough, so that the cooled surface maintains a temperature of 100 °C. Along the curve therefore the temperature difference between focus and cooled surface is constant and equal to 200 °C. At very small values of d the temperature gradient, and therefore also the heat flux in the anode which is proportional to it, must be very great in order that the temperature difference between the two surfaces may reach the figure mentioned. In this case therefore W_f is very great. At greater values of d the temperature difference mentioned is attained with a smaller gradient and therefore smaller value of W_f. Here again the region above the curve is "forbidden", since at every point in that region the focus temperature would be higher than 300 °C.

If it is now desired to satisfy both conditions, that for T_f and that for W_s, then only that region is "permitted" which lies not only below curve a, but also below curve b. It is clear that with a given thickness of anode d, namely that which corresponds to the intersection of the two curves, the greatest specific loading of the focus W_f is attained.

This optimum thickness of wall depends upon the rate of flow of the cooling water, since this rate of flow determines the permissible density of heat flux at the rear wall of the anode. Moreover it is, of course, not the same for different anode materials with their different physical constants. Since, however, these materials are only applied in the form of a thin plate to a base of copper, the mutual differences between the corresponding optimum thicknesses of wall are not great. The thinner the plates applied, the better the heat dissipation and the higher the specific load. In the tube here described the rate of flow of the water is 5 m/sec, for which a pressure of 1 atmosphere and a consumption of water of 6 litres/min. is necessary. The optimum wall thickness is then about 2 mm, and the permissible specific load with a copper anode is 60 W/sq.mm, when the tube works on alternating voltage and the line focus is 1.2 mm in width. For the sake of comparison it may be added that with the aircooled "Metalix" tube for structural analysis [6]) the specific load is 25 W/sq.mm and with another watercooled tube [7]) 30 W/sq.mm (both tubes having copper anodes).

[6]) A. Bouwers and W. Busse, Z. Krist. 77, 507, 1931.

[7]) R. Glocker, Materialprüfung mit Röntgenstrahlen, 1936, p. 167.

For iron, cobalt and chromium in the tube described a specific load of about 35 W/sq.mm is permissible, for tungsten and molybdenum about 70 W/sq.mm, when these metals are soldered to the copper base of the anode in the form of plates 0.5 mm thick.

Homogeneity of the focus loading

Fig. 8 shows the distribution of temperature over the width of the line focus. In the calculation of this curve (as in the curves of fig. 7) the line focus

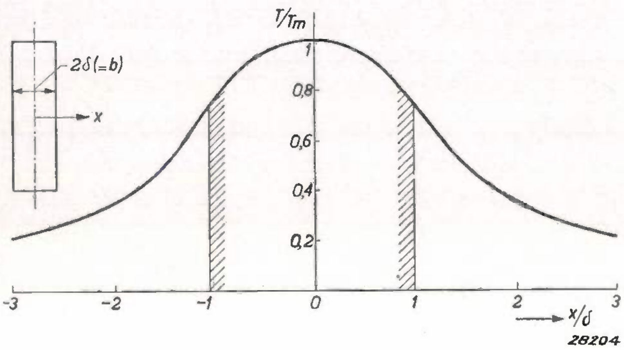


Fig. 8. Temperature distribution over the focus in the case of homogeneous loading. At the edges of the focus ($x = \delta$) the temperature is about 20 per cent lower than in the middle.

is assumed to be infinite in length, which gives a satisfactory approximation. It may be seen from the curve that the temperature at the edges of the focus is about 20 per cent lower than in the middle. This phenomenon, which may be ascribed to the lateral dissipation of heat from the focus to the adjacent anode material, occurs with the homogeneous loading of the whole surface of the focus which has been assumed until now. When, however, the cathode has the form of a spiral the loading of the focus is not homogeneous. Equal numbers of electrons are emitted per sq.cm of the whole surface of the spiral. They are drawn immediately toward the anode by the strong field. It is reasonable to suppose that the paths of the electrons which are emitted from parts of the spiral whose tangents are intercepted by the focus will lie closest to each other. Therefore the current density is greatest at the outside of the electron beam, and consequently the focus is more heavily loaded at the edges than in the middle. Therefore with a cathode in the form of a spiral the temperature distribution over the width of the focus will not have the form shown in fig. 8, but will be more nearly constant over the whole width (it may increase slightly at the edges). Considering the fact that the temperature permissible for the focus material is nowhere exceeded, and that nevertheless a higher average specific loading is obtained, one might think that the lack of homogeneity of the focus loading would even be advantageous. In the case of certain kinds of X-ray tubes for medical diagnosis use is actually made of this feature. For crystal structure analysis, however, non-homogeneous focus loading is not allowable. When the edges are more heavily loaded they cause separate interference images due to their greater brightness, and these images are shifted somewhat with respect to each other, so that the impression is given of splitting or lack of sharpness of the interference lines or points. This phenomenon is illustrated in *fig. 9* by a Laue diagram [8]), in which the loading of the focus was exceedingly bad

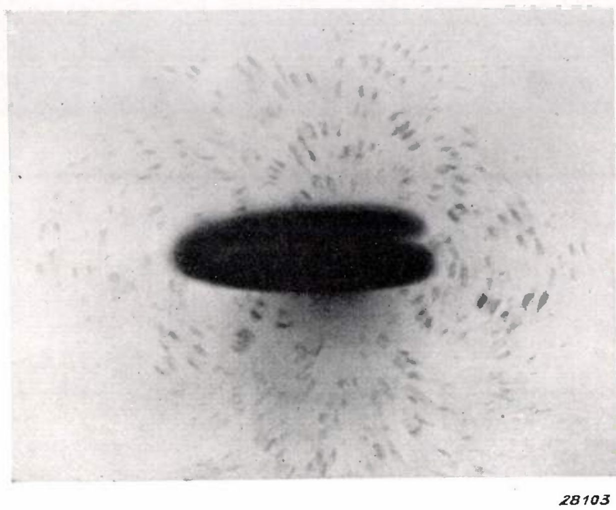


Fig. 9. A Laue diagram in which, due to very non-homogeneous loading of the focus, different mutually displaced interference images have been formed at the same time. This photograph should be compared with a good Laue photograph, such as given in this periodical 1, 213, 1936.

An ordinary spiral cathode may therefore not be used, the cathode must have such a form that a homogeneous loading of the focus is obtained with a temperature distribution like that in fig. 8. This has been achieved, in the tube described, by means of a filament the emitting parts of which lie in a single plane parallel to the anode. In *fig. 10* are photographs of a focus with non-homogeneous loading and one with homogeneous loading, as well as their brightness distributions recorded with a microphotometer.

Purity of the spectrum

In the introduction the requirement was mentioned that the X-ray spectrum produced by the tube must be pure, and reasons were given for this requirement. The bombarding electrons excite

[8]) This photograph was kindly supplied by R. Stephen of London.

in the anode metal an X-radiation with a continuous spectrum upon which is superposed a spectrum consisting of a few lines. This line spectrum is characteristic of each chemical element. In most of the applications of the tube, such, for example,

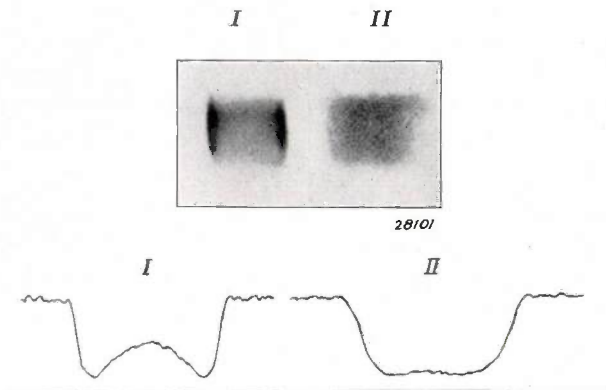


Fig. 10*a*) Photographs of a focus with non-homogeneous (*I*) and one with homogeneous (*II*) loading. The blackening is caused by the X-radiation emitted from the focus.

b) Photometer curves. The perpendicular deviation with respect to the horizontal zero line is proportional to the transmissibility at the successive points of a cross section of the photographs in *a*. Small ordinates thus correspond to heavy blackening and therefore to great X-ray intensity. With the non-homogeneously loaded focus (*I*) the brightness at the edges is considerably greater than in the middle.

as investigation by the Debye-Scherer method, a monochromatic X-radiation must be used. For every anode material a filter may be chosen of an element with a somewhat lower atomic number such that practically only one line of the characteristic spectrum is allowed to pass. If the surface of the anode is contaminated with another element, this element also emits its own characteristic radiation, which is, however, not entirely absorbed by the filter chosen, so that it is generally impossible in such a case to obtain a monochromatic radiation. For the sake of the purity of the spectrum, therefore, it is essential to avoid all contamination of the surface of the anode.

The most commonly occurring contamination is tungsten evaporated from the cathode. Since the anode is situated close to the cathode to improve the electron focussing, the evaporating tungsten can easily condense on the anode. At the usual cathode temperature of 2400 °K the rate of evaporation of metallic tungsten is still very low. If, however, the tube contains water vapour the glowing tungsten is oxidized and the oxide evaporates much more rapidly than the pure metal. The oxide condenses at a cooler spot (the anode, for example, whose focus temperature is much lower than the cathode temperature with most materials), and is here reduced to the metal again [9]). The oxygen liberated continues to attack more tungsten and makes the process progressive.

Water vapour might occur in the tube as the result of the liberation of hydrogen from metal parts and its oxidation by any oxide present; moreover, the glass may give off water vapour. A very rigorous outgassing and evacuation of the tube is therefore necessary. For this reason a getter is used in this tube, as in all "Metalix" tubes, which rapidly absorbs all gas freed during use. Due to these precautionary measures the intensity of the tungsten lines in the X-ray spectrum obtained amounts to only 1 to 2 per cent of that of the desired radiation after the tube has been in use for 1000 hours.

Moreover, a rigorous de-gassing is desirable for another reason: due to a process similar to the one described above, impurities also have an accelerating action on the evaporation of most of the anode metals, particularly on copper, so that the surface of the anode would become rough more rapidly if the tube were not well de-gassed.

Fig. 11. Photograph of the X-ray tube. The tube is drawn partially out of its container.

9) C. J. Smithells, Tungsten, 1936, p. 86.

Use of the tube

In conclusion we shall say a few words about the practical aspects of working with the tube described. The way in which complete protection is obtained against high tension and X-radiation ouside the active beam has already been discussed. The construction with a metal anode can makes it possible to place the lead jacket immediately around the can (with a glass tube the lead jacket must be at some distance from the wall as otherwise it would cause a distortion of the field and excessive local increases in the field strength). The lead jacket therefore need only be short and of small diameter, and the weight of the tube is therefore small. This has the advantage that one may work with light stands and simple suspension arrangements, and a fine adjustment on a definite point of the object such as is necessary in the measurement of stresses, is easily and quickly attained. The small size of the tube (see the photograph in *fig. 11*) also makes for ease in manipulation.

The characteristics mentioned are desirable in all X-ray tubes used in technical work. For crystal structure analysis moreover it is of advantage that the tube is quickly replaceable. It is sometimes necessary to change the type of radiation (*i.e.* the anode material), in order to adapt its wave length to the parameters of the crystal to be investigated. The tube described can be removed from its container by loosening two screws and may then be replaced by a tube with a different anode material. The water cooling system is taken out with the first tube and then transferred to the second. In this way the container remains dry.

A NEW FREQUENCY-CHANGING VALVE

by J. L. H. JONKER and A. J. W. M. VAN OVERBEEK. 621.385.5 : 621.386.694

An octode, *i.e.* a frequency changer and amplifier valve, is often used as frequency changer in superheterodyne receivers. The octode also serves as oscillator valve for generating the auxiliary frequency. In this article two objections to the use of an octode are discussed, namely the induction effect and the shift in the oscillator frequency when automatic volume control is employed. The induction effect may be neutralized in a simple way. In order to get rid of the frequency shift a new type of octode has been constructed in this laboratory in which the oscillator part and the frequency-changing part are practically independent of each other. The construction and characteristics of the new frequency changer are dealt with in detail.

Introduction

The transformation of the high-frequency oscillation into an intermediate-frequency oscillation which is applied in modern radio receiving sets is often obtained by means of an octode. This valve may be considered as a combination of a triode oscillator valve with a pentode amplifier valve, as is indicated in *fig. 1*. The oscillator, consisting of the cathode k and the grids *1* and *2* delivers a periodically varying current; as a result of this current the slope of the pentode changes, or in other words the degree to which the alternating voltage of the control grid *4* of the pentode acts upon the anode current. As one of the results of the action of grids *1* and *4* an alternating current component is obtained in the anode current, with a frequency equal to the difference between the oscillator frequency and the signal frequency on grid *4* which we shall call the signal grid.

This system of wave length transformation, when compared with other constructions in which separate valves are used for oscillator and fre-

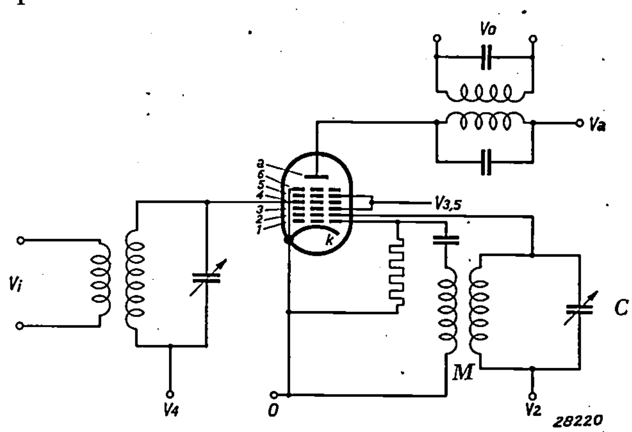


Fig. 1. Diagram of the principle of a frequency-changing stage when an octode is used. The cathode k with the grids *1* and *2* forms an oscillator triode; grids *4*, *5* and *6* form, with anode *a*. an amplifier and frequency-changing pentode. The screen grid *3* is the shield between these two parts of the octode; V_i is the input voltage and V_0 the output voltage.

Use of the tube

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The characteristics mentioned are desirable in all X-ray tubes used in technical work. For crystal structure analysis moreover it is of advantage that the tube is quickly replaceable. It is sometimes necessary to change the type of radiation (*i.e.* the anode material), in order to adapt its wave length to the parameters of the crystal to be investigated. The tube described can be removed from its container by loosening two screws and may then be replaced by a tube with a different anode material. The water cooling system is taken out with the first tube and then transferred to the second. In this way the container remains dry.

A NEW FREQUENCY-CHANGING VALVE

by J. L. H. JONKER and A. J. W. M. VAN OVERBEEK. 621.385.5 : 621.386.694

An octode, *i.e.* a frequency changer and amplifier valve, is often used as frequency changer in superheterodyne receivers. The octode also serves as oscillator valve for generating the auxiliary frequency. In this article two objections to the use of an octode are discussed, namely the induction effect and the shift in the oscillator frequency when automatic volume control is employed. The induction effect may be neutralized in a simple way. In order to get rid of the frequency shift a new type of octode has been constructed in this laboratory in which the oscillator part and the frequency-changing part are practically independent of each other. The construction and characteristics of the new frequency changer are dealt with in detail.

Introduction

The transformation of the high-frequency oscillation into an intermediate-frequency oscillation which is applied in modern radio receiving sets is often obtained by means of an octode. This valve may be considered as a combination of a triode oscillator valve with a pentode amplifier valve, as is indicated in *fig. 1*. The oscillator, consisting of the cathode k and the grids *1* and *2* delivers a periodically varying current; as a result of this current the slope of the pentode changes, or in other words the degree to which the alternating voltage of the control grid *4* of the pentode acts upon the anode current. As one of the results of the action of grids *1* and *4* an alternating current component is obtained in the anode current, with a frequency equal to the difference between the oscillator frequency and the signal frequency on grid *4* which we shall call the signal grid.

This system of wave length transformation, when compared with other constructions in which separate valves are used for oscillator and fre-

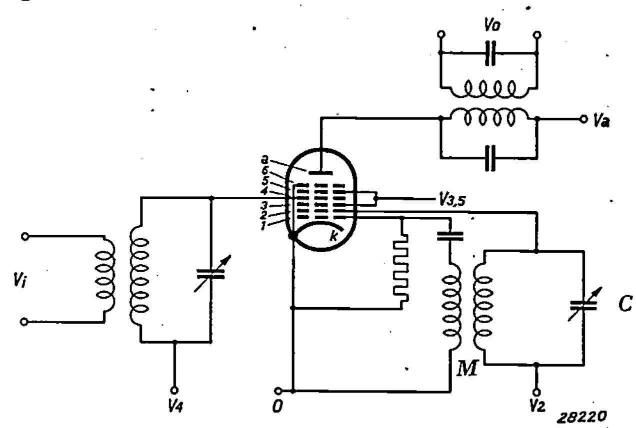


Fig. 1. Diagram of the principle of a frequency-changing stage when an octode is used. The cathode k with the grids *1* and *2* forms an oscillator triode; grids *4*, *5* and *6* form, with anode *a*. an amplifier and frequency-changing pentode. The screen grid *3* is the shield between these two parts of the octode; V_i is the input voltage and V_0 the output voltage.

quency-changing stages, has the advantage of simple construction and smaller current consumption, while a low oscillator voltage is sufficient to give a satisfactory ratio between high frequency and intermediate frequency. On the other hand there are several difficulties due to the reciprocal effects of the oscillator part and the amplifier part. The most important of these effects are the so-called induction effect and an effect on the oscillator frequency by the bias of the signal grid; the oscillator frequency varies, when for example automatic volume control is applied.

The induction effect has been discussed at length in an earlier number of this periodical [1]). This phenomenon consists in the fact that an alternating voltage of the oscillator grid *1* is induced on the signal grid *4* by the electrons which pass from grid *1* to grid *2*. This may result in the oscillator voltage also being conducted to the aerial which then radiates energy. Moreover, due to this voltage a direct current may flow to grid *1* and the amplification may be unfavourably affected.

Let us assume that grid *4* (see *fig. 2*) is at a negative potential, while the voltage of grid *1* is variable. The electrons which pass grid *1* will turn back at a definite plane in front of grid *4*; in the neighbourhood of this plane a strong negative space charge is formed which induces a positive charge on grid *4*. When the voltage of grid *1* now increases the space charge increases and therefore also the charge on grid *4*. On the other side of grid *4*, in the circuit of fig. 2, a negative charge is induced on that plate of the neutralising condenser *C* which is connected with grid *4*. The induction effect thus acts in the opposite way to the condenser and can be described by a negative capacity between grids *1* and *4*. By choosing a suitable value of the capacity *C* it may therefore be arranged that the only result of the rise and fall of the voltage of grid *1* is a passage of charge back and forth between grid *4* and the condenser plate connected with it, without any flow of current in the external circuit of grid *4*. The "negative capacity" of the induction effect is then neutralized by the positive capacity in parallel with it. This neutralization will, however, in general only be complete for a certain working point of the valve.

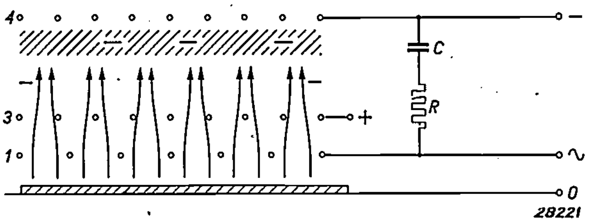


Fig. 2. Diagrammatic representation of the occurrence of the induction effect. When the voltage of grid *1* varies, the space charge due to the electrons which pass this grid and turn back in front of grid *4* also varies. As a result an alternating charge is induced on grid *4*. The induction effect may be compensated by a neutralising condenser *C* with a resistance *R* in series.

Actually the formation of the space charge in front of grid *4* occurs with a slight delay with respect to the voltage variations of grid *1* due to the transition time of the electrons. This delay may be taken into account be causing the charging of the neutralising condenser *C* to occur with a lag. This can be done by means of a suitably chosen series resistance *R*. In this way the induction effect may be so much reduced that it is no longer objectionable in practice.

The second difficulty mentioned above is an influence on the oscillator frequency by the bias of grid *4*. This may be explained as follows. Part of the electrons pass the signal grid with a low speed and the rest are reflected. Electrons may be reflected back and forth between grids *1* and *4* several times. They will finally be captured by the oscillator anode *2* or the first screen grid *3*, or they will escape between the meshes of signal grid *4*. The average number of reflections in an octode of the type EK 2 is two or three. The result of these repeated reflections is a considerable increase of the space charge in the neighbourhood of grids *1* and *2* (by which their capacity is increased) and an effect on the current-voltage characteristics of the oscillator triode part. If for example the negative bias of the signal grid is increased, the percentage of reflected electrons increases, and therefore also the space charge in the neighbourhood of grids *1* and *2*. The capacity of these grids is thereby increased, while in addition the current to grid *2* and consequently the slope of the oscillator triode changes. Both effects have an influence on the oscillator frequency [2]).

In this connection it is important to note that due to the increase in the space charge by reflected electrons the induction effect is also considerably increased. Apart from these undesired effects, the increase in the space charge also has an advantage: because of the repeated reflections in which each elec-

[1]) Some characteristics of receiving valves in short-wave reception, Philips techn. Rev. **1**, 171, 1936.

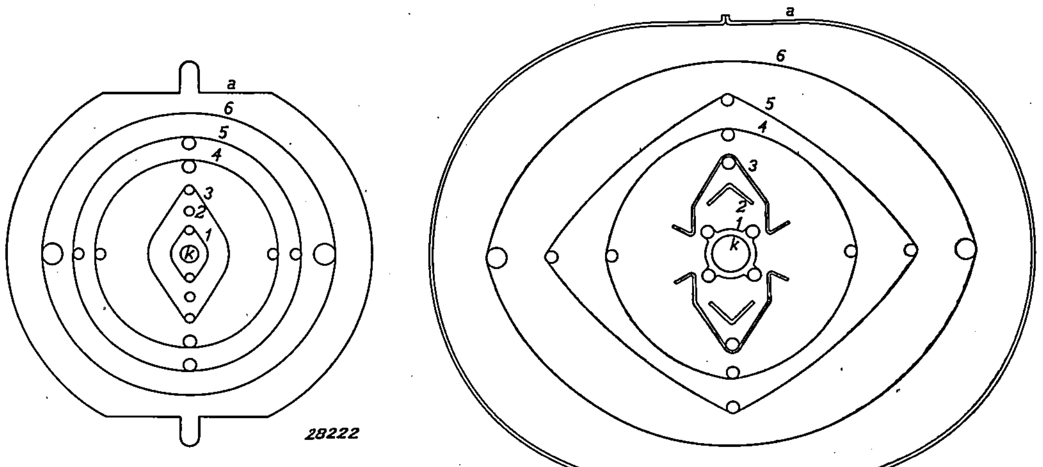
[2]) The influence of the capacity of grids *1* and *2* on the oscillator frequency may be deduced from fig. 1. The capacity of grid *2* is connected directly, that of grid *1* *via* the mutual induction *M* in parallel with the capacity *C* of the oscillator circuit. An increase therefore causes a lowering of the resonant frequency. The influence of the slope on the oscillator frequency is explained fully in the article: The Behaviour of Amplifier Valves at High Frequencies, Philips techn. Rev. **3**, 103, 1938 (*cf.* pp. 106 - 108).

tron is given several opportunities to pass or to be reflected, the slope of the anode current as a function of the voltage on the signal grid is increased.

In order to meet the objections to the octode discussed above, a modified type of octode was constructed in this laboratory, in which the electrons reflected at the signal grid are prevented from reaching the oscillator part. The frequency shift is thereby practically completely eliminated and the induction effect is considerably reduced. On the other hand the reflected electrons can no longer be utilized to increase the slope of the signal grid.

The oscillator part

Fig. 3b represents the construction of the four-beam octode (for the sake of comparison *fig. 3a* shows the construction of an ordinary type of octode, type EK 2). The oscillator grid *1* is circular like the cross section of the cathode *k*. By attaching four supports to this grid, four sharply separated beams are obtained. Grid *2*, the oscillator anode, is replaced by two plates which intercept two of the beams. The screen grid is formed by the two screens *3* which are kept at a positive potential of about 100 volts with respect to the cathode.



a *b*

Fig. 3*a*. Arrangement of the electrodes in the octode type EK 2. *k* cathode, *1* oscillator anode, *3* first screen grid, *4* signal grid, *5* second screen grid, *6* suppressor grid, *a* anode.
b) Arrangement of the electrodes in the four-beam octode, type EK 3. *k* cathode, oscillator grid, *2* oscillator anode, *3* screen grid, *4* signal grid, *5* screen grid, *6* suppressor grid, *a* anode.

The new construction, however, as we shall see later, offers new possibilities of increasing this slope, so that the slope of the new octode is actually even slightly greater than that of the old type.

The four-beam octode

The above considerations have shown that if the frequency shift upon regulation of the bias of the signal grid is to be avoided, the octode must satisfy the following two requirements: firstly the anode current of the oscillator part and secondly the capacities of the oscillator electrodes must be independent of the voltage of the second control grid.

The functioning of the oscillator part will only be independent of the action of the pentode part of the valve, if the paths of the electrons in the two parts are entirely separate. This was accomplished by dividing the electron current emitted from the cathode into four separate beams, two of which are directed to the oscillator part and two to the pentode part.

Both parts of the valve in which the oscillator current flows are in this way practically shut off from the rest of the valve.

Two small slits between the two halves of the screen grid *3* have been left open for the other two beams which are directed toward the pentode part of the valve. A current thus passes out through these slits whose intensity is dependent on the voltage on the oscillator grid. The electrons which move through the slit to the signal grid *4* are attracted by the projecting edges of the positive screen, so that they are deflected and form a fan-shaped electron beam. If an electron is reflected by the signal grid, and if its deflection is sufficiently great, it will no longer reach its starting point but be captured by the positive screen *3*. Reflected electrons therefore are no longer able to reach the neighbourhood of the oscillator grid, and in this way they are prevented from exercising any influence as a capacity on that grid.

Fig. 4 shows the paths described by the electrons in the space between grids *3* and *4*. By choosing a suitable form for the screen grid *3* and the signal grid *4*, electrons which pass through the slit only slightly to one side of its centre line may be deflected sufficiently so that they are no longer reflected to the slit. The electrons which pass through the slit at practically its centre line, and would otherwise be insufficiently deflected, are scattered by the support of the signal grid *4* (see fig. 3*b*), so that only a very small part of these electrons can be reflected back to the oscillator part.

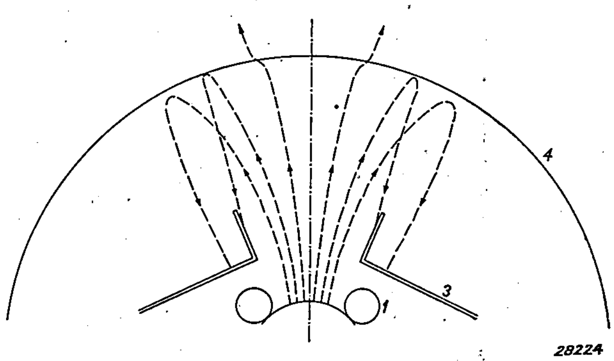


Fig. 4. Path of the electrons between the screen grid *3* and the signal grid *4* of the four-beam octode. The slightly deflected electrons pass grid *4*, while the strongly deflected electrons are reflected.

The slope of the pentode section

As was stated above, it was to be expected that the slope of the pentode part would be less with the new type of construction than in the earlier form. It was therefore important to devote attention to the various other factors which influence the slope favourably.

Upon a closer consideration of the paths of the electrons in fig. 4 it will be seen that the least deflected electrons in the case represented by the figure pass grid *4*, while the strongly deflected electrons turn back at a clearly visible distance in front of grid *4*. This was to be expected, because with increasing deflection a larger and larger part of the energy of the electrons is used in motion parallel to the plane of the signal grid, so that a smaller and smaller part of the energy remains available to overcome the potential difference between the screen grid *3* and the signal grid *4*.

With increasing voltage on the signal grid, therefore, a larger and larger part of the electron beam is allowed to pass. The slope will now be greater, the smaller the voltage interval between the grid voltage, at which the centre of the beam can just pass grid *4*, and the voltage at which the whole beam is allowed to pass.

In order to increase the slope it would now seem obvious to give grid *4* such a form that even in the case of the deflected electrons there is no motion parallel to the signal grid, so that all the electrons have the same amount of energy available to overcome the potential difference between screen grids and signal grid. This is actually possible as is shown in *fig. 5*. The electrons move perpendicular to the surface of the grid and are all reflected at the same very small distance from it. At a somewhat higher "critical" value of the grid voltage all the electrons would just be allowed to pass and the slope of the theoretical model at this value of the signal grid voltage would be infinite. Actually the slope is of course limited, in the first place because the electrons do not all have exactly the same energy, since they leave the cathode with different initial velocities (thermal motion), secondly because the distribution of potential over the surface of the signal grid is not uniform, but the potential between the wires of the grid is higher than that close to the wires, and thirdly because the reciprocal action of the electrons (space charge) has been neglected in the model. In spite of these deviations a maximum slope of the actual valve will correspond to the infinite slope of the model.

This method of obtaining a steep slope has, however, one serious objection: because the electrons are normally reflected they move along the paths on their return journey, as may be seen in fig. 5, so that they return again to their starting point, the slit, and the deflection does not have the desired effect. The tangential motion is therefore essential in order to lead the reflected electrons away from the slit.

Because of this the problem was solved in a different way, namely by altering, not the form of the signal grid, but the distribution of potential over the surface of the grid. It is possible to do this, since this potential does not depend only upon the signal grid voltage itself, but also upon the voltage and the distance of the screen grid *5*

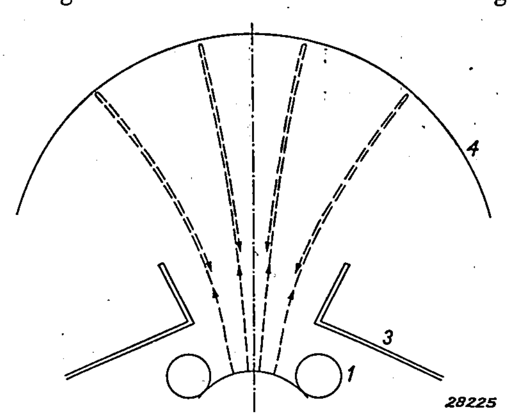


Fig. 5. Path of the electrons in the four-beam octode, when the signal grid has such a form that it is struck perpendicularly by the electrons. The electrons are reflected to the slit and are not captured by the screen grid.

lying behind it. As may be seen from fig. 3*b* this screen grid approaches the signal grid more and more closely with increasing distance from the middle of the slit, so that the effective potential

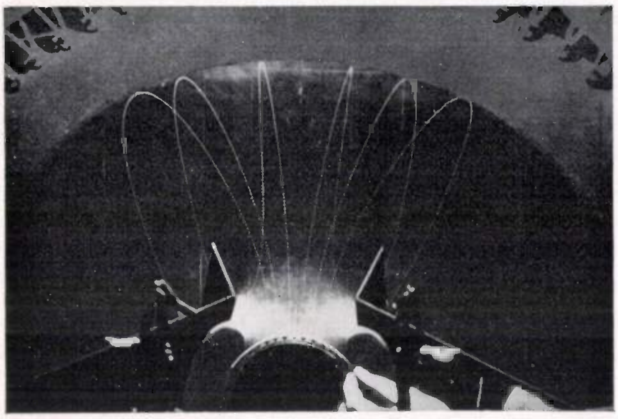


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Fig. 6. Photograph of the path of the electrons taken with a model of a four-beam octode with "adjusted" potential distribution over the signal grid. As in fig. 5 with the same critical voltage the electrons can just reach grid *4*; they are not, however, reflected back to the slit but are scattered and captured by the screening electrode *3*.

is higher at the point where the strongly deflected electrons reach the signal grid than in the neighbourhood of the middle line. Therefore the strongly deflected electrons, which have less energy available to overcome the potential difference, have a weaker opposing field to overcome.

The distribution of potential may now be so chosen that the "critical" control grid voltage is the same at all points of the grid. *Fig. 6* is a photograph of the paths of the electrons which are then obtained. This photograph was made with the help of the mechanical model described earlier in this periodical [3]). The electrons have very different tangential velocity components, nevertheless they all just reach the plane of the control grid, turn back and are then captured by the screen.

In *fig. 7* the current distribution over the anode of a valve is represented with varying potential at the signal grid, and compared with that of a valve in which the potential in the plane of the signal grid was constant. The distance between two successive curves is a measure of the slope. This figure demonstrates well that the control grid voltage at which the greatest slope is attained in the last case varies over the anode and that it increases with increasing deflection of the electrons. In the valve with adjusted potential distribution this is not the case. Moreover with an adjusted potential distribution the current density over the anode is fairly constant up to angles of deflection of 50°, except for angles smaller than 25° where the anode is very much screened by the previously mentioned support on grid *4*.

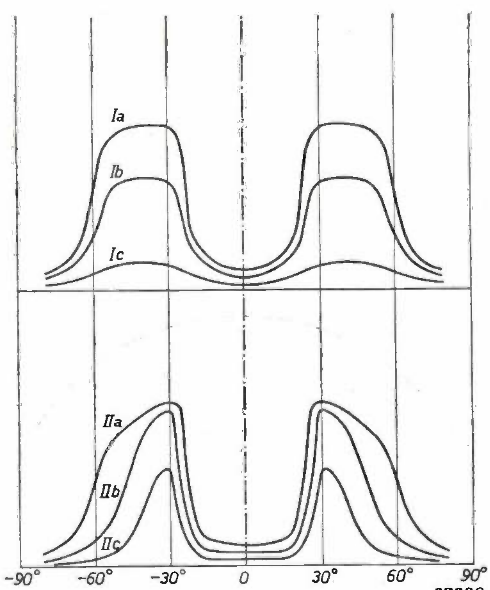

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Fig. 7. Current distribution over the anode of the four-beam octode.
I the potential in the plane of grid *4* is adjusted to the speed of the electrons,
II the potential is not adjusted.
a voltage of the signal grid is 0 volts
b ,, ,, ,, ,, ,, ,, — 2 volts
c ,, ,, ,, ,, ,, ,, — 3.5 volts

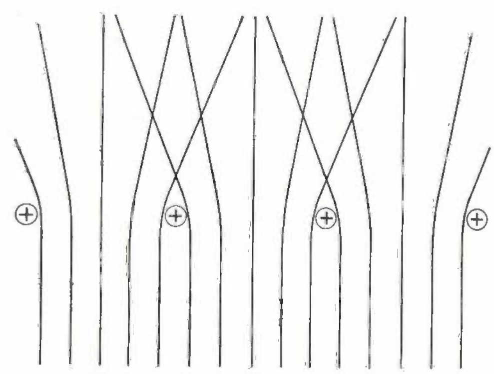

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Fig. 8. Deflection of the electrons at a grid with positive potential. Behind the grid the electrons move in criss-cross directions.

The construction of screen grid *3* with slits is an important factor governing the adjusting of the potential distribution along the signal grid. If a wire mesh grid were used instead of the screen grid the electrons would also be deflected, and would arrive on the signal grid with different speeds. The deflected electrons would, however, move in criss-cross directions (see *fig. 8*) with the result that at every point on the signal grid electrons with different perpendicular components of velocity would arrive, so that it would no longer be possible to

[3]) Philips techn. Rev. **2**, 338, 1937.

determine for every point a certain unique value of the critical potential.

Characteristic properties of the four-beam octode

In *fig. 9* the current to the oscillator anode *2* of the four-beam octode is represented as a function of the voltage of the signal grid *4*. It may be

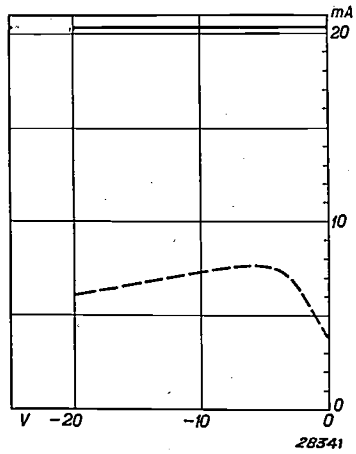

Fig. 9. Current of the oscillator anode *2* as a function of the voltage of the signal grid *4*. Full line: four-beam octode, dotted line: EK 2 octode.

seen that the oscillator current is entirely independent of the grid voltage in the pentode part, in contrast to the dotted line representing the oscillator anode current of the octode type EK 2.

In *fig. 10* is shown the change in the capacity of the oscillator grid with respect to the other electrodes as a function of the voltage of the signal grid. Here also the full line refers to the four-beam octode and the dotted line to the EK 2 octode. It may be seen that the variation in capacity is reduced to about one fourth.

The two causes of oscillator frequency shift have therefore been adequately eliminated in the four-beam octode. In other respects also the oscillator part has been considerably improved.

In the first place the slope has been considerably increased, in the second place the transition time of the electrons from the cathode to the oscillator anode has been shortened, since the electrons can follow a straight path which was impossible in the earlier construction as shewn in fig. 3*a*, where the supporting rod of grid *1* stood in the way.

Due to the greater slope the back coupling is more intense when the same coils are used, so that with fewer turns on the coil a sufficiently high oscillator voltage can be obtained. The minimum capacity of the oscillating circuit therefore becomes smaller, so that the ratio between the longest and the shortest wave length of a band which can be compassed with a given condenser range becomes greater. Furthermore, especially in the short-wave region, the frequency change upon fluctuations in the mains voltage is smaller because, due to the shortening of the transition times of the electrons, the effect of these times on the oscillator frequency is less.

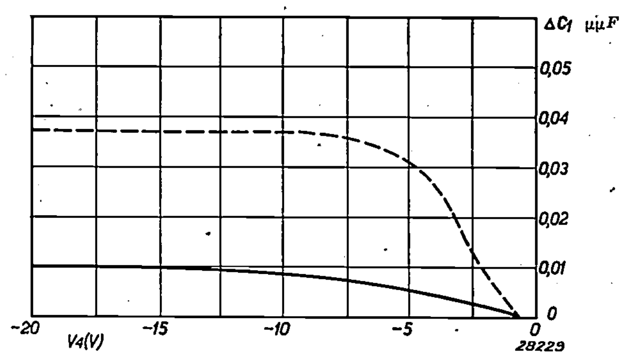

Fig. 10. Change in the capacity of the oscillator grid as a function of the voltage of the signal grid. Full line: four-beam octode, dotted line: EK 2 octode.

Since the space charge between the two grids *1* and *4* is less, the induction effect will also be present to a smaller degree, as mentioned above, while the remainder of the effect can be still further diminished by connecting these grids by means of a condenser and resistance in series.

LOW-PRESSURE MERCURY DISCHARGE WITHIN A LUMINESCENT TUBE

by W. UYTERHOEVEN and G. ZECHER. 537.525.8 : 621.327.3 : 537.37

Column discharges in mercury at low pressure, due to the strong ultraviolet resonance line 2537 Å, are particularly suitable for excitation of the luminescence of zinc silicate and other phosphorescent substances. The process of this light excitation is discussed in this article, as well as the various factors which determine the efficiency. The following also are dealt with: the use of phosphorescent substances in order to improve the efficiency and colour and for the purpose of rendering alternating current discharges free of flicker.

In the discussion in a previous article [1]) on the light emission in the positive column several cases were dealt with in which a considerable portion of the electrical energy supplied was transformed directly into visible light by excitation of favourably situated levels. In addition to the visible spectral lines, however, there are practically always other lines also which lie in the ultraviolet or infrared part of the spectrum, and which therefore do not contribute directly to the light flux. Attempts have been made to utilize this invisible radiation, and success has been attained in the case of the ultraviolet part by the use of luminescent substances. When these substances are exposed to radiation of certain wave lengths in the ultraviolet, they emit radiation in another region with longer wave lengths in the visible. Luminescent substances thus act as transformers of radiation. with the limitation that the wave length of the radiation is practically always increased in the transformation (Stokes' law). In this way it is possible to improve considerably the efficiency of certain discharges.

It has been found that the production of ultraviolet radiation and its transformation into visible light can take place in such a way that discharges which themselves emit almost no light can be made very efficient sources of light. As an example we shall discuss the low-pressure mercury discharge within luminescent boundary, confining our discussion to the column itself. Various factors, such as ignition and connections, which are very important when these discharges are used in lamps, will not be considered here, since they have no direct effect on the physical phenomena which occur in the column with a fluorescent wall.

Luminescent substances

The phenomenon of luminescence and the properties of various substances which exhibit this phenomenon, have been recently described in this periodical [2, 3]). We shall confine ourselves here to those substances which are used as light transformers in the low-pressure positive column of the mercury discharge. We are then chiefly concerned with "impurity luminophores" and "fluorescent glass". A typical impurity luminophore is zinc silicate. The zinc silicate is in this case the base material, which upon being mixed with small amounts of an "activator", manganese for example, and after a certain treatment, acts as a luminescent substance. The treatment consists mainly in a crystallization by heating to a high temperature. *Fig. 1* shows the emission curves for several zinc beryllium silicates activated with manganese in a low-pressure mercury discharge, in which the radiated energy is plotted as a function of the wave length; the maximum is put equal to 100. It may be seen that the emission takes place over a wide range of wave lengths, which is of great advantage for many purposes. In order to determine from these curves the (relative) light emission curves it is only necessary to multiply the ordinates

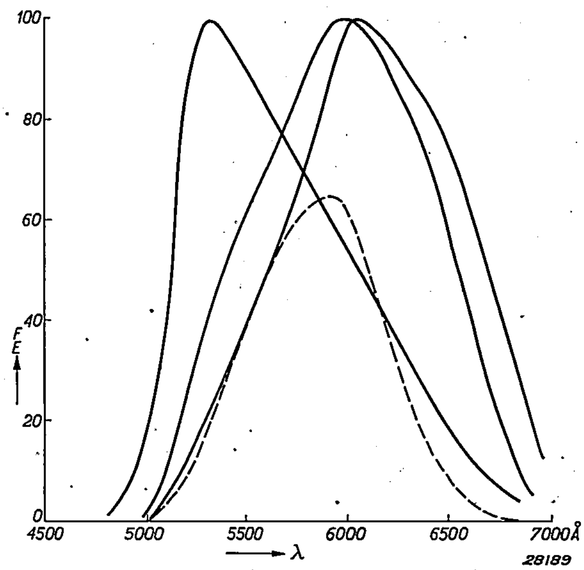

Fig. 1. Emission curves: energy as a function of wave length (full lines; maximum put equal to 100), for several zinc beryllium silicates activated with manganese in a low-pressure mercury discharge. The dotted line gives the light intensity as a function of wave length for the phosphorescent substance with its energy maximum at 6050 Å. At the wave length of maximum eye sensitivity the light intensity is put equal to the energy.

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1) W. Uyterhoeven, Philips techn. Rev. 3, 201, 1938.
2) W. de Groot, Philips techn. Rev. 3, 125, 1938.
3) J. H. Gisolf and W. de Groot, Philips techn. Rev. 3, 245, 1938.

by the corresponding values of the relative sensitivity of the eye, as has been done for the curve with the maximum at 6050 Å (dotted line in fig. 1).

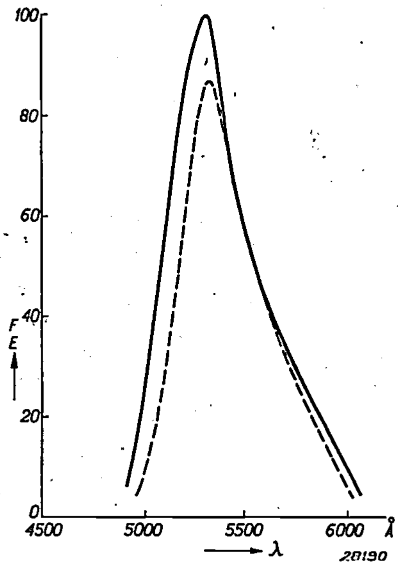


Fig. 2. Emission curves: energy (full line; maximum put equal to 100) and light (dotted line) as functions of the wave length for uranium glass irradiated with the light of a low-pressure mercury discharge.

A typical luminescent glass is uranium glass, in which certain groups of molecules of the added uranium compound are luminescent. In *fig. 2* may be seen the relative emission curves as a function of the wave length for uranium glass in a low-pressure mercury discharge. For further details we refer to the publications already mentioned. Other substances, such as copper and tin compounds, when added to glass can make it strongly fluorescent.

The production of light in luminescent column discharges

The production of visible light in positive columns with fluorescent substances is a rather complicated process in which we can distinguish several successive steps. In the first place there is the production of the ultraviolet radiation in the discharge, then the absorption of this radiation by the luminescent layer, which after a longer or a shorter time emits a part of the energy absorbed as visible light. In each of these steps in the transformation of the electrical energy supplied to the discharge into visible light, there will in general be a certain loss. It has, however, been found that an extraordinarily good efficiency can be attained, which, expressed in the usual way in lumens emitted per watt supplied, may reach 100 lumens per watt. We shall now study how this high efficiency is achieved.

The first step in the process of light excitation is the production of ultraviolet radiation, which, in the low-pressure mercury discharge, consists mainly of the resonance line 2537 Å (partially also of the resonance line 1850 Å).

In *fig. 3* is given the relative distribution of energy for a typical mercury discharge at low pressure, in which, in addition to the line 2537 Å, the intensity of which is taken as 100, several lines with longer wave lengths are also indicated. In order to obtain good efficiency in the production of the resonance line, the discharge will be made to occur preferably at a low current density. One of the results of this is that the energy taken up per cm of length is small (and therefore also the light flux produced), and the tube has a low brightness. A lamp with luminescence excited by the ultraviolet radiation of the low-pressure mercury discharge will therefore be made chiefly in relatively small units of, for instance, 25 watts. With a suitable choice of the mercury vapour pressure (*i.e.* the temperature), the tube diameter, and discharge current it is possible to transform about one half of the electric energy supplied per cm length of tube into radiation with a wave length of 2537 Å. If it is desired to excite luminescence by means of radiation in the near ultraviolet, for instance in the neighbourhood of 3600 Å, it is better to use a mercury discharge with high or very high pressure. For these lines as well as for the visible mercury lines, the considerations discussed in the article mentioned previously [1]) are valid. According to these considerations high pressure and current density are necessary for economical production of lines of longer wave length.

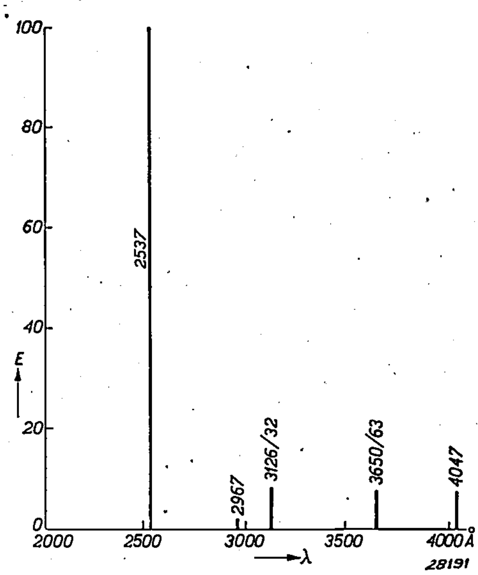


Fig. 3. Relative energy distribution of a typical low-pressure mercury discharge; intensity of the line 2537 Å set equal to 100.

In order that all the ultraviolet radiation may be used, it must all be absorbed by the luminescent layer, which may therefore not be too thin. It is, however, not always possible to use a sufficiently thick layer since such a thick layer may offer difficulties by not adhering to the inner wall. It may also happen that the phosphorescent substance absorbs too much of its own radiation so that light is lost. In these cases it is possible to use a thin layer of luminescent material and, instead of ordinary glass, a fluorescent glass; the ultraviolet radiation which is not absorbed by the phosphorescent substance is then transformed into visible radiation in the glass.

Efficiency of luminescence

There are various ways of defining the efficiency in the transformation of the ultraviolet radiation absorbed by the phosphorescent substance into visible light. One may determine how many quanta of visible light are emitted per 100 quanta of ultraviolet light absorbed. If the result is 80 for instance, the "quantum efficiency" is 80 per cent. Such values actually occur, and even values up to 100 per cent.

Even when the quantum efficiency is 100 per cent, the ratio of the energy radiated as visible light to the energy absorbed as ultraviolet light, the "energy efficiency", is considerably less than unity. If for every ultraviolet quantum with the frequency ν_1, one visible quantum with the frequency ν_2 is produced, where $\nu_2 < \nu_1$, the energy ratio is $h\nu_2/h\nu_1 = \nu_2/\nu_1$, since according to Planck's formula the energy of a quantum of radiation with the frequency ν is $h\nu$. Since the wave length is inversely proportional to the frequency the energy efficiency is $\nu_2/\nu_1 = \lambda_1/\lambda_2$. Thus when the wave length absorbed (λ_1) is 2537 Å and the wave length emitted (λ_2) is 5550 Å the energy efficiency is $\lambda_1/\lambda_2 = 2537/5550 = 0.46$.

If it is desired to express the efficiency in lumens per watt, the position and shape of the emission curve of the energy with respect to the eye sensitivity curve must be taken into account. The highest yield of light with a given energy efficiency is obtained with a narrow emission curve which has a maximum in the neighbourhood of the wave length $\lambda = 5550$ Å of maximum eye sensitivity. The colour of the light emitted by the phosphorescent substance is then a pronounced green which may be undesirable for some purposes.

The following example in figures may give some idea of the results attainable with the phosphorescent substances now used. We assume that 50 per cent of the electrical energy used in the column is transformed into radiation of the wave length 2537 Å, that the quantum efficiency is one, and the energy efficiency is 50/100 [4]), and that for each watt radiated in the visible 200 lumens are obtained. We then find for the total efficiency of the transformation: $\frac{50}{100} \cdot 1 \cdot \frac{50}{100} \cdot 200 =$ $= 50$ lumens per watt.

Adaptation of the discharge to the luminescent substance

In principle it is possible to find for a given phosphorescent substance a discharge, the emission of which in the ultraviolet has a favourable position with respect to the absorption spectrum of the phosphorescent substance. Such a substance absorbs mainly the wave lengths in a certain region several hundred Ångstrom units wide, in which region the absorption has a maximum. *Fig. 4* shows schematically the absorption spectrum of a zinc beryllium silicate. The maximum is considered to be 100. It may be seen that maximum absorption occurs in the neighbourhood of the line 2537 Å, which may be of advantage since the luminescent layer can have a minimum thickness when this wave length is used. In the case of the silicates and tungstates, activated with metals or not, it may in general be said that their absorption region in the ultraviolet lies at wave lengths shorter than 3000 Å. For these substances therefore the mercury discharge at

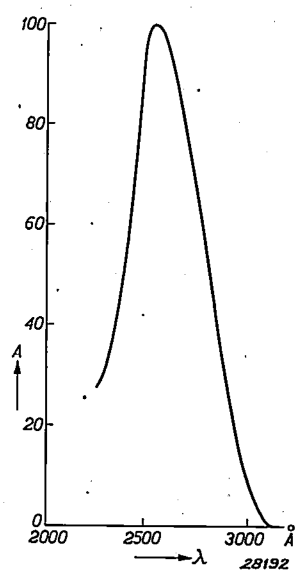
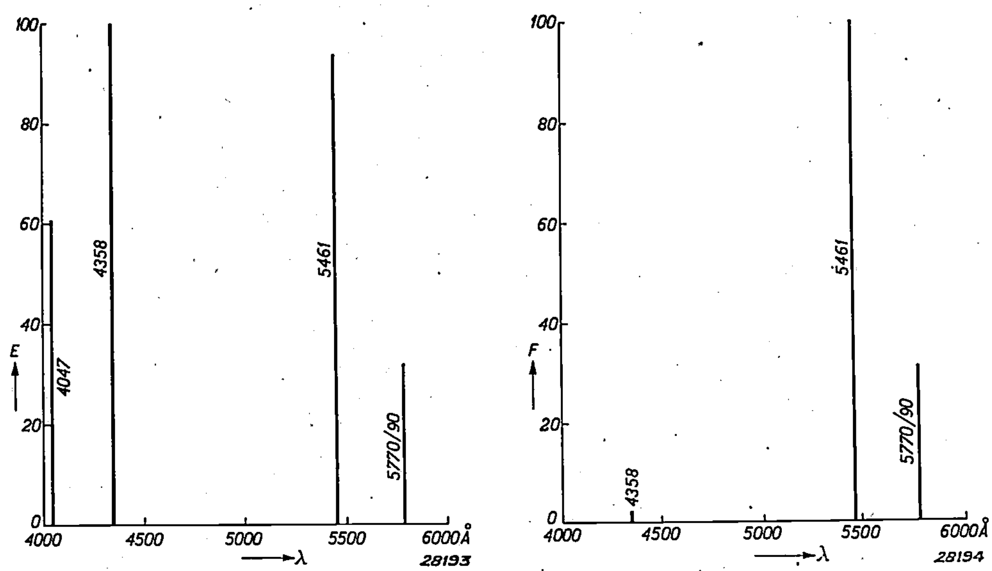


Fig. 4. Schematic representation of the absorption curve for a zinc beryllium silicate, maximum put equal to 100.

4) In an accurate calculation it must be taken into account that the energy efficiency λ_1/λ_2 is not the same for all wave lengths λ_2 of the emission band.

low pressure with the strong line at 2537 Å is the most suitable source of radiation for the excitation of light. Phosphorescent substances in which the basic material is a sulphide, for example the green phosphorescent zinc sulphide activated with copper, are, however, best rendered phosphorescent by irradiation with ultraviolet light of long wave length (3600 Å). For this purpose the mercury discharges at high and very high pressure are the best sources of radiation, since they have very strong emission lines in the neighbourhood of 3600 Å.

of purposes (for outdoor illumination, for instance). For illuminated advertising signs attempts are even expressly made to obtain various colours. For interior illumination, however, a colour reproduction more or less true to nature is indispensable, even if it is obtained at the cost of efficiency. In this case the adaptation of the phosphorescent substance to the discharge will therefore consist in the fact that the light emission of the discharge together with that of the phosphorescent substance must give a mixed light with which the observation of colours is satisfactorily true to nature.



a b

Fig. 5. Visible spectrum of the mercury discharge at low pressure: *a*) relative energy distribution, *b*) relative light emission, maximum put equal to 100.

In practice, however, we begin with a given discharge, and try to find a phosphorescent substance which is suitable. In doing this we may have different purposes in view: either improvement of the e f f i c i e n c y or improvement of the c o l o u r. In the first case the conditions which the luminescent substance must satisfy are obvious from the above discussion. The maximum in the absorption curve of the substance must lie in the neighbourhood of the maximum of the ultraviolet emission of the source of radiation, with low-pressure mercury discharges therefore in the neighbourhood of 2537 Å. The emission curve of the phosphorescent substance, which must have as high a quantum efficiency as possible, must have its maximum in the neighbourhood of maximum eye sensitivity (at $\lambda = 5550$ Å). In this way a very high efficiency is obtained, as high as 100 lumens per watt, but strongly coloured (green) sources of light are also obtained. This is not a great objection for a number

In *fig.* 5, *a* and *b*, may be seen the composition of the visible spectrum of the mercury discharge at low pressure, in fig. 5*a* the relative energy distribution and in fig. 5*b* the relative light emission. The influence of the low eye sensitivity at the extremities of the spectrum may be clearly seen upon comparing the two diagrams: with respect to energy the blue line 4538 Å is the strongest, with respect to light the green line 5461 Å is strongest.

Block method of colour determination [5])

In the development of sources of light consisting of a positive column discharge with a fluorescent boundary, the socalled "block method" for the determination of colour is very useful. Since the

[5]) For the block method see Philips techn. Rev. 2, 1, 1937. The limiting wave lengths have been changed somewhat since that article was written, so that the mercury lines fall about in the middle of the blocks, which improves the accuracy of the method.

measurement of complete emission curves is a fairly elaborate procedure, these curves have been replaced by a kind of block diagrams. The whole

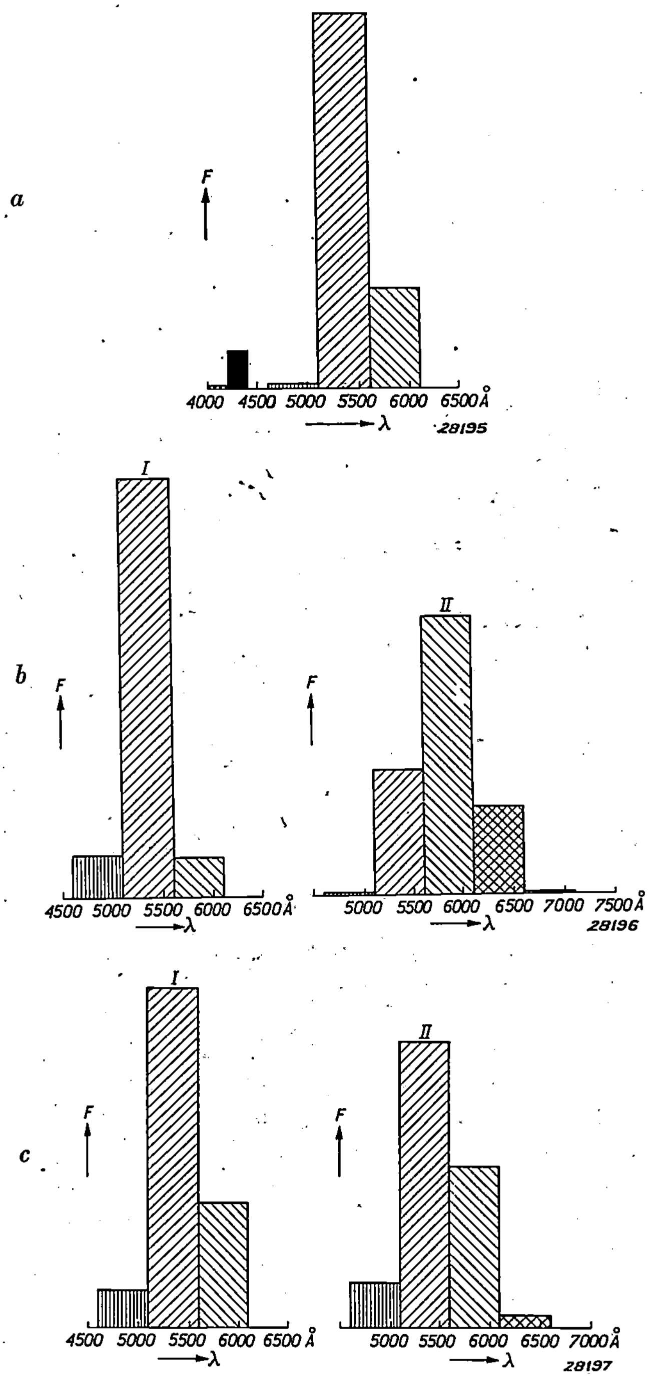


Fig. 6. Block diagrams of the light emission for *a*) a low-pressure mercury discharge *b*) two phosphorescent substances: zinc silicate (*I*) and zinc beryllium silicate (*II*), *c*) two luminescent glasses: uranium glass (*I*) and a combination of copper and uranium glass (*II*).

region of the visible spectrum is divided into eight blocks limited by the following wave lengths: 4000 - 4200 - 4400 - 4600 - 5100 - 5600 - 6100 - 6600 - 7200 Å. A spectrum of the light source to be investigated is thrown on a photocell by means of a double monochromator. In the path of the light rays various diaphragms are placed, each of which passes only the wave lengths which belong in one of the eight blocks. By a suitable choice of the height of each diaphragm, the deviation of galvanometer or electrometer due to the current of the photocell may be made proportional to the light flux emitted in each block, whereby the eye sensitivity and the spectral sensitivity of the photocell are taken into account. By means of a light source with a known energy distribution, a tungsten lamp for instance, the installation may be calibrated.

In *fig. 6* the block diagrams determined by this method are given for several cases. Fig. 6*a* represents the light emission of a low-pressure mercury discharge; it may be compared with fig. 5*b*. It may clearly be seen on the block diagram that the light flux is greatest in the green (5100 - 5600 Å), that there is, however, also a quite appreciable amount in the yellow region (5600 - 6100 Å) and in the blue (4200 - 4400 Å). Next to this in fig. 6*b* may be seen block diagrams for the radiation of two phosphorescent substances (without mercury light), namely zinc silicate (*I*) and zinc beryllium silicate

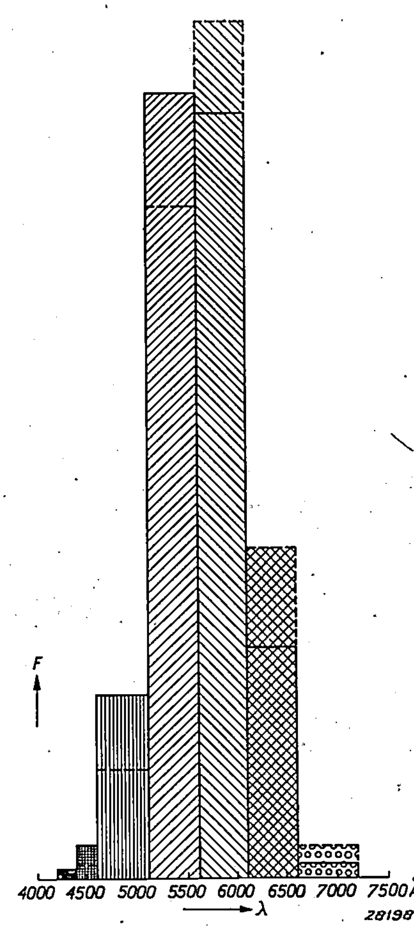


Fig. 7. Block diagram for sunlight (full line) and electric light, Bi-Arlita lamp type 78 W (dotted line).

(*II*), and in fig. 6*c* for the radiation of two phosphorescent glasses, namely uranium glass (*I*) and a combination of copper glass with uranium glass (*II*).

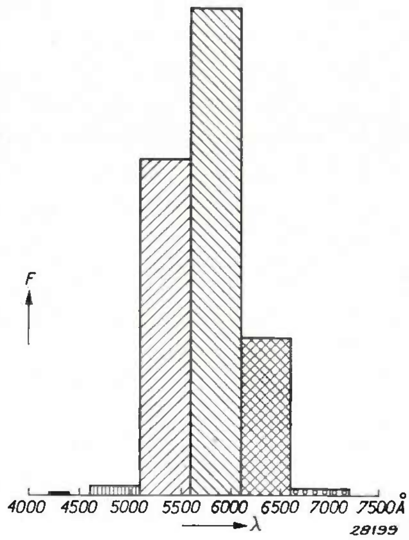


Fig. 8. Block diagram of the light of a mercury tube made of copper-uranium glass and containing zinc beryllium silicate.

The block diagram is very useful, particularly in judging the results of combining different luminescent substances. If one has as object the construction of a source of light which shall resemble sunlight or ordinary electric light for example, one begins by making the block diagram of these sources of light as is indicated in *fig. 7*. In the various experiments it is now possible in each case to find out whether the composition of the light satisfies the necessary requirements, and in which region any deficiency or excess occurs. If, for example, the dotted line of fig. 7 is compared with the diagram for zinc beryllium silicate of fig. 6*b*, it may be seen that the emission of the substance without the mercury spectrum resembles ordinary electric light fairly well. In this case the mercury light itself contributes only 15 per cent to the total light flux, so that the colour is practically determined by the phosphorescent substance. The combination of the light of the phosphorescent substance, glass and mercury discharge, as produced by a lamp for practical uses is finally given in *fig. 8*.

Alternating current light sources rendered free of flicker

In the foregoing we have only discussed the property of luminescent substances known as "light transformation". At the same time a number of these substances may also act as "light accumulators", *i.e.* they luminesce not only while they are being irradiated, but they continue to emit light for some time after the irradiation has been interrupted. The phosphorescent substance is therefore able to store the energy for some time and then emit it gradually. The time during which the phosphorescent substance continues to luminesce varies very much from one substance to another. *Fig. 9* gives the course of

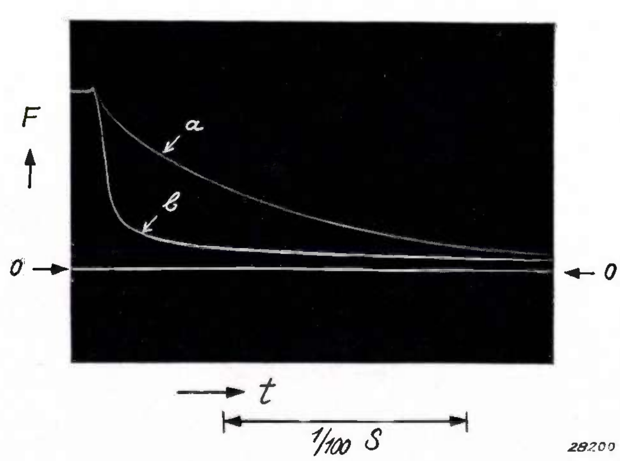


Fig. 9. Variation of the light intensity after interruption of the irradiation with ultraviolet light: *a*) for a phosphorescent silicate, *b*) for a phosphorescent sulphide.

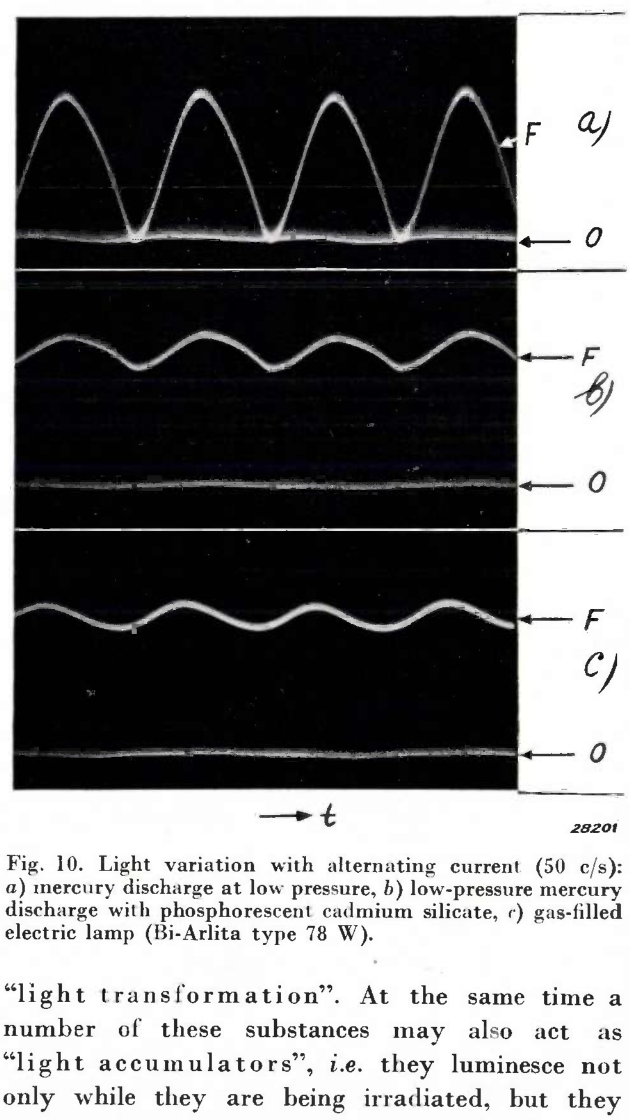


Fig. 10. Light variation with alternating current (50 c/s): *a*) mercury discharge at low pressure, *b*) low-pressure mercury discharge with phosphorescent cadmium silicate, *c*) gas-filled electric lamp (Bi-Arlita type 78 W).

the light emission after interruption of the irradiation for two types of substances: silicates (a) and sulphides (b). The silicates especially possess this property; tungstates on the other hand are practically without it.

The phenomenon of phosphorescence of these substances may be used to reduce the flicker with alternating current lamps. When the direction of the current is reversed, the current is zero for a short time, and practically no atoms are excited, so that the light emission becomes zero, since the phosphorescence of the discharge itself may usually be neglected. *Fig. 10a* shows the variation of the light with time in a low-pressure mercury discharge. In each half period the curve has practically the form of a sine curve in which the moments when the light is zero correspond to those when the current passes through zero. This periodic variation of the light intensity may be very disturbing in certain cases, especially in interior illumination. If the variation of the light from a low-pressure mercury discharge the boundary of which is covered with a luminescing layer (a cadmium silicate) is recorded in the same way, the diagram of *fig. 10b* is obtained. The fluctuation in the light is now reduced to a ripple which is much smaller than the average constant value upon which it is superposed. As a measure of the irregularity one may choose the ratio of the difference a between maximum and minimum to the value b of the minimum. In the case of fig. 10b the irregularity a/b defined in this way amounts to about 0.25. It has been found that under the prevailing circumstances such a value for the irregularity may be accepted without hesitation. For the sake of comparison the variation of the light is given in fig. 10c, for a gas-filled electric lamp (Bi-Arlita 78 W), for which the ratio a/b is about 0.15.

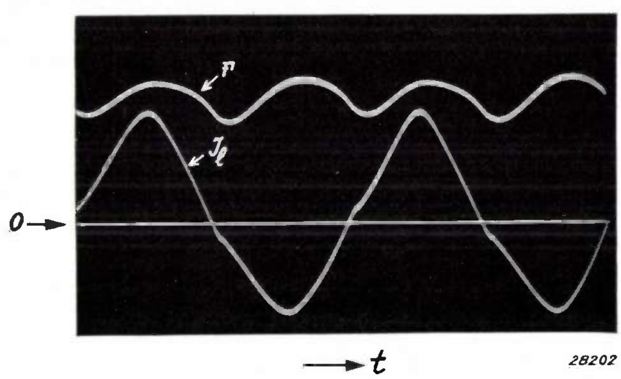


Fig. 11. Variation with time of the discharge current I_l and light flux F (sum of mercury light and luminescence light) for a low-pressure mercury discharge with phosphorescent zinc beryllium silicate (50 c/s).

In conclusion *fig. 11* gives the variation with time of the discharge current I_l, to which the intensity of the line 2537 Å is practically proportional, and of the combined light flux F of mercury and luminescence light from a tube containing zinc beryllium silicate. It may be seen clearly that there is a certain phase shift between the maxima of I_l and F, which facilitates the bridging over of the light minimum during the currentless period. The cause of the phase shift lies in what might be considered the reverse of the phosphorescence. Upon irradiation with a given, not too high, intensity of ultraviolet radiation, it takes some time before the phosphorescent substance reaches its full activity and the light emission reaches its maximum value, so that the light maximum lags somewhat behind the ultraviolet maximum.

THE TENSILE STRENGTH OF DEPOSITED WELD METAL

by J. SACK.

539.412 : 621.791.052

In this article it is shown that it is no longer correct to require in specifications an upper limit for the tensile strength of deposited weld metal. This requirement originated from an assumed relation between the variation with composition of the tensile strength and the brittleness of the material. The relation is indeed valid for ordinary carbon steels such as have long been used, but it no longer holds for modern welding materials. If this stipulation is made it leads only to the use of welding rods of poorer quality than would be used if the obsolete requirement were omitted.

Introduction

The specification accompanying an order must give the minimum conditions to be satisfied. In the case of welding rods for making joints in iron and steel the specification usually gives various requirements regarding the mechanical properties of the metal deposited and of the welded joint. For example it may be required that the tensile strength of the material deposited must be at least 42 kg/sq.mm., its elongation at least 25 per cent, and its notched bar toughness more than 8 kg/sq.cm. The elongation allows us to judge the capacity for deformation of the metal, while from its resistance to notching its toughness follows. Tensile strength and malleability are considered to be important properties of the metal and its quality increases in proportion to these characteristics properties.

It is obvious that the specifications must contain the minimum value of all these mechanical quantities. It is, however, strange that an upper limit should also be given for the tensile strength of the weld metal deposited. It would be unnecessary to concern oneself with this upper limit if only the requirement were sufficiently high ,but it often occurs that the specifications demand such a low upper limit for the strength of the weld metal that good welding rods fail to fulfil the requirement. The question therefore is why the tensile strength should not for instance exceed 47 kg/sq.mm and be 50 kg/sq.mm. The following will help to give an idea of the reason why specifications often contain such seemingly absurd requirements.

Relation between tensile strength and other mechanical properties

The tensile strength has always been the most important property for judging the quality of iron and steel. It formed the basis of the classification of steels, and on this basis the builder always made his calculations. Because tensile strength is such an important property of the metal it is understandable that attempts were made to correlate it with other properties of the material, in the first place with the hardness, *i.e.* with the resistance experienced by a ball or a diamond point upon being pressed into the material. The harder the material the greater its tensile strength. It is common knowledge that the tensile strength and the hardness of a metal vary equally with its composition.

Table I

Mechanical properties and chemical composition of several modern steels

Kind of steel		R.D.S. steel	Cor-ten steel	Cr-Va steel	Cromansil steel	Two examples of St 54	
Tensile str. (kg/sq.mm)		57.4	45-52	56-63	56-63	55.5-59.5	53.8-56.5
Yield (kg/sq.mm)		47.5	35-42	42-49	38-42	36.8-40.3	37.4-39.8
Elongation (%)		33	22-27	25-35	20-28	24.5-30.0	20.9-27.3
Decr. in cross section(%)		54	40-55	50-55	50-62	—	—
Percentage	C	0.095	0.10	0.10-0.20	0.12-0.20	0.18	0.19-0.21
	Mn	—	0.10-0.30	0.30-0.60	1.1-1.4	0.60	0.74-0.76
	Si	—	0.5-1.0	0.20-0.30	0.4-0.6	0.12	0.04-0.07
	Cr	—	0.5-1.5	0.80-1.10	0.7-0.8	0.62	0.23-0.28
	Cu	1.44	0.3-0.5	—	—	0.60	0.42-0.43
	Ni	0.75	—	—	—	—	—
	Mo	0.70	—	—	—	—	—
	Va	—	—	0.15-0.20	—	—	—
	P	0.04	0.1-0.2	0.04	—	0.04	0.06
	S	0.04	—	0.045	—	0.03	0.02

Table II

Mechanical properties and chemical composition of deposited weld steels

Material deposited		Welding rod "A"			Welding rod "B"			
		1	2	3	1	2	3	4
Tensile strength (kg/sq.mm)		46.8	47.6	52.1	48.2	48.4	51.3	52.3
Yield (kg/sq.mm)		37	37	42	38	38	40	39.5
Elongation (%)		26	27	26	22	26	29	28
Decrease in cross section (%)		42	44	42	45	51	55	52
Percentage	C	0.088	0.070	0.068	0.046	0.043	0.052	0.070
	Mn	0.44	0.47	0.48	0.40	0.41	0.48	0.44
	Si	0.14	0.13	0.21	0.08	0.09	0.13	0.15
	P	—	—	—	0.022	0.021	0.022	0.014
	S	—	—	—	0.016	0.022	0.016	0.019
	N	0.024	0.029	0.036	0.021	0.020	0.072	0.068

In the second place the tensile strength is related to the elongation at the breaking point, *i.e.* to the degree to which the metal can be deformed before it breaks. In the third place the tensile strength is related to the notched bar toughness, which quantity indicates the resistance of the metal to varying loads. With increasing tensile strength, elongation and toughness decrease. In the fourth place the tensile strength of a metal is related to its workability, which decreases with increasing tensile strength and hardness. Poorer workability is manifested for instance in the fact that the cutting speed must be reduced.

All these properties are related to each other in as far as they depend on the content of carbon which is one of the most important components added to iron to make steel. The greater the carbon content the greater the tensile strength and the smaller the elongation and toughness. This may be seen in *fig. 1* which refers to annealed carbon steel with 0.15 - 0.35 per cent of silicon and 0.50 - 0.70 per cent of manganese [1]).

If steel is hardened, a hardening structure is obtained, *i.e.* a martensite, troostite or sorbite structure, depending on the rate of cooling. Martensite is very hard and brittle, while pearlite which may be obtained by annealing is less hard and tougher. In welding, the metal on both sides of the weld is heated to a rather high temperature and then cooled more or less rapidly.

If welding is done with steel having a high carbon content, hard and brittle zones must be expected. In order to avoid these undesirable structures the rate of cooling of the piece of work and the weld must be decreased. This is accomplished in practice by pre-heating. For these reasons carbon is a detrimental component in either the piece of work or the weld.

Modern steel industries pay particular attention to the development of weldable steels, and with this end in view they limit the content of carbon to, for example 1 0.20 per cent. It is now even possible to obtain steels of high strength whose

[1]) E. Houdremont, Einführung in die Sonderstahlkunde.

Table III

Mechanical properties and chemical composition of the material deposited from different types of welding rods

Type of welding rod		PH 38	PH 40	PH 46	PH 48	PH 50
Tensile strength (kg/sq.mm)		38-42	38-42	40-46	48-52	48-52
Elongation (%)		10-13	18-22	22-26	24-28	28-32
Notched bar toughness (kg/cm/sq.cm)		—	4-6	6-8	11-14	10-13
Bending angle (°)		—	60	90	180	180
Percentage	C	0.015	0.03	0.04	0.07	0.07
	Mn	0.07	0.02	—	0.50	0.40
	Si	—	0.01	—	0.15	0.10
	P	0.04	0.04	0.04	0.02	0.02
	S	0.04	0.04	0.04	0.02	0.02

carbon content is low. This may be seen in the case of the steels 52 and 54 which possess tensile strengths of more than 52 and 54 kg/sq.mm respectively. In America steel 60 is being used more and more. All these types of steel have been developed on the basis of a low carbon content, so that the high tensile strength is obtained by the addition of other elements such as chromium, silicon, copper, nickel, manganese, vanadium, etc.

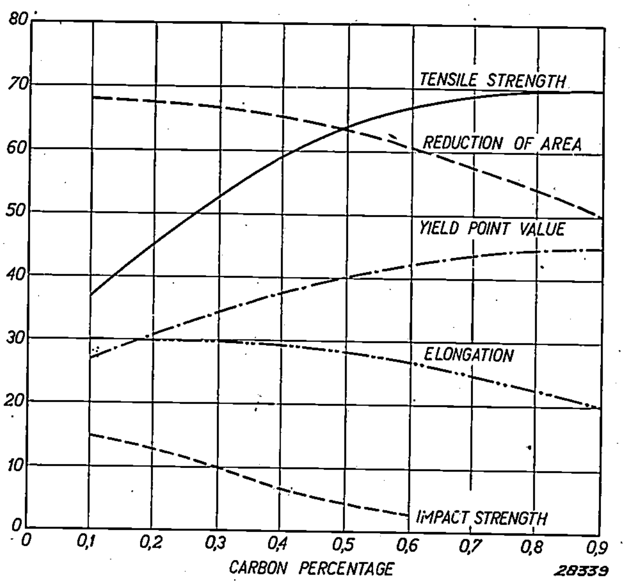


Fig. 1. The following properties are plotted as functions of the per cent of carbon: the tensile strength in kg/sq.mm, the reduction of area at the breaking point in per cent, the yield in kg/sq.mm, the elongation at the breaking point in per cent and the notched bar toughness in kg/sq.mm.

In *table I* several examples are given of modern steels which do not contain more than 0.2 per cent of carbon. Many of these steels show a remarkably high malleability and toughness.

Like the steel manufacturer, the manufacturer of welding rods must provide for a low carbon content of the fused weld metal. Usually the content is not higher than 0.1 per cent, as may be seen in *table II*. This table contains several experimental results of extensometer tests carried out in this laboratory on welding metal fused down into rods, as well as their chemical composition, including the nitrogen content. It may be seen from the table that tensile strengths of 47 to 52 kg/sq.mm. may be attained with steels containing less than 0.09 per cent of carbon. The material with the highest tensile strength contains less than 0.07 per cent of carbon. It may be seen moreover that small elongation at the breaking point need not accompany high tensile strength.

In *table III* several mechanical properties and the chemical composition of the material deposited from several types of welding rods have been collected. Welding rods with a thin coating (type PH 38 and PH 40), with thick coating (type PH 46 and PH 50) and also welding rods with organic components in the coating (type PH 48) are included. It may be seen that in general, with the welding rods in question, those with a high tensile strength also have the best mechanical properties in other respects, namely the greatest elongation and angle of bending. In any case the rule that a hard metal is also brittle is found not to be valid here. It is valid for annealed or rolled carbon steel, but not for the fused down weld metal in question.

It is true that there is a more or less close relation between the various mechanical properties of carbon steels, but it does not follow therefore that the tensile strength can be used as a measure for all the other mechanical properties, and it is certainly wrong to do so in the case of deposited weld metal in which there appears to be absolutely no relation between the different mechanical properties, so that they cannot be deduced from the value of the tensile strength.

If a high malleability or a definite toughness is desired, the desired value of that property itself must be stated in the specification.

With respect to the workability of ordinary lightly alloyed steels, it depends in the first place on the hardness and in the second place on the chemical composition and the metallographic structure. It is obvious that the workability decreases with increasing hardness, but for hardnesses with which we are concerned (about 150) the decrease in cutting speed is of no significance.

This follows from *fig. 2* [2]). The limit of the filing

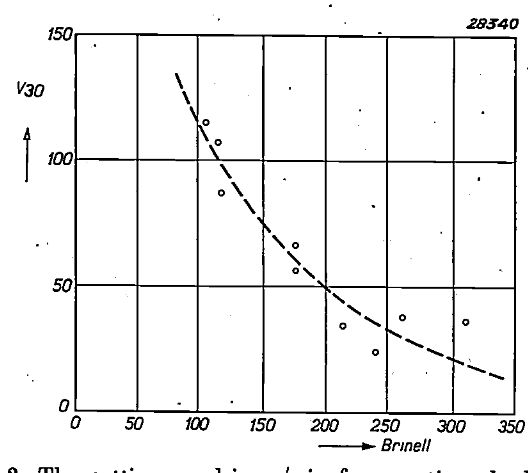


Fig. 2. The cutting speed in m/min. for a cutting edge life of 30 min (v_{30}) as a function of the hardness in Brinell units with a ball diameter of 10 mm, a load of 3000 kg and a loading time of 30 sec.

[2]) J. R. J. van Dongen and J. G. C. Stegwee, Metaalbewerking, 3, 49, 1936; *cf.* also P. Clausing, Philips techn. Rev. 1, 183, 1936.

quality of the metal is so high that the differences in hardness with which we are here concerned may be neglected.

We may state in conclusion that great care must be taken in applying information obtained from experiments on rolled carbon steel to deposited weld metal. The stipulation of a low upper limit of tensile strength of the deposited weld metal can at the present time only result in the use of welding rods of poorer quality than would be used if this obsolete requirement had been omitted from the specifications.

Nature of the material deposited

We shall now study the actual differences between the metal deposited in a weld and rolled metal, and shall consider particularly the hardness. The metal under consideration consists of ferrite or α-iron and cementite or iron carbide. The hardness of the metal depends upon that of its components. Now the hardness of ferrite is very much influenced by the impurities which are dissolved in the α-iron.

Silicon and phosphorus dissolve completely in α-iron. The solubility of gases like nitrogen and oxygen is not very great but it nevertheless plays an important part. Solid solutions in general exhibit a greater hardness than pure metals, but the data on this subject are scanty. For every one-hundredth of a per cent of nitrogen the hardness increases by $2^1/_2$ to $4^1/_2$ Brinell units for example. In general the nitrogen content is lower than 0.07 per cent.

The problem in the case of oxygen is not so simple. Its solubility in α-iron is very small. The content of oxygen in the deposited metal may sometimes be very high, 0.1 per cent for instance. Oxygen combined with silicon or aluminium in the form of SiO_2 and Al_2O_3 has no influence of the mechanical properties. Oxygen dissolved in α-iron, however, increases its hardness, while the rest of the oxygen combines with the iron and manganese forming FeO and MnO whereby the hardness of the metal is decreased.

As may be seen from the following list, that deposited weld material contains more nitrogen than rolled metal:

Siemens-Martin steel: 0.001 - 0.01%,
Thomas steel : 0.01 - 0.03%,
Deposited weld metal : 0.01 - 0.07%.

This is one of the causes of its hardness.

A second cause is the fine-grained structure of the metal deposited in a weld consisting of several layers. From investigations of rolled steels it follows that a fine-grained structure produces a greater hardness: this may also be seen from the behaviour of pearlite in this respect. According to Belajev the hardness is 200 Brinell units for a mutual separation of the lamellae of 0.4 μ, while it increases to 300 at 0.26 μ.

Furthermore a slight hardening of the material deposited occurs during welding, and can be reversed by annealing the welded piece of work (*cf. table IV a* and *b*). This hardening may be caused by an incipient precipitation (so-called pre-precipitation) of compounds of carbon, nitrogen and oxygen.

It is also possible that martensite is formed. There may also be other unknown causes for the hardening of the deposited metal.

Table IV

Influence of annealing and hardening on the mechanical properties of the deposited metal; a) absolute values, b) relative values

Type of welding rod	Heat treatment	Tensile strength (kg/sq.mm)		Yield (kg/sq.mm)		Elongation (%) $l = 5d$		Decrease in cross section (%)	
		a	b	a	b	a	b	a	b
PH 50	normally annealed	44.6	100	33.5	100	32.5	100	66	100
PH 50	untreated	48.7	109	39.5	118	29	89	55	83
PH 50	hardened *)	62.0	134	45.0	134	23	71	45	68
PH 48	normally annealed	47.4	100	34.8	100	32	100	68	100
PH 48	untreated	50.5	107	37.8	109	25	79	43	64
PH 48	hardened *)	63.0	133	41.4	119	19	60	36	53

*) annealed at 920 °C and quenched in water.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS GLOEILAMPENFABRIEKEN

1300: J. D. Fast: Spanlose Formung von Zirkon und Titan (Metallwirtschaft **17**, 459 - 462, Apr. 1938).

This article discusses the working of rods of ductile zirconium and titanium to sheets and wires of only 30 μ thickness. Various properties of zirconium are determined, among others the tensile strength and stretch. The measurements were carried out on cold worked sheets and wires, as well as on material which had been annealed at various temperatures after the cold working. In connection with X-ray determinations a survey is obtained of the changes in structure which occur upon annealing.

1301: M. J. O. Strutt and K. S. Knol: On the absolute measurement of alternating currents and the calibration of thermocouples in the decimetre-wave range up to 500 megacycles per second (Physica **5**, 205 - 214, Apr. 1938).

With the help of a diode voltmeter which measures accurately to within about 1% at 60 megacycles per sec it is shown that thermocouples of different models indicate the absolute value of an alternating current with a maximum error of 2%, up to 60 megacycles/sec, when the calibration curve for direct current is used. With frequencies up to 500 megacycles/sec, for the calibration of thermocouples and the absolute measurement of alternating currents, a hot-wire air thermometer according to Scheibe is used. With these instruments and a suitable Lecher bridge circuit alternating currents of several mA may be measured accurately within 5% at 500 megacycles/sec and within 1% at 325 megacycles/sec. The errors in indication of specially constructed thermocouples also lie within these limits.

1302: F. M. Penning: The energy balance for an infinitely small electron current in a uniform electric field (Physica **5**, 286 - 297, Apr. 1938).

The portion of the energy taken up in a homogeneous field by an infinitely small electron current which is used for elastic energy losses, for excitation of higher electron and vibration states, for ionization and for the acceleration of the electrons, is calculated as a function of the quotient of field strength E and gas density for neon, argon, mixtures of neon and argon and for nitrogen and hydrogen.

1303: W. G. Burgers and J. J. A. Ploos van Amstel: Electron optical observation of metal surfaces III. Crystal growth and allotropic transition in zirconium (Physica **5**, 305 - 312, Apr. 1938).

By means of the electron microscope it was observed that upon passing the transition temperature to lower temperatures the cubic β-crystallites of zirconium pass over into a lamellar, hexagonal α-texture, whose lamellae are different for each region where a given β-crystallite was at first situated. This α-texture recalls that of martensite steels. Upon increasing the temperature the original β-crystallites are recovered, and they may again be transformed into practically the same α-lamellae and so on. From this behaviour it follows that we are concerned with a homogeneous change in form of coherent lattice regions, which by slip and stretch in definite crystallographic directions pass from the one state to the other and the reverse, so that the whole process resembles the manner in which martensite needles occur in steel. At temperatures above 1150° C the β-crystallites exhibit a crystal growth in which the velocity of growth of the individual β-crystallites may be influenced by causing a temporary transition into the α-modification.

1304: W. G. Burgers and J. J. A. Ploos van Amstel: Electron optical observation of metal surfaces IV. Appearance of "lines" of high emissivity on nickel-iron crystals (Physica **5**, 313 - 319, Apr. 1938).

A crystal of nickel-iron activated by strontium and barium exhibits under the electron microscope a "normal" emission pattern according to its crystal structure. If an additional amount of strontium and strontium oxide is now applied to the cathode while it is kept at about 900° C, the emission increases in intensity not only over the whole surface of the cathode, but in addition it becomes unusually intense at individual points which are arranged along parallel lines. The direction of these lines is connected with the orientation of the crystal lattice, because the lines usually run from one crystal boundary to the other. Furthermore when the excess activating substance is removed by heating for a short time at a higher temperature, and when activating atoms are then deposited again on the surface under the same conditions, the lines reappear with practically the same direction.

1305: J. H. de Boer and J. D. Fast: The diffusion of hydrogen through regenerated cellulose and some cellulose derivatives (Rec. Trav. chim. Pays Bas **57,** 317 - 332, April 1938).

In the diffusion of hydrogen through metals the molecules are dissociated into atoms. In other media, such as acetylcellulose for example, hydrogen dissolves in the form of molecules, so that a diffusion of molecules may be expected, *i.e.* a diffusion proportional to the gas pressure. On the basis of the structure of cellulose and cellulose derivatives it may be expected that diffusion will occur more easily through the cellulose derivatives than through cellulose itself. Nitro and acetylcellulose are practically impermeable to air although not to hydrogen. Regenerated cellulose is practically impermeable to either gas.

The dependence of the rate of diffusion of hydrogen on temperature was investigated. For the case of nitrocellulose the dependence on pressure was studied and a proportionality was found. The diffusion was found to be practically the same for nitrocellulose and technical celluloid. The increased permeability of acetylcellulose for hydrogen compared with regenerated cellulose is found not to be due to a smaller activation energy, but to a greater number of paths per sq.cm suitable for diffusion, which in turn is perhaps due to a smaller size of micelles. In the case of nitrocellulose a smaller activating energy plays a part.

1306: E. J. W. Verwey and J. H. de Boer: Dilatancy (Rec. trav. chim. Pays Bas **57,** 383 - 398, April 1938).

A plastic mass for instance of iron filings with a non-polar liquid can be made fluid with a trace of oleic acid. The liquid so obtained can then still be considerably thickened, *i.e.* the particles of iron may be packed much closer together. Certain other substances exhibit the same behaviour. The liquids thickened to the correct concentration exhibit remarkably well the phenomenon described by Osborne Reynolds of becoming solid under the influence of rapid and acute deformations, it is called "dilatancy".

From the investigations there seems to be a simple explanation of this strange phenomenon. The solidifying under the influence of rapid and energetic deformations takes place because of the fact that the repulsive forces between the particles, due for instance to the polar film of oleic acid, are overcome locally.

1307: J. F. Schouten: Properties of variable width sound film as an optical diffraction grating (Nature, London **141,** 914 - 915, May 1938).

The employment of sound film as an optical diffraction grating suggests the investigation of the diffraction patterns which are produced by sound film with a variable width of the sound track. These diffraction spectra are much more complicated than those produced by sound film with variable transmissibility of the track. At the same time, however, they make possible an immediate analysis of the sound into its different components. The action of ordinary diffraction gratings is based on change either of the amplitude or of the phase of the incident light. Sound film with variable transmissibility of the sound track belongs to the first type. Sound film with variable width of the sound track, however, is a third, new type of optical diffraction grating.

1308*: F. A. Heyn: Radio-activity induced by neutrons (Diss. Delft 1938; 96 p.).

Two types of apparatus are described for the production of neutrons by means of high voltages (up to 600 kilovolts). Many substances were bombarded with neutrons produced by these apparatus, and new radio-active nuclei were found as well as a new type of reaction. A survey is given of the nucleus reactions which may be obtained by bombarding different substances with neutrons.

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*) An adequate number of reprints for the purpose of distribution is not available of those publications marked with an asterisk. Reprints of other publications may be obtained on application to the Natuurkundig Laboratorium, N.V. Philips' Gloeilampenfabrieken, Eindhoven (Holland), Kastanjelaan.

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EDITED BY THE RESEARCH LABORATORY OF N.V. PHILIPS' GLOEILAMPENFABRIEKEN, EINDHOVEN, HOLLAND

TELEVISION WITH NIPKOW DISC AND INTERLACED SCANNING

by H. RINIA.

621.397.33

The previously described television system with a Nipkow disc for the transmission of films has now been reconstructed for interlaced scanning in order to improve the quality of the picture. A description is given of the arrangement for obtaining interlaced scanning. The synchronization required special care. In order to insure good interlacing of the reproduction, the picture synchronising signals are now derived from the Nipkow disc (instead of from the film motor). This necessitated a special synchronization of the Nipkow disc with the film *i.e.* with the mains. The method by which this problem which was solved, and which provides satisfactory damping of any fluctuations in the speed, is described in detail.

In the previous volume of this periodical[1]) a television system for film transmission was described in which a Nipkow disc was employed. In this system the picture was resolved into 405 lines which were successively scanned. Since then this apparatus has been reconstructed for interlaced scanning.

In interlaced scanning the picture, which is scanned in 1/25 sec., is again resolved into 405 lines.

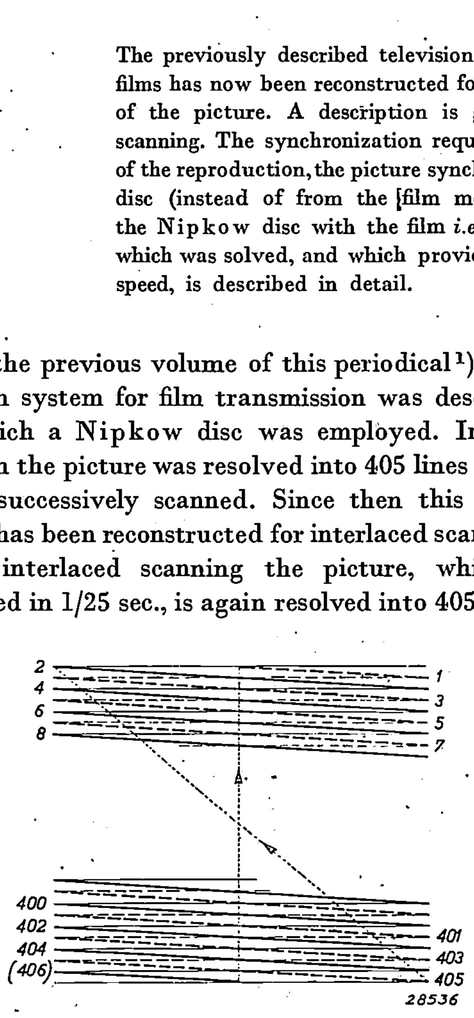


Fig. 1. Interlaced scanning with 405 lines per picture. The light spot first scans all the even-numbered lines, 2, 4, ... (heavy line). At the middle of the 406th line (*i.e.* the first line of the next picture) it flies back to the same height as at the beginning of scanning and now scans all the odd-numbered lines 1, 3, ... (dotted line), after which it again begins at the beginning of the second line. The representation here given is somewhat simplified. Actually the fly-back does not take place with infinite velocity, but it occupies a finite time. This has, however, no influence on the interlacing when the fly-back always occupies the same length of time.

1) H. Rinia and C. Dorsman, Philips techn. Rev. 2, 72, 1937.

However the scanning light spot first scans successively all the even-numbered lines and then, when it has reached the middle of line 406, (this being the 203rd line which it scans and actually the first line of the following picture) it flies back to the middle of the first line and proceeds to scan all the lines with an odd number (*fig. 1*). The advantage of this method lies in the fact that the picture reproduced flickers less: the effect is the same as if we had fifty complete pictures per second instead of twenty-five. The definition of the picture (fineness of the raster) is however the same as that of non-interlaced pictures with 405 lines per picture, so that the quality of the observed picture, in which both definition and freedom from flicker are important, is considerably improved.

Interlaced scanning

With interlaced scanning the film is moved continuously in the same way as was described previously for non-interlaced scanning, while the holes along the edge of the Nipkow disc move past the film in a direction perpendicular to the direction of motion of the film. Since the film advances a distance of one picture in 1/25 of a second, 405 holes must move past the film in this time. This is attained by providing the disc with 81 holes and causing it to make 5 revolutions in 1/25 sec, *i.e.* 7500 revolutions per minute.

In order to illustrate the way in which interlacing is obtained, *fig. 2* shows a length of film with three pictures. The scanning light spot L (*i.e.* the point at which the ray of light, coming through the hole which is just passing the film, touches the film) is in the upper left-hand side of the film picture I at a height a. It describes an almost horizontal line toward the right; as soon as it has reached the edge it jumps to the left again (actually a new ray of light appears coming from the next hole in the disc). However, because of the continuous vertical displacement of the film (in the direction of the arrow on the left), the light spot is now slightly lower than the beginning of the first line. The second line begins at this point, and so on. If now the light spot is given a vertical motion, by means of an arrangement to be described later, opposite and equal in magnitude to that of the film, the lines are twice as far apart: after the first line, the third one is scanned. After 1/50 of a second the light spot has reached the height b, the upper edge of picture II. 1/50 of a second later the whole of picture II has been scanned in $202^1/_2$ lines, and the light spot is at c, which point has also been reached by the upper edge of picture III, while the upper edge of II now lies at a. If at this moment the light spot jumps back to a, picture II is scanned once more. The fly-back of the light spot to the upper edge of the picture takes place after exactly $202^1/_2$ lines, *i.e.* at a moment when it is exactly in the middle of a line. The section of film has at that moment advanced only one half the distance between the lines since the beginning of the last line. Therefore the lines of the second scanning of picture II lie exactly symmetrically between the first $202^1/_2$ lines scanned. In this way interlaced scanning is achieved.

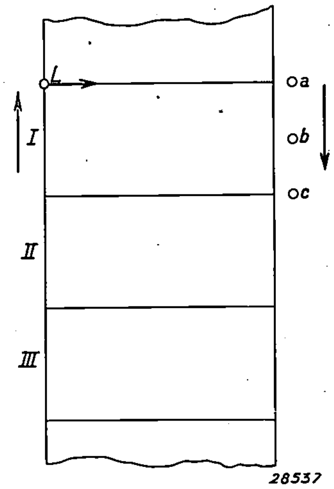


Fig. 2. Interlaced scanning of the film. The section of film moves in the direction of the left-hand arrow. The light spot L is projected through the holes in the Nipkow disc which move across the film picture. The light spot must be given a vertical motion equal and opposite to the motion of the film (arrow, right), and it must jump from the height c to a periodically.

The vertical motion of the light spot is obtained in the following way *(fig. 3)*. The light ray which passes through a hole in the Nipkow disc N passes through the objective O and falls upon a mirror S which reflects it back to the film F. The mirror can be rotated about an axis a, which is perpendicular to the plane of the drawing and lies in the plane of the mirror surface. A lever h is attached to the mirror and its extremity is pressed by a spring v against the cam disc K. The cam disc is driven by the film motor and makes one revolution in the time in which one picture passes, *i.e.* 1500 revolutions per minute. The circumference of the disc K has such a shape that the mirror, and therefore also the reflected light spot scanning the picture, is given exactly the motion described above.

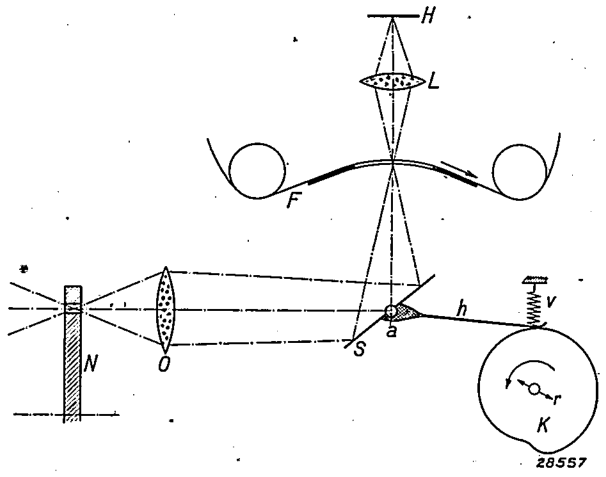


Fig. 3. Arrangement for obtaining interlaced scanning of films. N = Nipkow disc. O = lens which focusses the images of the holes in N sharply on the film F. The mirror S can be rotated about the axis a (perpendicular to the plane of the drawing) and is given a dipping motion by the lever h, which is pressed by the spring v on the circumference of the cam disc K. The axis of the cam K can be shifted slightly with a very fine setting screw in the direction of the arrows r. In this way the amplitude of the motion of the mirror may be accurately regulated. Lens L focusses the image of the opening of lens O on the photocathode H of the electron multiplier which amplifies the picture signals.

Fig. 4 shows the development of the disc, *i.e.* the radius as a function of the angle. The curve at the same time represents approximately the angle of rotation of the mirror and consequently the displacement in height *(fig. 2)* of the scanning spot as a function of the time [2]). It may be seen that the light spot is

[2]) This is not absolutely true because the point of contact between lever and cam disc experiences a slight sideward motion in the rising and falling of the lever; this has been taken into account in determining the shape of the cam disc. The cam therefore has not exactly the shape represented in fig. 4, but such a shape that the angle of rotation of the mirror as a function of the time varies according to fig. 4.

displaced uniformly during part I of one revolution of disc K, and during the very short part II it flies back to its initial position. This jumping back of the mirror must take place during the time of

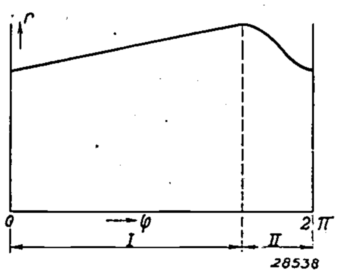


Fig. 4. Fundamental form of the development of the cam disc K (radius r as a function of the angle φ). The curve also gives the variation of the displacement of the light spot on the film with the time. In part I scanning takes place, in part II the back flash.

the picture synchronization signal which is transmitted between two successive pictures of the film (in this interval scanning would be useless in any case). The picture synchronization signal however occupies only 0.002 sec. In order to have the mirror jump back in such a short time without the appearance of any very large forces of inertia, the moment of inertia of the mirror, of its holder, and especially of the lever h, must be made very small. The mirror is no wider than is necessary for the required movement of the scanning spot, and the lever is made short and as light as possible for the necessary stiffness. The length of the lever cannot be made indefinitely short since the tolerances of the cam disc and its bearings would otherwise become too small. In our case the lever is 32 mm long; the tolerances are $\pm 3\mu$ and the acceleration at the end of the lever is 300 times the acceleration due to gravity. The mass of lever and mirror together acting at the end of the lever is only 0.5 g. The shape of the cam disc in part II of the curve in fig. 4 is such that the lever is uniformly accelerated during the first half of this period and then slowed down in the same way. This may be seen in *fig. 5*. (This line is actually the second derivative of the curve in fig. 4.) In this way shocks are avoided and the mechanism runs fairly noiselessly.

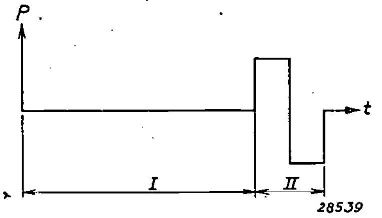


Fig. 5. Variation of the force of inertia P on the extremity of the lever h during one revolution of the cam disc K (this line is approximately the second derivative of the curve of fig. 4).

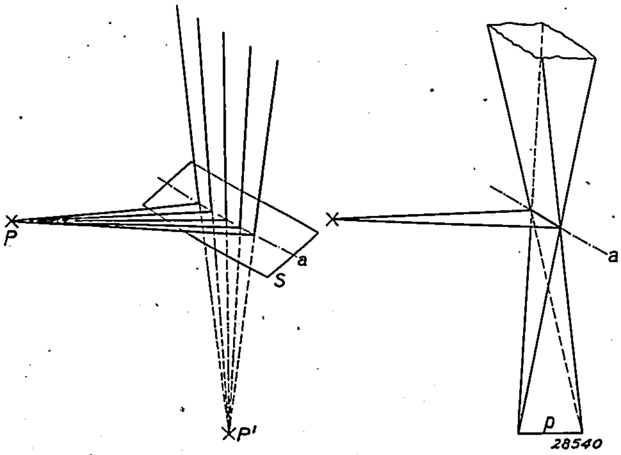
The spring v must of course be strong enough to prevent the lever from leaving the cam disc at any point [3]).

Image on the photocathode

The path of the film F at the point where the pictures are scanned is part of the surface of a cylinder whose axis coincides with the axis about which the mirror turns. Because of this the image of the light spot on the film always remains sharp when the mirror turns. The light transmitted through the film is collected by lens L and thrown upon the photocathode H of the electron multiplier which amplifies the picture signals. It is essential that the image of the light spot shall always fall upon the same points of the photocathode, so that, so to say, the photocathode remains unaware of the fact that the scanning spot comes continually from different directions. The sensitivity of a photocathode is never constant over its entire surface, so that if the light spot were displaced on the photocathode during scanning, the proportionality factor between film transmissibility and excited signal current would not be the same for all points of the picture. Such a displacement can be prevented by focussing with lens L an image on the photocathode of the opening of lens O, where the distribution of brightness is independent of the direction of the rays, so that the image of a point P of this opening always falls upon the same point of the photocathode. In previous arrangements without the rotating mirror in the path of the rays, this could be achieved directly. The fan-shaped beam of rays lying in one plane, which has its origin at P (the fact that the rays appear successively instead of simultaneously is of no importance here), may be concentrated by an ordinary spherical lens to a point again. By the action of the dipping mirror, however, the beam of rays takes on another form. This is represented in *fig. 6*, in which for the sake of simplicity the centre point of the opening of lens O is chosen for the point P. In fig. 6a a definite position of the mirror is assumed; due to the reflection the horizontal "fan" is transformed into a vertical "fan" with the image P' of point P as origin. Point P' might again be projected upon the photocathode with an ordinary lens. Due to the dipping motion of the mirror, however, about the

[3]) By giving the cam disc a somewhat more complicated shape it is also possible to scan the film pictures alternately twice and three times. With 24 pictures per second one then obtains exactly 60 scannings per second. This is desirable for countries like the United States where the mains voltage has a frequency of 60 c/s (instead of 50 c/s as in this country). Due to the synchronism of scanning and mains voltage many disturbing phenomena are avoided.

axis a the vertical "fan" as a whole is given a rotating motion about the same axis. The origin P' describes approximatelly a line p which lies in a plane perpendicular to a. Thus the beam of rays finally assumes the form shown in fig. 6b: it narrows twice to a line, in a and p; the two lines are perpendicular to each other. For every other point of the opening of O a similar figure is obtained which differs only slightly from that for the centre P.



a. b

Fig. 6. The fan-shaped beam of rays which originates from a point P of the opening of the lens O (see fig. 3) assumes a different form due to the dipping mirror S.
a) At a definite position of the mirror the horizontal fan is transformed into a vertical one.
b) By the motion of the mirror the vertical fan is turned back and forth around the axis a, the image P' describes a line p. The resulting beam narrows twice, at a and p, to a line.

It is known from elementary optics that the image of a point through an astigmatic lens is not a single point, but two mutually perpendicular lines lying one behind the other. Since in every optical arrangement the path of the rays may be reversed, it is clear that a beam of rays which has as origin two mutually perpendicular lines one behind the other, *i.e.* the form of fig. 6b, can be focussed to a definite point by means of a suitable astigmatic lens. This fact has been applied in the case under consideration. The astigmatism consists of the different curvature of the surface of a lens in two mutually perpendicular directions. Such a lens is obtained here by combination of a spherical with a cylindrical lens. One surface of both lenses is plane. By cementing the two plane surfaces together loss of light due to reflection is avoided as far as possible.

Synchronization

In the transition from ordinary scanning to interlaced scanning especial attention had to be paid to synchronization. We must distinguish here between the interlacing of the scanning and that of the reproduction.

Interlacing in scanning

In order to obtain satisfactory interlacing in scanning, *i.e.* in order to make the set of lines of the second scanning of a picture always fall symmetrically between the lines of the first scanning, the first requirement is that the dipping of the mirror shall take place synchronously with the changing of the film pictures. This is achieved mechanically due to the fact that the film and the cam disc are driven by the same motor (a synchronous motor of 3 000 revolutions per minute connected to the mains). When the adjustment is once correctly made the jumping back of the mirror always occurs at the right moment. In addition, however, the amplitude of the motion of the scanning spot, *i.e.* the distance a-c in fig. 2, must be exactly equal to the distance between corresponding points of two successive film pictures (the height of a picture plus the distance between two successive pictures). If this were not so, then, according to the explanation given with fig. 2, a different part of the picture would be scanned in the first scanning than in the second scanning, while these different parts would fall upon the same places in the picture reproduced. In order to be able to fulfil the above condition in every case, even when the height of the film pictures has been somewhat altered by shrinkage or stretching of the film as a whole, the axis of the cam disc (see fig. 3) can be shifted in position. By means of a fine set screw its distance from the axis of the mirror can be varied slightly. The effective length of the lever h is hereby changed and it is therefore possible to regulate accurately the angle through which the mirror turns during one revolution of disc K.

Interlacing in reproduction; picture synchronization signals

For satisfactory interlacing of the reproduction it is necessary that the times of the picture synchronization signals be very accurately determined with respect to the line synchronization signals: the light spot on the fluorescent screen of the receiver must at the moment of fly-back be alternately exactly at the end and exactly in the middle of a line (see fig. 1). In order to obtain some idea of the requirements to be met, it must be remembered that the scanning of one line lasts about 1/10 000 of a second. The time of the picture synchronization signal may only exhibit fluctuations which are small with respect to this time interval. A discrepancy of 1/10 of the length of a line, *i.e.* 1/100 000 of an second is already observable in reproduction as an asymmetry in the position of the two sets of lines.

Since the line synchronization signals are obtained through the holes in the Nipkow disc, as previously described [1]), it seemed natural to insure the correct mutual phase relation of line and picture synchronization signals in a simple way by deriving the picture synchronization signals also from the rotating disc. The following arrangement serves to do this (see *fig.* 7). A disc B_1 is attached to the axis

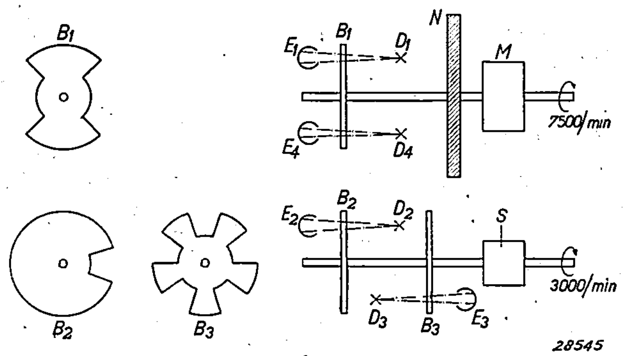


Fig. 7. Arrangement for obtaining picture synchronization signals. M = collector motor which drives the Nipkow disc, S = synchronous motor connected to the mains. B_1, B_2 = sector discs, D_1, D_2 = lamps, E_1, E_2 = photocells. The sector disc B_3 and the photocells E_3, E_4 with their lamps serve for synchronization of the Nipkow disc with the mains, as described on page 294.

of the Nipkow disc. Two sectors, each of 90°, and diametrically opposite each other are cut out of the disc B_1. On opposite sides of B_1 are placed a lamp D_1 and a photocell E_1. The light from the lamp is interrupted twice during one revolution of the disc, *i.e.* 250 times per second; the light flux I_1 on the photocell varies as shown in *fig. 8a*. In series with E_1 is a second photocell E_2, upon which the

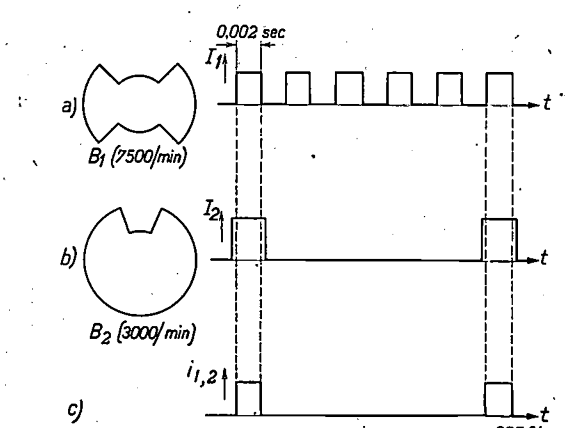


Fig. 8. Production of the picture synchronization signals. The light fluxes I_1 and I_2 on the photocells E_1 and E_2 in series with each other (see fig. 7) are interrupted by the discs B_1 and B_2 in the rhythm shown in the drawing. The light flux I_1 consists of 220 impulses per second (a), the light flux I_2 of 50 (b). The resultant photocurrent $i_{1,2}$ consists of 50 impulses per second (c) each lasting 0.002 of a second. These impulses are used as picture synchronization signals.

light from lamp D_2 falls. The light from D_2 is interrupted periodically by a disc B_2. This disc is driven by a small synchronous 3 000 r.p.m. motor S connected to the mains. Disc B_2 has only one opening of such a width that the transmitted light impulses last somewhat longer than the light impulses from disc B_1. The light flux I_2 on photocell E_2 varies as shown in fig. 8*b*. In the electrical circuit which includes the two photocells in series, a current can however flow only at moments when both photocells are illuminated, which is the case at every fifth impulse of the light flux on E_1 (fig. 8*c*). In this way 50 impulses per second are obtained which, due to the unchanging relative positions of disc N and B_1, always have the same phase relation relative to the line synchronization signals. These impulses are used as picture synchronization signals.

In the earlier arrangement [1]) the picture synchronization signals were obtained by a rotating sector on the axis of the film motor which interrupted the light falling on a photocell in the desired rhythm. This arrangement is now used to suppress the picture signal during the time when the picture synchronization signal is being transmitted. The impulses of the photocurrent, which are somewhat wider than the picture synchronization signals, are amplified and fed to one of the plates of the electron multiplier which amplifies the picture signal. By such a large voltage impulse on one of the plates, the secondary emission of that plate is interrupted since electrons no longer fall upon it; the amplification of the whole mutiplier thereby immediately becomes zero [4]). The picture signal is therefore suppressed during the voltage impulse. and the magnitude of the picture synchronization signal is consequently constant.

Synchronization of the Nipkow disc.

Because of the fact that the picture synchronization signals are now no longer derived from the film motor as in the previous arrangement, but from the Nipkow disc, a special synchronization became necessary between the film and the Nipkow disc. In order to attain correspondence between the picture scanned and the picture reproduced, the picture synchronization signals must be synchronous with the changing of the film pictures. A deviation from the correct ratio of revolutions of film motor to those of the Nipkow disc

[4]) One particularly satisfactory characteristic of the electron multiplier is the fact that, due the absence of a current when the multiplier is not illuminated, no extra voltage impulse occurs (in contrast to the case of an ordinary amplifier valve).

(actually disc B_1, see fig. 4), would lead to the reproduced picture being not stationary but moving over the screen. The Nipkow disc must therefore be synchronized with the film motor and in such a way that no appreciable fluctuations occur about an average value of the speed, in order the prevent a disturbing rise and fall of the picture [5]).

The problem of keeping the number of revolutions of the Nipkow disc constant, at 7 500 revolutions per minute as mentioned above, is made more difficult by the fact that due to the large moment of inertia of the disc, the characteristic frequency at which the system can oscillate becomes low (of the order of 1 c/s). The oscillations occurring cannot therefore be satisfactorily suppressed by an eddy current damping in the driving motor, but would have to be avoided by the use of a sufficiently powerful motor [6]). Apart from the fact that the use of a very heavy motor would not be exactly an elegant solution of the problem, the synchronizing of the asynchronously starting motor is also difficult.

A very satisfactory solution was obtained in the following way. The axis of the Nipkow disc N is driven by a commutator motor. On opposite sides of the sector disc B_1 (fig. 7), which serves for the excitation of the picture synchronization signals, there is a second photocell E_4 and a lamp D_4. The light flux I_4 on E_4 again has the form of fig. 8*a* with 250 impulses per second, or 15 000 per minute (see *fig. 9a*). On the axle of the synchronous motor S (see fig. 7), besides disc B_2, there is also a second disc B_3, out of which 5 sectors have been cut, so that the light falling on the photocell E_3 is interrupted in the same rhythm (15 000 times per minute) as the light on photocell E_4. The light flux I_4 on photocell E_4 varies as shown in fig. 9*b*. Photocells E_3 and E_4 are again connected in series, a photocurrent flows only when both of them are illuminated at the same time. A pulsating-direct current $i_{3,4}$ occurs (fig. 9*c*), whose average strength obviously depends upon the phase relation between discs B_1 and B_3. By means of the current $i_{3,4}$ the excitation of the commutator motor M which drives the Nipkow-disc, is regulated as follows (see *fig. 10*). The pulsating direct current, which is

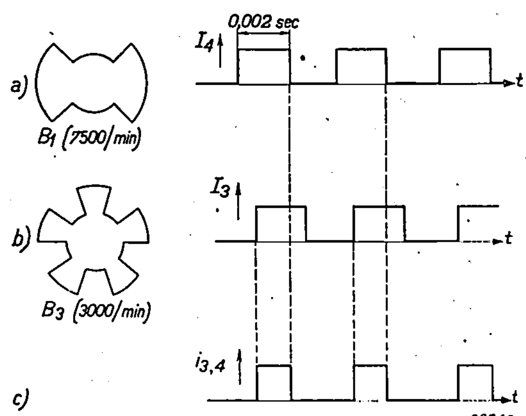


Fig. 9. a) Variation with time of the light flux I_4, falling on the photocell E_4 and periodically interrupted by the disc B_1.
b) Variation with time of the flux I_3, falling on the photocell E_3 and periodically interrupted by the disc B_3.
c) Variation with time of the resultant photocurrent $i_{3,4}$ at a definite relative phase displacement of the discs B_1 and B_3.

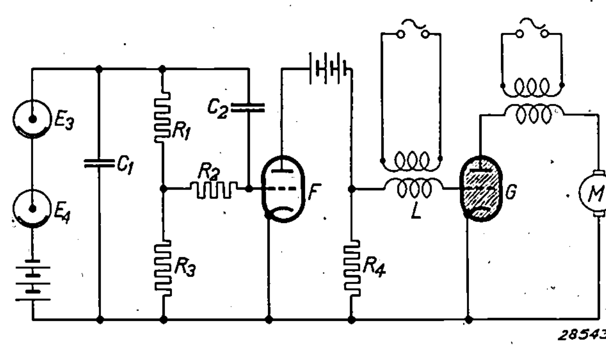


Fig. 10. Diagram showing the principle of the circuit for synchronization of the Nipkow disc. E_3 and E_4 = photocells, C_1 = smoothing condenser for the photocurrent $i_{3,4}$. The voltage on R_3 is amplified in the valve F. The voltage on R_4 is fed to the grid of the relay valve G which supplies the excitation current for the motor M which drives the Nipkow disc. The transformer L supplies a sinusoidal grid voltage to G which is 90° in phase behind the anode voltage. Due to the resistance R_2 and the condenser C_2 the regulatory photocurrent obtains the necessary forward phase shift with respect to the deviation occurring from the correct number of revolutions of the disc at that frequency at which the system can oscillate.

smoothed by the condenser C_1, causes a fall in potential along the resistances R_1 and R_3. The fall in potential along R_3 acts, over R_2, between grid and cathode of the triode F. The anode current of F causes a voltage drop along the resistance R_4. This voltage influences the moment of ignition of a relay valve G which supplies the excitation current of the motor M (actually three relay valves are connected in parallel for the three phases of the supply current). An alternating voltage V_g is supplied to the grid of G by the transformer L, which voltage is 90° in phase behind the anode alternating voltage V_a (*fig. 11*). The anode current flows only from the moment t_1 when the grid voltage becomes positive until moment t_2 when the

[5]) A slight rise and fall of the picture, as sometimes also occurs in the picture in a cinema, is not disturbing.

[6]) In any case it would not be possible to use an ordinary synchronous motor connected with the mains, since such a motor makes 3000 r.p.m at most. A synchronized converter from 50 to 125 c/s would have to be employed.

anode voltage V_a becomes negative. If now a direct voltage is superposed on the grid voltage V_g, the curve V_g in fig. 11 is shifted somewhat in height

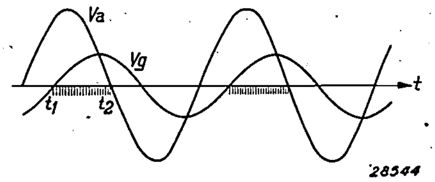


Fig. 11. The relay valve G (fig. 10) ignites as soon as the grid voltage V_g becomes positive (time t_2). The valve works only during the time intervals indicated by cross hatching. If a direct voltage is superposed on V_g, *i.e.* if the curve V_g is raised somewhat in height the moments t_1 occur somewhat sooner, the average anode current of the relay valve becomes greater.

and the moment t_1 occurs somewhat sooner. The anode current impulses from G become longer, *i.e.* the average anode current which excites the motor M becomes greater, the motor runs faster. Therefore the speed of the motor is influenced by the magnitude of the photocurrent $i_{3,4}$, *i.e.* by the phase relation of the discs B_1 and B_3 in fig. 7 and 9. At a definitive position of the discs B_1 and B_3 the beginning of a relative speeding up of the Nipkow disc will result in a decrease in the photocurrent $i_{3,4}$ and therefore also in a retarding influence on the disc. Nevertheless this system would not be sufficient without additional precautions not only to obtain synchronization but also to prevent oscillations. The synchronizing force is in phase with the deviation which occurs in the relative position of the discs B_1 and B_3, and by which, as in the case of the pendulum where the force in the backward direction is in phase with the displacement from the mean relative position of the discs, *i.e.* synchronism of motion, (the position of rest of the pendulum) is established, but no damping is obtained. In the case of the pendulum for damping a frictional force is necessary which is always a quarter of a period in phase ahead of the backward-acting force of gravitation. By analogy with this in our synchronization arrangement also the synchronizing force must be a certain amount ahead in phase of the deviation from the correct relative position. This is realized in a simple way by means of the condenser C_2 in fig. 10, which is of such a size that the voltage variations fed to the grid of F are ahead in phase by a certain amount with respect to the variations of the photocurrent $i_{3,4}$ at that frequency at which the Nipkow disc can oscillate.

Excellent damping and synchronization is actually obtained in this way. With any slight changes in the mains frequency, the Nipkow disc, due to the action of the synchronous motor S, continues to move synchronously with the film which is in turn moved synchronously with the mains.

THE "ROTALIX" TUBE FOR X-RAY DIAGNOSIS

by J. H. VAN DER TUUK.

621.386.1 : 616-073

In X-ray tubes for medical diagnosis in which moving organs (heart, lungs, stomach) are to be photographed, a very high specific focus loading of very short duration is essential. In contrast to the case of the previously described tube for the investigation of crystal structure, an intensive cooling with running water would be quite inadequate here. The specific loading can however be considerably increased by allowing the anode to rotate during the exposure. The improvement obtained by this method and the manner in which this idea is realized in the "Rotalix" tubes are discussed in this article.

Sharpness of the X-ray picture

As in ordinary photography, sharpness is also essential in X-ray photography. In medical X-ray examination the primary problem is to obtain sharp pictures of certain parts of the body or organs. What are the requirements which must be made of an X-ray tube for this purpose, and how is the desired tube to be realized technically? The problems here are somewhat different from those in the case of the recently described X-ray tube for crystal structure analysis.

It is clear that the size of the focus of the X-ray tube plays an important part. Let us assume for the sake of simplicity that the focus is at $d = 1$ metre distance from the object to be photographed; the distance object-to-film may be $e = 10$ cm (*fig. 1*). The lack of definition with which a point

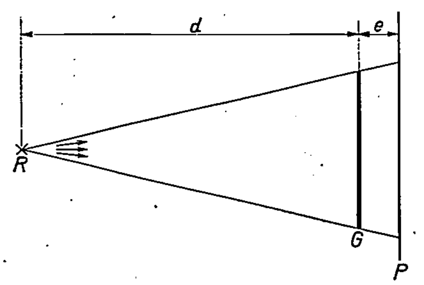


Fig. 1. Arrangement in taking an X-ray photograph. The object G is at a distance d from the square focus R. The distance between object and film P is e. With an area F of the focus every point of the object is projected as a shadow spot of width $e\sqrt{F}/d$ (geometrical lack of sharpness O_g).

of the object is projected as a shadow image on the film is then 10 per cent of the dimensions of the focus; if the focus is 3 mm wide a point is projected as a spot 0.3 mm in diameter. From this one might conclude that the X-ray tube with the smallest focus gives the sharpest picture (at definite values of d and e; by increasing d the geometrical lack of definition is decreased in any case). This is indeed approximately true in the photography of simple fractures, and in general in the photography of stationary objects.

The situation, however, is quite different if the object to be photographed moves during the exposure, as in the case of heart, lung and stomach photographs. In addition to the geometrical lack of definition of the picture due to the finite dimensions of the focus, the so-called kinematic lack of definition plays a part. The most rapidly moving parts of the lung move for instance at a speed of 15 mm/see [1]). The lack of sharpness caused by this movement depends upon the exposure time, and may be very considerable. In order to limit the kinematic lack of sharpness the necessary exposure time must be made as short as possible, which means that at a definite permissible loading per sq.cm. of the focus the surface of the focus would have to be increased in order to obtain a greater intensity. Thus we make two contradictory requirements of the focus: it must be small in order to keep the geometric lack of sharpness (O_g) small, and it must be large in order to keep the kinematic lack of sharpness (O_b) within certain limits. The total lack of sharpness [2]) $O = O_g + O_b$ is as small as possible when:

$$O_g = 2\,O_b \quad \cdots\cdots \quad (1)$$

This may be explained as follows. The kinematic lack of sharpness O_b, with the very short exposure times t to be considered, is proportional to these times: $O_b = a.t.$ The geometrical lack of sharpness O_g is proportional to the linear dimensions of the focus, *i.e.* to the square root of the area F of the focus: $O_g = b\cdot\sqrt{F}$ is however proportional to the total intensity of the X-radiation, and therefore inversely proportional to the necessary exposure time t; therefore $O_g = c/\sqrt{t}.$ The total lack of sharpness becomes $O = O_b + O_g = a.t. + c/\sqrt{t}.$ O reaches a minimum for the case when $a - (1/2\,t)\cdot c/\sqrt{t} = 0$, *i.e.* when $c/\sqrt{t} = 2\cdot a\,t$, *i.e.* condition (1).

With a given exposure technique and a given speed of the object to be photographed, for a given

1) H. Chantraine, Fortschr. Röntgenstr. 9, 659, 1933.
2) The lack of sharpness of the picture due to the photographic material is here left out of consideration.
3) A. Bouwers, Acta Radiologica 12, 175, 1931.

distance between focus and object, the most favourable size of the focus of the X-ray tube to be used is determined by equation (1). (For more slowly moving objects a smaller focus is in every case still more satisfactory).

If the distance d between focus and object (fig. 1) is changed, then the optimum area for the surface of the focus is also changed. The attainable sharpness however remains the same. If we choose the dimensions of the focus twice as great, the total intensity will of course become 4 times as great; however, in order to obtain the same geometrical sharpness we must decrease the area of the focus by one half, and the intensity of the X-radiation becomes one fourth, so that we have gained nothing in the end.

Specific loading

The sharpness of the picture, as far as the X-ray tube is concerned, can only be improved according to the above by simultaneously diminishing the area of the focus and shortening the exposure time. This is possible if we cau increase the load per sq.cm. on the focus. This specific loading has however a natural limit due to the maximum permissible temperature of the focus. As in the previously described X-ray tube for structural analysis [4]), the problem is here encountered of how to make the specific loading as high as possible without overheating the focus. The solution reached in that case (cooling with running water) is not satisfactory in our case. The exposure times in structure analysis are always at least several minutes, while in the X-ray photography of moving organs we are often concerned with exposure times of only several hundredths of a second. The loading here has more the character of an "explosion", and it is technically impossible to transfer the heat developed during the required specific loadings of the focus to running water so rapidly.

It is however possible to obtain an increase in the specific loading by allowing the anode to rotate during the exposure time. In that case fresh, still cold parts of the anode pass successively under the beam of electrons, and each point is loaded for only a fraction of the exposure time. This idea is already quite old (Breton 1898—1899); it was however only realized in a practical form by Bouwers in 1929 [5]).

[4]) J. E. de Graaf and W. J. Oosterkamp Philips techn. Rev. 3, 263, 1938.

[5]) A. Bouwers, Kongressheft Fortschr. Röntgenstr. 20, 103, 1929.

Improvement attainable with rotating anode

We shall now examine what improvement of the specific loading can be theoretically and practically attained with a rotating anode. For times $t < 0.04$ sec (in the range of temperatures with which we are concerned) the rise in temperature $T - T_0$ at the surface of a loaded anode is given approximately by

$$T - T_0 = AW\sqrt{t} \quad \cdots \cdots \quad (2)$$

W is here the specific loading, and A is a constant depending upon the heat conductivity and specific heat of the anode material. Instead of allowing the anode to turn under the electron beam we may imagine for the sake of simplicity that the focus is displaced over a stationary anode, which amounts to the same thing. If during the exposure the focus of width f moves over a distance $m.f$ the time during which a given point of the anode is loaded becomes m times as small. According to (2) the increase in temperature at the surface then becomes $\sqrt{m}$ times as small, or, with a specific loading $\sqrt{m}$ times as great, the original increase in temperature is again obtained. This is illustrated in *fig. 2*. With a stationary anode the temperature at the focus varies during loading according to parabola *1* (see equation (2)). Line *2* represents the variation of the temperature with a rotating anode under the same load and with a speed of rotation such that the anode is displaced a distance of $16\,f$ during the exposure time. At a specific loading $\sqrt{16} = 4$ times as great (line *4*) the same maximum increase in temperature is then obtained as with the stationary anode. If in general the displacement of the anode during loading is p^2 times the width of the focus, the permissible loading is improved by a factor p. Now the following relation holds:

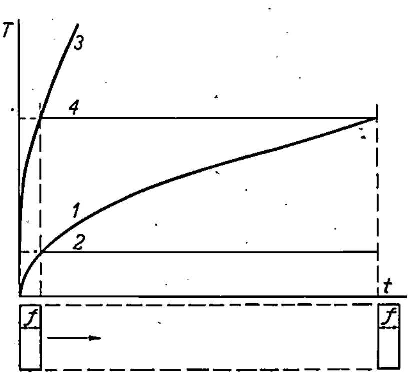


Fig. 2. Variation of the temperature T at the focus during loading. The initial temperature T_0 is set equal to zero for the sake of simplicity. Curves *1* and *3* with stationary anode, curves *2* and *4* with rotating anode. The specific loading in curves *3* and *4* is four times as great as in curves *1* and *2*.

$$p^2 f = 2\pi r n t,$$

where r is the distance from the middle of the focus to the axis of the anode, n the number of revolutions and t the loading time. The theoretically attainable improvement factor p of the specific loading upon transition from stationary to rotating anode therefore becomes:

$$p = \sqrt{2\pi r n t/f} \quad \cdots \cdots \quad (3)$$

From the derivation of this formula it may be seen that it is only valid for a limited range of loading times t. If t is so small that $mf \leqq f$, *i.e.* if the anode is displaced by less than one focus width during the loading, no improvement is obtained by the rotation of the anode: a strip of the focus which may be wider or narrower is in the latter case always loaded during the time t and the permissible loading is limited by this critical strip, which amounts to the same thing as if the whole focus were loaded during the whole time t (stationary anode). For shorter loading times than those at which $mf = 2\pi r n t = f$, *i.e.* for $t \leqq f/2\pi r n$, therefore, formula (3) is not valid, but p is a constant equal to unity.

On the other hand formula (3), as well as (2) from which it is derived, holds only for exposure times $t < 0.004$ sec. For longer times the rate of increase in temperature with a stationary anode is less than proportional to $\sqrt{t}$; therefore, under otherwise similar conditions, the improvement factor increases less and less rapidly with increasing exposure time, and would finally approach a definite maximum value. Furthermore, however, with longer exposure times, *i.e.* when $t > 1/n$, one spot on the rotating anode passes under the electron beam more than once during the whole loading. The second time that a spot is loaded it will have retained a certain temperature increase from the first time, and this will be even greater the following times; therefore the loading may not be increased by the theoretical factor p, but only by a smaller factor if the same maximum focus temperature is to be reached. There is therefore even a decrease in the improvement factor p with increasing exposure time t. For extremely long exposure times p is independent of the time t, it is then determined by the material, the form and cooling of the anode.

In the practical construction of "Rotalix" tubes, which we shall describe later, $r = 2$ cm, $n = 2\,900$ rev/min., $f = 1.3$ mm. From equation (3) it then follows that specific loading with the rotating anode for $t = 0.04$ sec could be more than 13 times as great as with a stationary anode. With $n \doteq 2\,900$ rev/min. during one exposure of 0.04 sec about two revolutions occur, so that the actual improvement factor is smaller (the temperature increase retained after one revolution is in the practical case about 60°).

Moreover in the case of the anode of the "Rotalix" tube for practical reasons (see page 301) one must calculate with a basic temperature of $Tg = 450$°C instead of the temperature assumed up to now of 0°C. This also makes the practically attainable improvement factor smaller. It amounts to about 8 or 9. In *fig. 3* the measured improvement factor

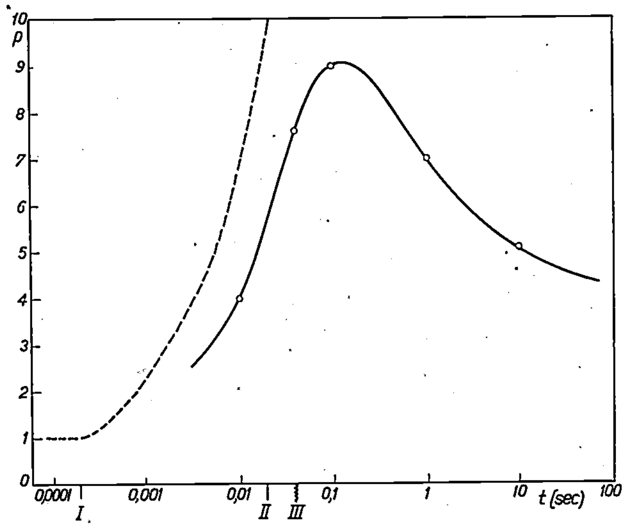


Fig. 3. Variation of the improvement factor p of the specific loading as a function of the exposure time t upon transition from stationary to rotating anode (t is plotted logarithmically). Width of focus $f = 1.3$ mm, radius of anode $r = 2$ cm, number of revolutions $n = 2\,900$ per min. At $t < 0.002$ sec (left of I) no improvement is obtained ($p = 1$). For $0.0002 < t < 0.02$ sec. p varies theoretically according to formula (3) (dotted line). Due to the higher basic temperature of the anode after continued use the practical value of p is less than the theoretical one (continuous line). At $t > 0.02$ sec (right of II) a single spot on the anode passes more than once under the electron beam. Finally at $t > 0.04$ sec (right of III) the focus temperature with stationary anode increases at a rate less than proportional to $\sqrt{t}$. For these two reasons the improvement factor falls with higher values of t, and approaches a value independent of t.

is plotted as a function of the time. The lower limit of validity of formula (3) lies in our case at $t = 0.0002$ sec. The theoretical variation of p according to (3) is indicated by a dotted line.

An attempt might be made to increase p still further by giving the anode a greater diameter ($2r$) and a higher speed of revolution (n). This is possible technically. Such an improvement however would at present still serve no useful purpose. With the increase of the specific focus loading it is only in the geometrical and kinematic lack of sharpness of the picture that a decrease is obtained. There is still a third source of lack of sharpness which is inherent in the photographic material used (namely the fluorescent intensifying screens). As long as this lack of sharpness remains unchanged it is useless to decrease the other contributions to the total lack of sharpness below a certain limit. This limit has been almost reached with the photographic material at present available.

In *fig. 4* the improvement obtained in the specific loading is illustrated in a different way: a focus of a tube with a stationary anode (*a*) and the equivalent focus with a rotating anode (*b*) are shown next to each other, both to be loaded with 21 kW during 0.1 sec.

Fig. 4. Focus (in projection, see below) of a tube with stationary (*a*) and with rotating (*b*) anode, both loaded with 21 kW during 0.1 sec.

Form of the focus

In the article quoted in footnote [4]) it has already been explained in detail that, due to the non-validity of Lambert's law in X-ray emission, the focus exhibits a considerably greater brightness in an oblique direction of observation than in a normal direction. Use has been made of this fact in "Rotalix" tubes also. These tubes have a so-called line focus with the dimensions $1.3 \cdot 5.2$ mm and the central ray of the effective cone makes an angle of 15° with the surface of the anode. In this projection the focus appears as a square $1.3 \cdot 1.3$ mm, and would then, as far as the total X-ray intensity in this direction is concerned, be equivalent to a circular focus observed normally of almost 3.0 mm diameter with a stationary anode. The much smaller apparent dimensions of the line focus are an important advantage from the point of view of sharpness of the picture. For a tube with a rotating anode, however, the line focus offers an additional advantage. According to equation (3) upon rotation the improvement factor p is proportional to $\sqrt{n/f}$, *i.e.* a given number of revolutions accomplishes, for the line focus of $f \doteq 1.3$ mm, as much as double the number of revolutions would accomplish for the circular focus of width (diameter) $f = 3.0$ mm.

Finally in *table I* the permissible loading is given of a "Rotalix" tube with a line focus of 1.3 mm width, for several loading times (with direct voltage).

Table I

Loading time in seconds	Permissible loading in kW
0.01	24.5
0.04	23
0.1	20.5
1.0	11
10.0	6

Construction of "Rotalix" tube

In *fig. 5* the construction of a "Rotalix" tube is given diagrammatically. For the general construction

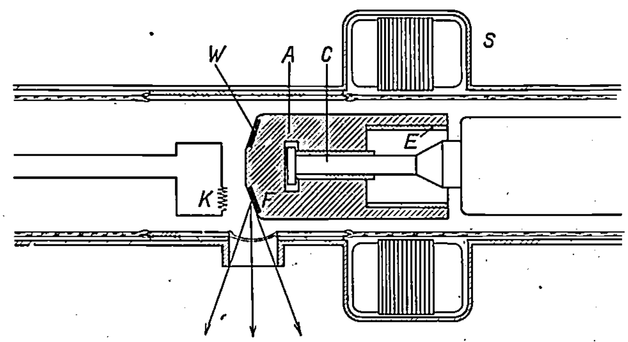


Fig. 5. Construction of a "Rotalix" tube (diagrammatic). The discharge space between the cathode K and the anode A is completely surrounded by metal. The cathode is mounted excentrically in the tube and constructed in such a way that a line focus F occurs along a line of revolution of the conical front surface of the anode, upon which a layer of tungsten W is applied. The anode turns about the axis C. It is driven in the manner of the armature of an induction motor by a polyphase current in the stator windings S. The iron field ring E causes as many lines of force as possible to pass through the copper of the anode.

(small discharge space surrounded by metal, window for the emission of the X-rays, screening from the undesired X-rays by lead, protection against high voltage by earthed metal jacket) we may refer to the X-ray tube previously described for the investigation of crystal structure [4]), which exhibits the same characteristics. The anode A can rotate about the axis C. The driving force is obtained as in an induction motor: the windings on the stator S are supplied with polyphase current, the copper anode itself forms the rotor turning in a vacuum, upon which a couple acts due to the induced eddy current. The iron field ring E serves to cause as many lines of force as possible to pass through the copper. For the same reason the distance between stator and anode is made as small as possible. Since the stator is earthed and the anode is under high tension, this distance is limited by the distance necessary to avoid break-down. The high tension is supplied to the anode *via* a brush contact.

The problem of lubrication of the bearings of the rotating anode in the evacuated tube is solved in the case of "Rotalix" tubes by the use of a special grease (a high distillation product of petroleum) in combination with ordinary ball bearings. The vapour pressure of this grease is only 10^{-6} mm at 150°C.

Special care is taken that the bearings, which must be well outgassed before mounting, do not become too warm during running, so that the vapour pressure of the lubricant remains low during use. Moreover, in the "Rotalix" tube as well as in all "Metalix" tubes a getter is employed [6]), which is introduced in such a position that it can easily be reached by gases and vapour from the grease which are freed in the bearings. Thanks to these precautions the vacuum in the tube is not impaired by the presence of the lubricant.

Construction of the anode

For several reasons the most suitable material for the anode of an X-ray tube is tungsten. This is especially true of tubes for diagnosis which are particularly constructed for X-ray photography of moving objects, in which, as we have seen, high permissible loading during a short time is essential. The intensity of the X-radiation is proportional to the atomic number which is very high in the case of tungsten, namely 74. The melting point of tungsten is higher than that of any other suitable metal; (approximately 3 400°C). The vapour pressure is small so that the material may be heated to more than 2 500°C for a short time without causing the vapour pressure in the X-ray tube to rise unduly. The heat conductivity is 40 per cent of that of copper, and therefore quite satisfactory. Considering that for the parts of the anode lying behind the focus the high melting point, the high emission of X-radiation, etc. of tungsten are no longer important factors, and, that there only is the best possible heat conduction to the back (perpendicular to the surface of the focus) desired, copper can be used to advantage for these parts. The anode consists of a massive copper block upon which the focus is applied in the form of a thin layer of tungsten. The surface of the focus reaches temperatures of about 2 500°C during the very brief loading times; the temperature of the boundary surface tungsten-copper may however not exceed 1 000°C, since the melting point of copper is about 1 080°C. At a given specific loading therefore the optimum thickness of the tungsten layer depends upon the loading time. With loading times of the order of 1 sec the most suitable thickness is about 1.5 mm, for shorter times it is considerably smaller, for instance only about 0.3 mm with a loading time of 0.04 sec.

It is obvious that the anode material is put under severe strain during loading. In the tungsten layer within a distance of 0.3 mm between front and back there is a temperature difference of about 1 500°C, and the front surface reaches temperatures of more than 250°C. In addition there is the alternate heating and cooling of the material with resulting buckling and shrinking. A tungsten layer in the form of a plate proves unable to withstand all this for a long time. Therefore in "Rotalix" tubes the tungsten layer on the anode is composed of spirally-wound tungsten ribbon (with the axis of the anode as axis of winding) 0.12 mm. thick and 0.3 mm wide. The layer is then 0.3 mm thick. The copper of the anode must also be reinforced to prevent the occurrence of cracks. This is done by means of so-called tungsten wool consisting of a bunch of tungsten wires (each wire 0.1 mm thick, 5 wires per sq.mm.), with which the individual crystals of copper are anchored, as it were, during the casting of the copper block of the anode.

Cooling of the anode

The increase in temperature of the massive anode block is fairly slight during one exposure. It is therefore possible to make a number of successive photographs without the anode as a whole becoming very much warmer. In the early models of "Rotalix" tubes therefore no special provision was made for cooling the anode. It then took some time before a heated anode was again quite cold, since convection and heat conduction have no effect in the evacuated tube, and the anode could lose accumulated heat only by radiation. The use to which the tubes were put made a change necessary here. Contrary to the original intention the "Rotalix" tube was used not only for taking photographs of moving organs, but also of practically stationary objects; at present it is even used for fluoroscope examinations (in this case the anode need not even turn, since it can take up such small loads when stationary). In order to make the tube also suitable for such continuous use the anode had to be cooled.

Fig. 6 gives a diagrammatic representation of the anode of a "Rotalix" tube with a cooling system [7]). As already mentioned, dissipation of heat can only be obtained by radiation. Therefore

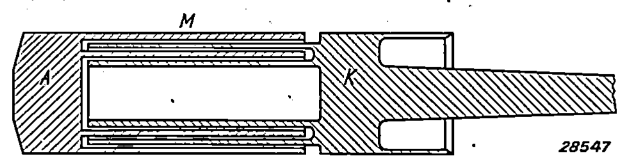


Fig. 6. The cooling of the anode A rotating in a vacuum, which is necessary for continuous use of the tube, takes place by radiation of the blackened cylindrical shells M upon similar shells of the stationary support K.

[6]) J. H. van der Tuuk, Ned. T. Natur. 3, 129, 1936.

[7]) A. Bouwers, Fortschr. Röntgenstr. 48, 232, 1933.

the radiation is made as great as possible. Blackened cylindrical shells M projecting from the rotating anode A radiate the heat on similarly blackened shells of the stationary support K. The amount of heat transferred per second then amounts to the following, according to the Stefan-Boltzmann law

$$Q = c \cdot O \cdot \sigma \, (T_A^4 - T_K^4) \quad \cdot \cdot \cdot \cdot \quad (4)$$

where c is the heat transfer coefficient between the shells, O the total surface of the shells M, σ the radiation constant ($5.3 \cdot 10^{-12}$ watt cm^{-2} degree^{-4}), and T_A and T_K are the absolute temperatures of the anode and support respectively. The coefficient c for cylindrical surfaces is approximately $z/(2 - z)$ where z is the blackening coefficient. The surface of the shells M is for instance 400 sq.cm, and at an easily attainable value of the blackening cofficient

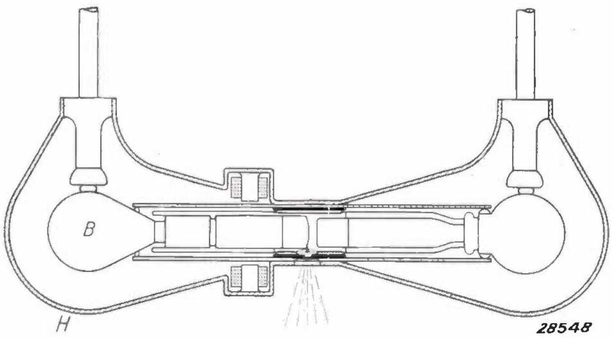


Fig. 7. For the cooling of the support K in fig. 6 a cooling bulb B is mounted on it, which gives off its heat by convection to the surrounding air and by radiation to the earthed container H of the tube.

z of 0.7 the coefficient of heat transfer $c = 0.54$. It can be estimated that the anode as a whole reaches a basic temperature of about 450°C, while the support K which can give off its heat to the air around the tube reaches a temperature of about 200° with intensive use of the tube [8]). With these values it can be calculated from (4) that an amount of energy of more than 200 W can be dissipated continuously through the anode, this is a load which is very seldom necessary for several consecutive hours in X-ray diagnosis. The support which receives the heat by radiation from the anode must in its turn get rid of it, and the necessary electrical insulation forms an obstacle to this process. The simplest solution is sketched in *fig. 7*. A blackened cooling bulb B is mounted on the support, and by convection through the air, and partly also by radiation, it gives off its heat to the earthed container of the whole tube.

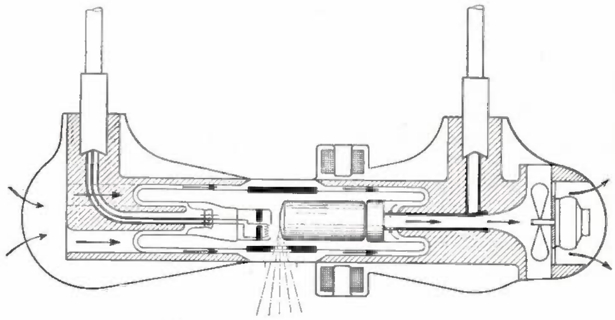


Fig. 8. The support in fig. 6 and the whole tube may be cooled by means of a fan by air circulation. This makes it possible, especially when the insulation between tube and earthed container is obtained with the help of high-tension "Philite" (instead of air), to make the dimensions of the tube considerably smaller.

Because of the necessary large surface of the cooling bulb the tube then becomes fairly large. This can be avoided by cooling the support and tube directly by air circulation, by means of a fan. The reduction in the dimensions of the tube is then limited only by the necessary insulation distance between the tube proper and its container. If this space is filled with high tension "Philite", which has a much higher break-down potential than air, the whole tube can be still further reduced in size. In *fig. 8* is shown a cross section diagram of a tube insulated in this way with fan cooling. *Fig. 9* gives photographs of two "Rotalix" tubes, the lower with cooling bulbs according to fig. 7 and the upper with fan cooling according to fig. 8.

Because of the great mass of the anode (1200 g) the "Rotalix" tube is also specially suitable for a particular application, namely X-ray cinematog-

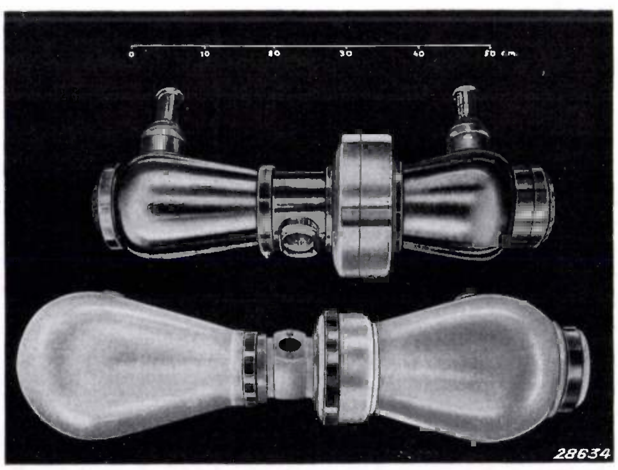


Fig. 9. Two "Rotalix" tubes projected against high tension and X-radiation: the lower with cooling bulb according to fig. 7, the upper with fan cooling according to fig. 8.

8) Since the radiation increases sharply with the temperature, it seems tempting for constructional reasons to choose the basic temperature T_A of the anode mass higher, in order to be able to use a smaller surface. A higher basic temperature however means, as we have seen above, a lower specific loading, while the problem is just to make this as high as possible.

raphy [9]). In the latter 12 exposures per sec. are made for 20 sec. for example. The X-ray tube is loaded for 20 consecutive seconds with more than 5 kW; the total amount of heat supplied thus amounts to more than 100 kW sec. ≈ 25000 cal. This heat supplied, in a mass of 1200 g and a specific heat of copper of about 0.08, caused a rise in temperature of the whole anode block of about 250°C. The cooling of the anode by radiation makes it possible to repeat such extremely heavy loadings at fairly short intervals.

[9]) van de Maele, J. belge Radol. Nr. 137, 1935.

THE DIFFRACTION OF LIGHT BY SOUND FILM

by J. F. SCHOUTEN

534.44 : 535.42 : 778.534.45

The light diffraction phenomena which are obtained with a strip of sound film may used to analyse the sound. If a pure sinusoidal oscillation is registered on a film according to the intensity system, the grating so obtained produces, in addition to the non-diffracted rays, only first order spectra. If the oscillation contains higher harmonics, diffraction spectra of higher orders also occur. The intensity of each order is determined by the amplitude of the corresponding harmonic. If the sound is registered according to the amplitude system the diffraction phenomena become more complicated. A two-dimensional diffraction pattern with two axes of symmetry is obtained. The distribution of intensity on the horizontal axis of symmetry, which is parallel to the axis of symmetry of the sound track on the film, corresponds to the diffraction spectrum of the film registered according to the intensity system, and may therefore, like that spectrum, be used to analyse the sound. In this article several examples are given: sinusoidal tone, tone with second harmonic, tone with third harmonic, transmission grating, tone with frequency modulation. Finally the frequency characteristic of an amplifier is determined according to the above method.

Introduction

In general light is composed of radiation of various wave lengths. An ordinary electric lamp, for example, emits all wave lengths to which the eye is sensitive to a certain degree depending uniformly on the wave length. This can best be seen by observing the light through a spectroscope. In such an instrument the kinds of light of different wave length are projected next to each other in a plane, so that an immediate impression is obtained of the spectral composition of the light.

If for example we observe sunlight through a spectroscope we see dark lines, the Frauenhofer lines, on a continuous spectrum. This means that those wave lengths are present in sunlight in only a slight amount. If, on the other hand, we consider a gas-discharge lamp, for example a sodium, mercury or a neon lamp, through a spectroscope, we observe, instead of a continuous spectrum, only a few narrow coloured lines on a dark background. This means that this light is composed of several definite wave length regions.

The presence and the position of these regions could also be determined roughly by observing the light source through coloured filters. Such filters transmit certain regions of the spectrum and absorb the rest of the light. We can now find out through which filters the light source is visible or invisible. The spectroscope, however, and that is its fundamental advantage over coloured filters, offers the possibility of obtaining an immediate impression of the number, position, intensity and extensiveness of these regions *i.e.* of the spectral composition of the light.

In the case of sound we encounter the same problems as with other waves. Every sound may be considered to be composed of pure sinusoidal oscillations of different frequencies, each of which by itself causes a sound impression of different pitch, and which together give the impression of a certain tone.

Among the sounds, with which we are practically concerned, the frequencies of which they are composed do not usually have such continuously distributed values as is the case for example with light. From the nature of the origin of the tones, the motion of the sources of sound is repeated periodically. This is seen most clearly when sound is registered with a cathode ray oscillograph. With this instrument a graphic representation is obtained of the displacement of the vibrating air raphy [9]). In the latter 12 exposures per sec. are made for 20 sec. for example. The X-ray tube is loaded for 20 consecutive seconds with more than 5 kW; the total amount of heat supplied thus amounts to more than 100 kW sec. ≈ 25000 cal. This heat supplied, in a mass of 1200 g and a specific heat of copper of about 0.08, caused a rise in temperature of the whole anode block of about 250°C. The cooling of the anode by radiation makes it possible to repeat such extremely heavy loadings at fairly short intervals.

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In the case of any purely periodic oscillation, it can be shown that it may be built up of a sinusoidal basic tone and sinusoidal overtones of double or triple frequencies, etc. The sound spectrum thus consists here of a number of discrete frequencies which are integral multiples of the basic frequency. The sound impression of such a tone is determined by the relative strengths of the basic tone and the various overtones.

It will have been noticed that in the analysis of light we always spoke of the wave lengths which compose it, and when speaking of sound, of the frequencies of which it is composed. The result is the same, because wave length and frequency can be expressed in terms of each other by means of the velocity of propagation. The difference in terms used is chiefly due to the fact that our elementary idea of the spectral composition of light is supported by the diffraction phenomena in which wave length is a decisive factor, while the spectral composition of sound is known to us mainly from resonance phenomena in which frequency is the decisive factor. Considering the fundamental analogy between the composition of light and that of sound, the obvious question arises as to the possibility of obtaining a direct spectrographic analysis of sound similar to that of light. Instruments actually exist for this purpose, but due to their complex construction they are only seldom used. The problem has therefore been solved by calculating from recorded oscillograms the sinusoidal components of which they are composed with the help of Fourier analysis. This is, however, generally a fairly elaborate method.

Nevertheless the question of the spectral composition of sounds is often very important. This is true of the study of the human voice and of the sounds produced by musical instruments. It is especially true in technical acoustics for the study of the distortion which may be caused by acoustic and electrical apparatus. This distortion is manifested in the fact that with a pure sinusoidal input signal of a given frequency, overtones also occur in the output signal. The problem is solved in the investigation of these overtones by going over the whole region with an electrical filter of adjustable frequency. How much more satisfactory it would be to be able to obtain an immediate idea of the spectral composition of the sound.

According to an original idea of Brown [1]) immediate insight into the spectral composition of sound can be obtained if the temporal periodicity of the sound is transformed into a special periodicity by recording it on a sound film. The latter can then be used as an optical diffraction grating.

We shall develop this idea further in this article, with special emphasis on the close analogy between the sound spectrum of a composite sound and the diffraction spectrum of an optical grating, an analogy which is expressed in the possibility of mathematical analysis into Fourier series.

The diffraction spectrum of a transmission grating

Sound film consists of a black strip in which a transparant strip has been left free. We distinguish between two systems: the intensity system and the amplitude system. In the intensity system the width of the transparent strip is kept constant and its transmissibility is varied proportionally to the amplitude of the sound recorded. In the amplitude system the transmissibility of the strip is kept constant, but its width varies proportionally to the amplitude of the sound recorded. If a narrow beam of light is projected on the film perpendicular to the direction of length of the film, and if the film is moved in the direction of its length, the amount of light transmitted will in both cases vary proportionally to the amplitude of the sound registered on the film, and may therefore be transformed into the original sound.

In order to understand the behaviour of such a section of sound film as a diffraction grating, we must refer to the theory of optical gratings.

We assume (*fig. 1*) the plane of the section of film to be vertical and the direction of its length to be horizontal. Then (see also fig. 4) we suppose that on this strip of film there are an infinite number of very fine equidistant lines of equal height which transmit light. We shall now investigate the spectrum of this transmission grating. A parallel beam of light falls horizontally on the film. The light may be obtained from a small source at a great distance. The light which passes through the slits (*fig. 2*) will be diffracted in a horizontal direction [2]). The direct rays will all be in the same phase and therefore reinforce each other. We can also find a certain angle of diffraction α at which the ray diffracted from one slit is exactly one wave length ahead or behind in phase to that from the two adjacent

[1]) D. Brown, Nature, **140**, 1099, 1937.

[2]) We shall in this article concern ourselves only with the diffraction in the horizontal direction and shall not discuss the diffraction in the vertical direction.

slits. For that angle the rays coming from the different slits will again be in phase and therefore also reinforce each other.

The same is true for every angle of diffraction at which two adjacent rays have a phase difference of a whole number of wave lengths. At certain angles therefore maxima will appear in the intensity transmitted. The condition for the appearance of these maxima is the following:

$$\sin \alpha_n = \frac{n\lambda}{l}, \quad n = 0, \pm 1, \pm 2, \pm 3, \ldots \quad (1)$$

λ is here the wave length of the light used and l the distance between the slits.

If a positive lens is placed behind the film, a point image of the source would be formed at the focus of the lens if the film were not there. Due to the presence of the film there is now a vertical line instead of the point, due to the refraction of light in a vertical direction. To the left and right occur a number of equidistant lines of equal intensity which we shall call the first, second, third, etc. orders according to the customary usage in spectroscopy (fig. 4).

A few words must be said about the order of magnitude of the angles of diffraction which prevail. With optical gratings l is about 2μ. At a wave length of 0.6μ for the first order $\sin \alpha = 0.3$ and α therefore about 18 degrees. With sound film, at 300 cycles the length of one cycle on the film, l, is 1 mm. Therefore $\sin \alpha = 0.006$ and $\alpha = 0.006$ radians $= 2$ minutes. We may therefore conveniently substitute α for $\sin \alpha$ and write

$$\alpha_n = \frac{n\lambda}{l}, \quad n = 0, \pm 1, \pm 2, \pm 3, \ldots \quad (2)$$

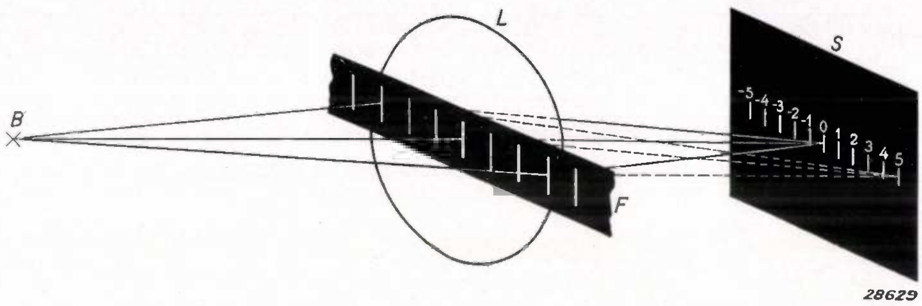


Fig. 1. Experimental arrangement for obtaining diffraction spectra. A point source of light B illuminates the strip of film F with practically parallel monochromatic light. On the screen S at the focus of lens L the diffraction spectrum may be seen. The strip of film F has a number of very narrow equidistant light transmitting slits. The spectrum now consists of a number of equidistant vertical lines.

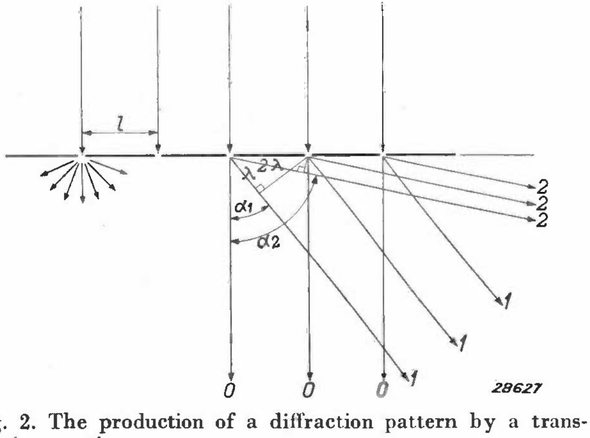


Fig. 2. The production of a diffraction pattern by a transmission grating.

Left: The light rays passing through the slits are refracted in all directions in a horizontal plane.

Right: The rays in a given direction coming from the different slits interfere with each other. The result is that the resulting intensity is equal to zero except for those directions in which two neighbouring rays have a phase difference equal to a whole number of wave lengths. In the figure are given the rays of the zero order (no phase difference), the first order (phase difference λ), and the second order (phase difference 2λ).

Except at the maxima the intensity of the diffraction spectrum must always be zero. For every angle which does not satisfy condition (2), in the case of every slit of the infinitely long grating there is always another slit such that the two rays are exactly in opposite phase and thus cancel each other.

The spectrum of an infinitely long ideal transmission grating contains all the higher orders in equal intensity.

The diffraction spectrum of a sinusoidal intensity grating

Let us now consider a strip of sound film on which a pure sine is recorded by the intensity system. The width of the strip is constant and the transmissibility as a function of the place x on the film is variable. Let the transmissibility D be given by

$$D = a + b \cos \frac{2\pi x}{l}, \quad (3)$$

i.e., by a constant component a and a sinusoidal variable component with amplitude b and period l.

The equidistant points on this "grating", $x = 0$, $x = \pm 1$, $x = \pm 2$, etc. all have the same transmissibility. Together they may be considered analogous to the transmission grating described in the previous section. In the focal plane this grating gives the following light oscillation for all orders

$$(a + b) \sin 2\pi\nu t,$$

where ν is the frequency of the light used (the amplitude is thus set equal to the transmissibility). In the same way the points $x, x \pm l, x \pm 2l$, etc. all have the same transmissibility $a + b \cos 2\pi x/l$. They may therefore also be considered as a transmission grating which, however, is shifted a distance x with respect to the first grating. The rays from this latter at an angle of α, cover a distance αx, and are therefore a phase angle $2\pi\alpha x/\lambda$, ahead of the first ones. They thus contribute to the light oscillations the amount:

$$(a + b \cos 2\pi x/l) \sin (2\pi\nu t + 2\pi\alpha x/\lambda);$$

or substituting the value of α in (2):

$$(a + b \cos 2\pi x/l) \sin (2\pi\nu t + 2\pi n x/l).$$

The total light oscillation M_n which is produced by the whole grating in the n^{th} order is

$$M_n = \int_0^l \left(a + b \cos \frac{2\pi x}{l}\right) \sin \left(2\pi\nu t + \frac{2\pi n x}{l}\right) dx \quad \cdot \quad (4)$$

When we set $2\pi x/l$ equal to φ, this gives, except for a constant factor, the following,

$$M_n = \int_0^{2\pi} (a + b \cos \varphi) \sin (2\pi\nu t + n\varphi)\, d\varphi =$$

$$= a\int_0^{2\pi} \sin (2\pi\nu t + n\varphi)\, d\varphi + b\int_0^{2\pi} \cos \varphi \cdot \sin (2\pi\nu t + n\varphi)\, d\varphi. \quad (5)$$

The first integral of (5) is zero for all values of n except $n = 0$, which gives

$$M_0 = 2\pi a \sin 2\pi\nu t \quad \cdot \cdot \cdot \cdot \cdot \quad (6)$$

The second integral of (5) may be written as follows:

$$\frac{b}{2}\int_0^{2\pi} \sin[2\pi\nu t + (n-1)\varphi]\, d\varphi +$$

$$+ \frac{b}{2}\int_0^{2\pi} \sin[2\pi\nu t + (n+1)\varphi]\, d\varphi.$$

The first integral of this expression is always equal to zero except for $n = +1$, when it gives

$$M_{+1} = \pi b \cdot \sin 2\pi\nu t, \quad \cdot \cdot \cdot \cdot \cdot \quad (7)$$

while the second integral is always equal to zero except when $n = -1$, when it gives

$$M_{-1} = \pi b \cdot \sin 2\pi\nu t \quad \cdot \cdot \cdot \cdot \cdot \quad (8)$$

Therefore if the signal $a + b \cos \varphi$ is registered on the film, we see in the zero order of the spectrum a light oscillation whose amplitude is proportional to the average value a of the signal, and in the first order to the right and left ($n = +1$ and -1) an oscillation whose amplitude is proportional to the amplitude b of the sinusoidal variations, while all higher orders are missing. The intensity of the light here becomes proportional to the squares of a and b respectively.

It is easy to show that when the registered sound does not consist of one, but of several pure tones, several lines will also occur in the diffraction spectrum, whose positions will be determined by the frequency, and whose intensity by the amplitude of the tones in question.

The diffraction spectrum of a strip of sound film recorded by the intensity system gives exactly the Fourier series of the sound recorded on the film.

This is substantially a well-known proposition of the theory of optical gratings, which now, however, may be observed from a different point of view through its interpretation for the study of sound. With optical gratings our problem is to deduce the spectral analysis of light from the spectrum. The characteristics of the diffraction grating in that case interest us only in so far as they affect the greatest attainable accuracy. In the use of sound film as optical grating our problem on the other hand is to deduce from the diffraction spectrum the characteristics of the grating. In this latter case it is the spectral composition of the light which may appear as a disturbing factor, since according to (2) the position of the maxima is a function of the wave length used. Therefore all records are made with monochromatic light (the green mercury line $\lambda = 5461$ Å).

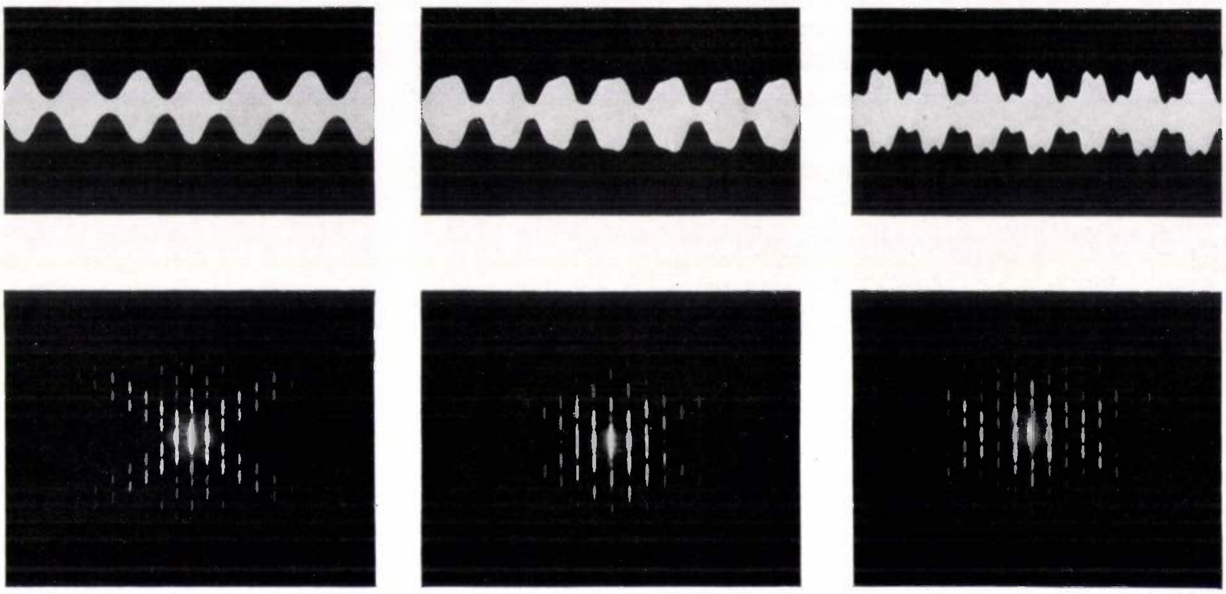
With a sound film recorded by the amplitude system, diffraction phenomena occur which are much more complicated than with a sound film recorded by the intensity system [3]).

In most cases a two-dimensional symmetrical pattern is obtained as shown in *fig. 3* for different sound tracks. Even when a pure sine is recorded in the strip, all the higher orders appear in the spectrum. On the horizontal axis of symmetry, however, the same proposition is valid as that deduced for the film recorded by the intensity system.

[3]) J. F. Schouten, Nature **141**, 914, 1938.

In fig. 3*a* a tenfold enlargement is given of a practically pure sinusoidal sound track and of the spectrum obtained from it. On the horizontal axis, in addition to the zero and first orders, only a very weak third order is visible. Fig. 3*b* refers to a sound track with strong second and weak third harmonics, and fig. 3*c* to a sound track with practically only a third harmonic.

In this way it is possible to analyse the spectrum weaker when the experiment is actually carried out. The explanation of this is simple from the point of view of Fourier analysis. The slits are of course not infinitesimally narrow, but have a definite, though small width. This is manifested in Fourier analysis by the fact that the strength of the overtones gradually decreases with increasing order, passes through zero and then varies periodically. The first decrease is shown clearly in fig. 4.



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a b c

Fig. 3. Harmonic analysis of sound registered on a strip of film by means of diffraction patterns.
Above: the strip of film used (tenfold enlargement).
Below: the diffraction pattern.
a) Practically sinusoidal oscillation. On the horizontal axis of symmetry of the diffraction pattern, except for a weak third order, only the zero and first orders appear.
b) Basic tone with strong second and weak third harmonics.
c) Basic tone with strong third harmonic. The second harmonic is missing.

of periodic phenomena. In the following we shall discuss several characteristic examples, and conclude with a special application: the determination of the frequency characteristic of a system.

The spectrum of a periodically repeated impulse

We deduced the proposition formulated above from the simple spectrum of a transmission grating. We may now, conversely, carry out experimentally the analysis into Fourier series of the transmission grating (in acoustics: a periodically repeated impulse) with the help of this proposition. As previously mentioned, all orders have the same intensity in the diffraction spectrum of a transmission grating. In the Fourier analysis of a periodically repeated impulse, therefore, all overtones occur with equal amplitude. As may be seen from *fig. 4*, however, the higher orders become gradually

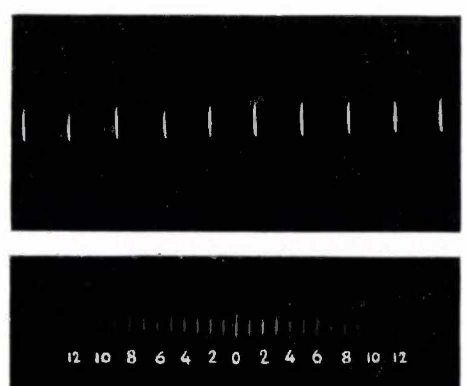

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Fig. 4. The spectrum of a transmission grating (periodically repeated impulse).
Above: tenfold enlargement of the strip of film used. Frequency 250 c/s.
Below: the spectrum of the transmission grating. All orders must appear in equal strength. The gradual decrease in intensity of the higher orders is due to the finite width of the slits on the film.

The spectrum of a finite train of waves

In the derivation of the spectra we have assumed until now that the strip of film was infinitely long. In practice the strip will always have a finite length, so that the spectrum always refers to a "wave train" of, for instance, several centimetres. How will this fact be manifested in the spectrum? As an example let us consider a finite portion of a pure sine. The Fourier analysis of an infinitely long sine gives us a discrete frequency, namely that of the sine itself. If, however, the oscillations are interrupted at a definite point the Fourier analysis of this function, which is actually no longer purely periodic, does not give a single frequency, but a frequency band, flanked by weaker side bands. It can be proved that the width of the central band in periods is equal to twice the reciprocal of the time interval in which the wave train passes.

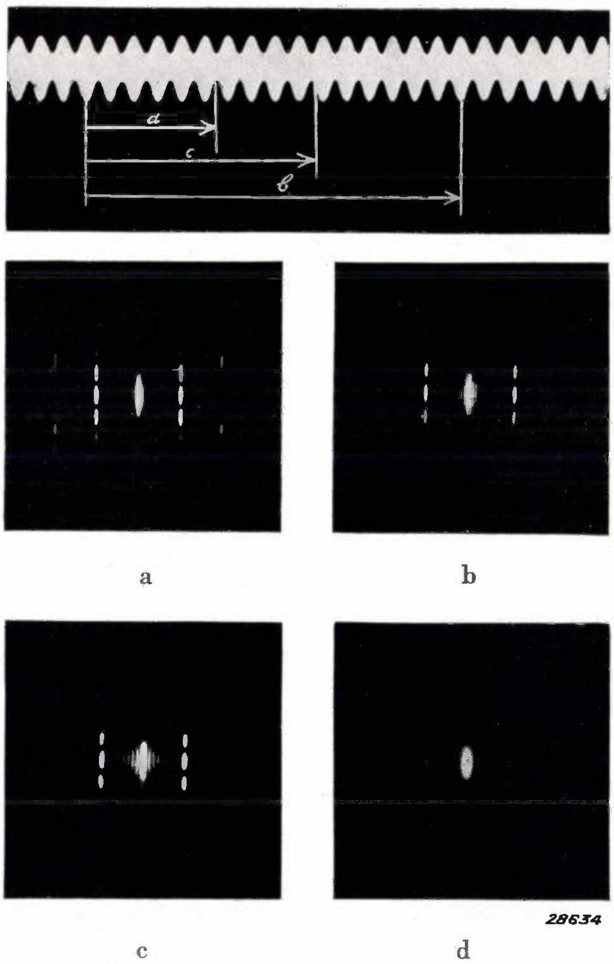

Fig. 5. Influence of the length of the strip of film on the width of the diffraction lines.
Above: tenfold enlargement of the strip of sound film. A pure tone of 730 c/s is recorded upon the film according to the amplitude system.
Below: spectra of strips of film of different lengths.
a) Length 33 mm = 0.11 sec. The width of the lines is still determined by the finite dimensions of the light source used.
b) Length 6.6 mm = 0.022 sec. The zero order exhibits the expected extra maxima. The width of the main line is 90 periods.
c) Length 4.0 mm = 0.01234 sec. The lines are clearly widened to 150 periods.
d) Length 2.3 mm = 0.0077 sec. The lines are very much widened, to 260 periods.

In *fig.* 5 may be seen the spectrum of four strips of sound film of the amplitude system upon which a pure sine of 730 c/s is recorded. The length of the strips is 33, 6.6, 4.0 2.3 mm. respectively, corresponding to 80, 16, 9.7 and 5.6 periods and to times of 0.11, 0.022. 0.0134 and 0.0077 seconds respectively. The width of the main lines must therefore be 18, 90, 150 and 260 periods respectively.

In fig. 5*a* this width is still small in comparison to the widening caused by the finite dimensions of the light source which is not exactly a point. In 5*b* the widening is scarcely noticeable. Next to the zero order, however, the side bands are now clearly visible. The width of the main line in the zero order is exactly 1/8 of the distance between two orders, which is equal to 730 periods, and therefore corresponds to the expected width of 90 periods. The films *c* and *d* show, as was to be expected, a continually increasing widening. The higher orders are not visible on the photographs since their intensity is too slight. Therefore upon shortening the strip of film the spectrum becomes steadily less sharp, so that fine structure can no longer be resolved. The well-known rule of grating spectroscopy, that the resolving power of a ruled grating is directly proportional to the width of the grating, is thus encountered here in the case of the spectra of short sections of film.

The spectrum of a warble tone

Warble tones are tones whose frequency varies periodically. They are used in acoustics in measurements of reverberation, and derive their name from the sound impression which they give.

The remarkable thing about these warble tones is that their Fourier spectrum consists of a number of discrete lines lying on both sides of their average frequency. In *fig. 6* may be seen the spectrum of a repeated impulse with periodically changing distances. When this spectrum is compared with that of fig. 4 it may be seen how each line has been more and more clearly resolved into a number of lines according as the order increases.

This phenomenon recalls the phenomenon of ghosts in grating spectroscopy. Almost every optical grating gives several weak extra lines called ghosts on both sides of the main line of the diffraction spectrum. These lines are due to periodic changes in the distance between the rulings of the grating, which in turn are due to periodic irregularities in the transmission mechanism of the ruling machine.

A similar phenomenon is met with in the frequency modulation in radio waves, whereby several side bands occur in addition to the carrier wave.

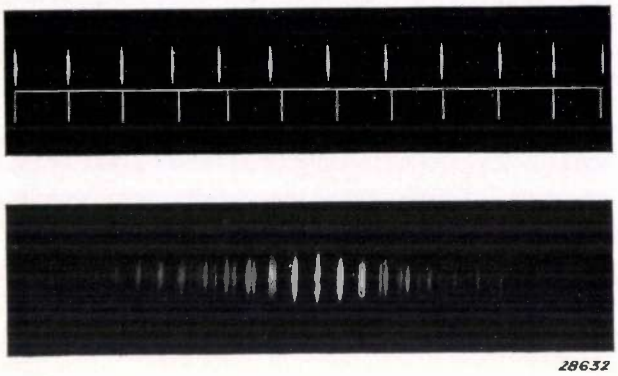

Fig. 6. Spectrum of a warble tone.
Above: tenfold enlargement of the strip of sound film, upon which a periodic impulse with periodically varying frequency is registered. Average frequency 200 c/s. The maximum or minimum of the frequency occurs 23 times per second.
Below: The spectrum of the warble tone. If this is compared with the spectrum of fig. 4, it will be seen that the higher orders have split into a number of lines.

The determination of the frequency characteristic of a system from the manner of motion under the influence of an impulse

The characteristics of an amplifier can be determined by the frequency characteristic. This is the ratio of output to input signal as a function of the frequency. For faithful reproduction it is essential that this ratio be independent of the frequency.

An amplifier may, however, be characterized in another way, namely by giving the amplifier a short impulse and registering on an oscillograph the consequent voltage occurring at the output terminals as a function of the time. If the amplifier is ideal the same impulse will appear in the output signal, *i.e.* on the oscillograph. If the apparatus is not ideal, and if, for example, the frequency characteristic exhibits a more or less pronounced maximum, instead of the impulse, a more or less damped oscillation is seen on the oscillograph, having a frequency practically equal to the resonance frequency. It is well known that the frequency characteristic can be deduced from this oscillogram. What does it mean when we give an amplifier a single impulse?

In the foregoing we have seen that a periodic impulse contains the main tone and the overtones all in equal intensities. If now we continually increase the time intervals at which the impulse is repeated, the main frequency becomes steadily lower and the overtones, which all have the same intensity and are equidistant in the frequency spectrum, begin to lie closer and closer to each other. If the time interval becomes infinitely great, only one impulse remains, and the series of lines passes over into a continuous spectrum in which all frequencies are equally strongly represented. The giving of a single impulse therefore comes down to the simultaneous application of all frequencies.

If the amplifier has a resonance in the frequency characteristic, the frequency corresponding to it will appear relatively strongest in the output signal. The resulting voltage will no longer have purely the form of an impulse, but will show a periodicity corresponding with the resonance frequency.

Another example may serve to clarify the relation between frequency characteristic and output signal for a single impulse. We take an ordinary pendulum and apply to it a periodic force. The pendulum will then move with the same period. For frequencies below the characteristic frequency this movement will be small in amplitude. As the frequency of the force applied rises and approaches the characteristic frequency of the pendulum, the amplitude of the pendulum will increase steadily and decrease again for frequencies higher than the characteristic frequency of the pendulum. In this way we have determined the frequency characteristic of the pendulum. When we now give the pendulum a blow, *i.e.* apply a sudden force to the pendulum for a short time, it will not immediately move with a jerk and then return quickly to the position of rest. The pendulum will rather begin to move with a finite velocity and then oscillating in its characteristic frequency return gradually to its position of rest.

This is precisely the same behaviour as above. By means of the impulse, forces of all frequencies in equal strengths are applied the pendulum. Because of the maximum in the frequency characteristic, however, the frequency corresponding to the maximum appears more strongly in the amplitude, and thereby leads to a periodic motion of the pendulum. According as the maximum of the frequency is more or less flat, the motion of the pendulum is more or less damped.

We may therefore determine the frequency characteristic of a system by resolving the oscillogram obtained with a sinusoidal input signal in the form of an impulse into its Fourier components. By means of the spectra method this latter is easily possible. In *fig.* 7 may be seen a strip of sound film registered by the amplitude system. A periodic impulse was here applied to a system with a characteristic resonance, and the output signal was recorded on the film. It may be seen that amplitude returns to zero in the form of a damped oscillation after every impulse. The Fourier analysis of the strip of film shows that the oscillation no longer contains the various overtones all with the same strength, but with a strength which is directly proportional to the amplification factor of the system for the frequency in question. In the diffrac-

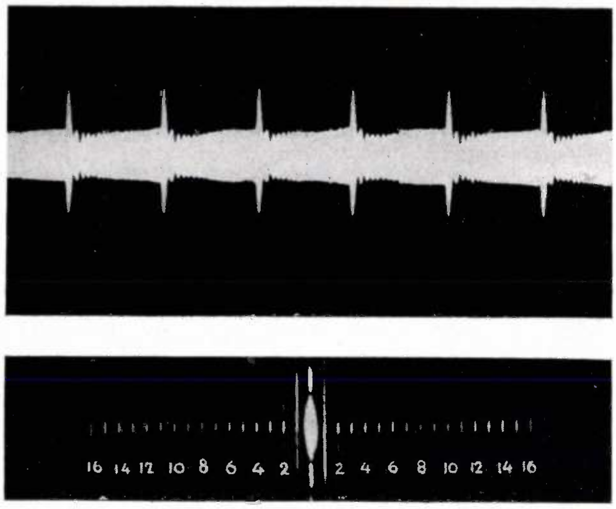

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Fig. 7. Determination of the frequency characteristic.
Above: A periodic impulse of 250 c/s is applied to a system with a characteristic resonance in the neigbourhood of 4 000 c/s. The input is then recorded according to the amplitude system. The amplitude is now not that of the pure impulse, but that of a damped oscillation (tenfold enlargement of the film).
Below: The spectrum of the strip of film. If this is compared with the spectrum of fig. 4, it may be seen that the higher orders are relatively stronger. This is due to the rise in the frequency characteristic for those frequencies (see fig. 8).

tion spectrum of fig. 7 this is manifested by the fact that the strength of the overtones is not proportional to that in fig. 4, but that here the strength after an initial decrease, increases again and reaches a maximum at about the 15th overtone. The frequency characteristic of the system is now obtained by measuring the intensity of every overtone, in fig. 4 and 7, and extracting the square root of their ratio. The characteristic is given in *fig. 8*.

It may be seen from this how it is possible to determine the whole frequency characteristic of a system from a strip of film several centimetres long. In principle a still much shorter length, namely that of one period, should even be sufficient. Then instead of a series of equidistant lines a continuous diffraction spectrum would be obtained, whose variation in intensity corresponds to the frequency characteristic of the system. The light intensity would then however be very small, and this would make the measurements difficult.

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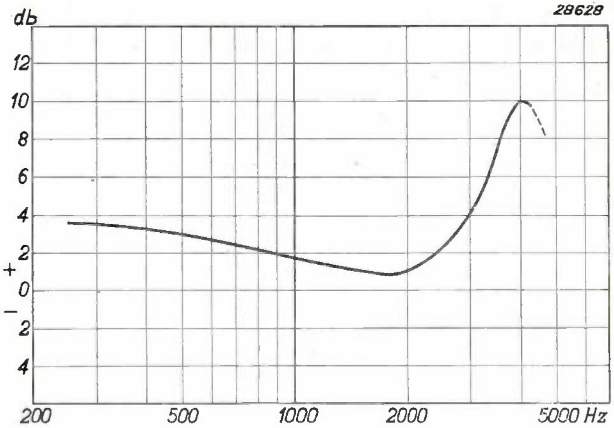


Fig. 8. The frequency characteristic calculated from the spectra of figs. 4 and 7. The ratio between the intensities of corresponding orders of figs. 4 and 7 is a measure for the corresponding ordinate of the frequency characteristic of the system for the frequency in question. By the measurement of intensities in the spectrum the whole frquency characteristic can be determined in this way.

IMPULSE VOLTAGE INSTALLATIONS

by W. HONDIUS BOLDINGH.

621.319.5

By means of standardized components a series of generators may be build to give impulse voltages of 800 kV to 4 million volts, and impulse energies of 3.2 to 80 kilowatt seconds. The base of the apparatus is 2 × 2 metres, its height about 2 m/million volt. The condensers have an energy content of up to 80 wattseconds per cubic decimetre. A switch mechanism makes possible whole or partial series-parallel connection of the condensers in a simple way, so that impulses of lower voltage with the full energy can also be produced. Installations for impulse voltages up to 1200 kilovolts can be transported ready for use.

Introduction

Since an impulse voltage installation was described in the first volume of this periodical [1]) considerable technical progress has been made in this field, and we shall discuss the result in the present article. This development coincided with the increasing interest on the part of power engineers in impulse voltage research, for which higher and higher voltages and especially larger and larger capacities of the impulse generator were demanded.

The knowledge that a large percentage of disturbances in high tension mains, power plants and transformer stations is caused by brief excess voltages due to atmospheric discharges has emphasized more and more the desirability of impulse voltage tests, of high tension installations ready for use as well as their separate elements. The great capacity toward earth of long transmission lines, of the high tension windings of large transformer units and of other test objects explains the demand for large impulse capacity. The charge of the generator must be high with respect to that necessary to raise the test object to the required potential, so that no excessif voltage loss will occur during the impulse.

Since as a rule the insulation of the test object to earth must be tested, standard generators will be constructed exclusively with one pole earthed, but in such a way that they can give positive as well as negative voltage with respect to earth.

In practice it is furthermore important that the whole impulse installation can be moved to any point of the high tension mains to be tested, without the necessity of elaborate assembly dismantling, so that the investigation need not be confined to the factory or laboratory.

For the examination of outdoor lines it is sometimes desirable to be able to use the full energy at a low voltage, which can be attained by connecting the impulse condensers wholly or partially in parallel. Then with the maximum energy an impulse of low voltage and high current can be produced, in order to burn through with this so-called surge current a defect caused with high tension, and thus bring it to light.

The above considerations led to the design of the impulse voltage generators to be described in this article. Although they naturally differ in many respects from the generators for high constant voltage which were developed for research on the problems of nuclear physics [2]), there are nevertheless many common structural features. The chief common characteristic is that by adequate designing of each composing part the main dimensions, even for very high voltages and energies could be kept within very narrow limits.

As for the impulse generators, it was made possible in this way to house an installation for several megavolts in a laboratory or factory hall of reasonable size, and on the other hand it was made possible to transport installations for up to 1.2 megavolts in working order.

Scheme of the impulse generator

Before we go into the particulars of the circuit, we shall briefly recall the principle of the Marx circuit which is the one used here. A charging generator charges a number of condensers (the "impulse condensers") connected in parallel *via* suitable resistances to a high voltage with the condensers. As soon as a definite voltage is reached a series of spark gaps break down and the impulse condensers are suddenly connected in series. A voltage is obtained which is equal to the sum of the voltages of the impulse condensers, which now discharge (again through suitable resistances). A damping resistance prevents the occurrence of any oscillations on the voltage wave.

Even when we confine ourselves to unipolar

[1]) A. Kuntke: Philips techn. Rev. 1, 235, 1936.

[2]) A. Bouwers and A. Kuntke: Z. techn. Phys. 18, 209, 1937; cf. Philips techn. Rev. 1, 6, 1936 and 2, 161, 1937.

impulse generators, there are in general two fundamentally different possible circuits for the charging generator: the symmetrical and the uni-polar circuit. *Fig. 1 a* and *b* give diagrammatically the principle

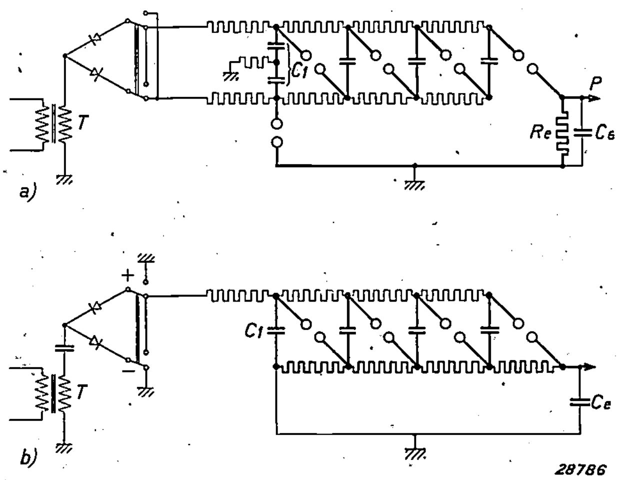


Fig. 1. Fundamental scheme of a uni-polar impulse voltage installation (four stages) *a*) with symmetrical charging generator, *b*) with uni-polar charging generator. T = high tension transformer, C_1 = first impulse condenser, C_e = end condenser, R_e = discharging resistance. The test object is connected at P.

of these two cases. (For the sake of simplicity the damping resistance is not drawn here). The symmetrical circuit is more often used, and from the stand point of insulation it is also the simpler. If the high tension transformer T gives a peak voltage E_s then with the symmetrical charging generator (fig. 1*a*) each of the two supply connections is only at the potential E_s (either plus and minus) with respect to earth. With the uni-polar charging generator, on the other hand, where the charging voltage 2 E_s is obtained by cascade arrangement of the two rectifier valves (fig. 1*b*), the supply connection is at the full charging potential 2 E_s with respect to earth. Therefore with the uni-polar charging generator more insulation is required especially since the polarity must be able to be reversed, and thus both terminals must be insulated for the full voltage of 2 E_s.

Nevertheless a uni-polar charging generator was chosen for the Philips impulse voltage installations, since it has various advantages.

1) Part of the charging resistances can also serve as discharging resistance (thick lines in fig. 1*b*). Apart from a saving in material by the omission of the special discharging resistance R_e (fig. 1*a*) this has the advantage that there is no large current loop during the discharge: each condenser is discharged through its own discharging resistance. The external inductance which increases with the area of the current loop is in this way reduced to a minimum.
2) No spark gap to earth is necessary (*cf*, for contrast fig. 1*a*), since the first condenser C_1 is already uni-polarly earthed even during charging. In addition to the centre tap of C_1, the special arrangement for regulating the spark gap to earth, which must always be adjusted to half of the sparking potential of the other spark gaps, may also be omitted.

By the abolition of the spark to earth an important source of oscillations on the voltage wave is avoided, since C_1 is not first earthed by the break down of the spark earth. Both of the arguments mentioned above contribute to obtain a smooth voltage wave without using large damping resistances causing excessive voltage loss, or any other complication.

In designing an impulse generator the voltage of the charging generator, *i.e.* the voltage per stage, is very important, since it determines the number of stages for the total voltage required. The charging generator in the Philips impulse voltage installations gives a maximum charging voltage of 200 kilovolts. It is so generously designed that even at the highest impulse energy the desired rapid succession of impulses (for instance every 10 seconds) can be obtained. The charging generator, as well as the impulse generator itself, is mounted on a base plate, which may be provided with wheels, so that it can easily be transported. Because of the uni-polar circuit, in which the full direct voltage of 200 kilovolts with respect to earth is obtained, the charging generator by itself (separate from the impulse generator) may very well be used for other purposes, such for example as the testing of cables.

Main components of the installations

In the construction of the impulse generators to be described here the use of "high-tension Philite", which has already been described in this periodical [3]), has been of great value. This insulating material has a break-down strength of 30 kilovolts/mm, and may therefore be compared with the best high-tension porcelain. It has thereby the advantage of very small dimension tolerances so that it can be made with screw-threads. "Philite" is used in the Philips impulse generators chiefly in the form of cylinders, with or without a bottom and with or without a flange, which may be screwed together and to other composing elements. The

[3]) L. L. C. Polis: Philips techn. Rev. 3, 9, 1938.

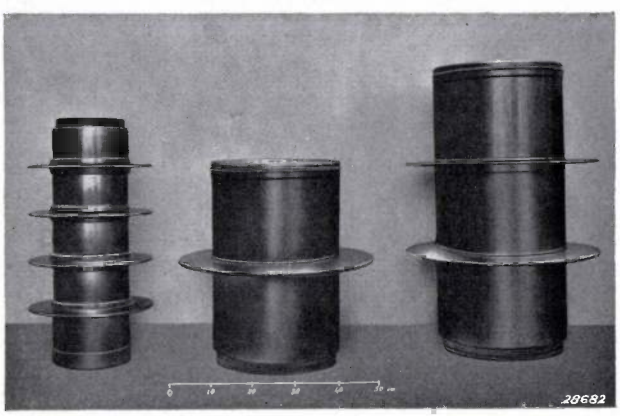
cylinders are made in two sizes: with a diameter (not including the flange) of 18 and 30 cm and a height of 11.1 and 22.2 cm respectively, and used for insulating supports, condensers (with the necessary number of condenser units connected in series simply piled up in the "Philite" cylinder), for resistances, etc.

Condensers

The impulse condensers are naturally the most important parts of the impulse generator. Their shape and size determine the design of the whole installation. The fact that great progress has been made in this field is obvious when the apparatus described here is compared with that of two years ago. Then condensers for 150 kilovolts were used with a capacity of 0.01 μF; the energy content ($^1/_2\ C.V^2$) per stage therefore amounted to 112.5 W sec. The length was 30 cm, while the diameter was 18 cm, so that 15 W sec per dm³ is taken up. This was then a very high figure.

At present condensers for 200 kilovolts maximum working voltage are used with a capacity of 0.125 μF, which can therefore take up 2500 W sec. with a length of 48 cm and a diameter of 30 cm. With a condenser length of 70 cm these values rise to 0.2 μF and 4 kilowatt sec., which corresponds to *80 W sec/dm³*. These condensers, whose length and therefore energy content can be made still greater, can easily meet present requirements as to impulse energy at high tensions, without it being necessary to connect several of them in parallel.

For impulse generators with smaller energy, for instance for educational purposes, smaller condensers are used (see in *table I* at 800 and 1200 kilovolts). They are built up out of the previously mentioned "Philite" rings 18 cm in diameter. A five-ring condenser of this type for 200 kilovolts



a b c

Fig. 2. Impulse condensers for 200 kV. *a*) 0.04 μF, 800 W sec; *b*) 0.125 μF, 2.5 kW sec; *c*) 0.2 μF, 4 kW sec.

has a capacity of 0.04 μF, *i.e.* an energy content of 800 W sec. Such condensers are used especially in constant voltage cascade generators.

The three types of condensers for 200 kilovolts, of 0.8, 2.5 and 4 kW sec respectively, are shown in *fig. 2*.

A remarkable characteristic of these condensers is that they are short circuit proof. They can be short circuited in the fully charged state without any series resistance. If this is done with a condenser as in fig. 2*b* in which 2500 W sec are thereby transformed into heat (600 cal) in a few microseconds, and if the discharge voltage is recorded, it will be seen from the damped oscillation which appears that the internal resistance of the condensers is less than 0.5 Ohm, and the inductance at the most 0.7 μH at a frequency of about 10^6 c/s.

In order to be able to obtain the required steepness of the wave front, it is desirable to employ a so-called end condenser. The duration of the wave front, t_s, which the voltage at the test object needs to reach its maximum value, is determined approximately by the relation:

$$t_s = 2\, R_d \cdot C_p. \qquad (1)$$

It will be smaller the smaller the total capacity C_p in the circuit, and the smaller the total damping resistance R_d. In order to test all objects with about the same standardized duration of the wave front t_s, the latter could be regulated only by changing the damping resistance. With test objects with small capacities a large damping resistance would then be necessary. An upper limit is, however, prescribed for the damping resistance in connection with the permissible voltage drop: R_d may not amount to more than 10 per cent of the discharging resistance R_e. The latter is already limited: its value is determined mainly by the standardized value of the time to half value of the wave tail t_h, and by the capacity $C_s + C_p$ of the generator (impulse capacity and load capacity) according to the formula:

$$t_h = R_e \cdot (C_s + C_p) \ln 2. \qquad (2)$$

In generators with a high impulse capacity therefore the discharging resistance as well as the damping resistance must be small. The smaller the latter resistance can be made the better, from the standpoint of voltage loss.

It is therefore preferable to regulate the time t_s by the addition of the end condenser already mentioned. This may then have so large a value that the capacity of the test object has little influence within wide limits on the time t_s. The total loading capacity, however, may also not be too great since this also causes voltage drop. 20 per cent of the impulse capacity is accepted as a maximum value. The end condenser is then also always considerably smaller than the impulse condenser. In the Philips impulse voltage installations it consists of units of small diameter, for instance for 400 kilovolts, which may be connected in series or in parallel as desired. With test objects of large capacity the end condenser can of course be omitted.

Resistances

While formerly liquid or carbon resistances were commonly used because of the high value of the resistance necessary to obtain the normal wave shape with the small impulse capacities then in use metal resistances can and must always be used now with the so much greater impulse energies. The resistance necessary for charging, discharging, damping and measuring are all mounted in dust-proof insulating material, and are of such generous dimensions that no excessive heating occurs, even with the highest impulse energies and short impulse intervals.

While the resistances which serve exclusively for charging may be wound in the ordinary way, the smallest possible inductance must be attained for the damping, discharging and measuring resistances. The damping resistance which, as we have seen, must be small in connection with the voltage drop, must on the other hand be large enough to insure the non-periodicity of the circuit $C_s - R_d - C_p$. For this the following condition is valid as an approximation:

$$R_d > 2\sqrt{L/C_p}, \quad \cdots\cdots \quad (3)$$

where L is the total inductance of the circuit in question. It was possible to reduce self-induction of the damping resistance to such a low value that its resistance value can even be chosen smaller than the permissible limiting value of 10 per cent of the discharging resistance.

The total damping resistance is subdivided in such a way (see fig. 4) that each impulse condenser is connected in series between two equal sections of it. The damping is hereby so large that no high frequency oscillations on the main wave can occur in each separate stage due to parasitic extra capacities.

In order to be able to discharge the impulse condensers under all circumstances even without an impulse on the test object, an earthing resistance is built into the charging generator (not shown in fig. 1). This resistance has such dimensions that if necessary it can take up the full energy of all the impulse condensers together. The earthing switch earths the whole installation automatically when charging is interrupted.

Construction of the impulse generator

The impulse generator consists of four columns, in each of which condensers and insulators are mounted alternately one above the other. The condensers are connected through resistances in such a way that the Marx circuit is built up in the form of a screw (see *fig.* 3). With this method of construction any number of stages may be completed, and

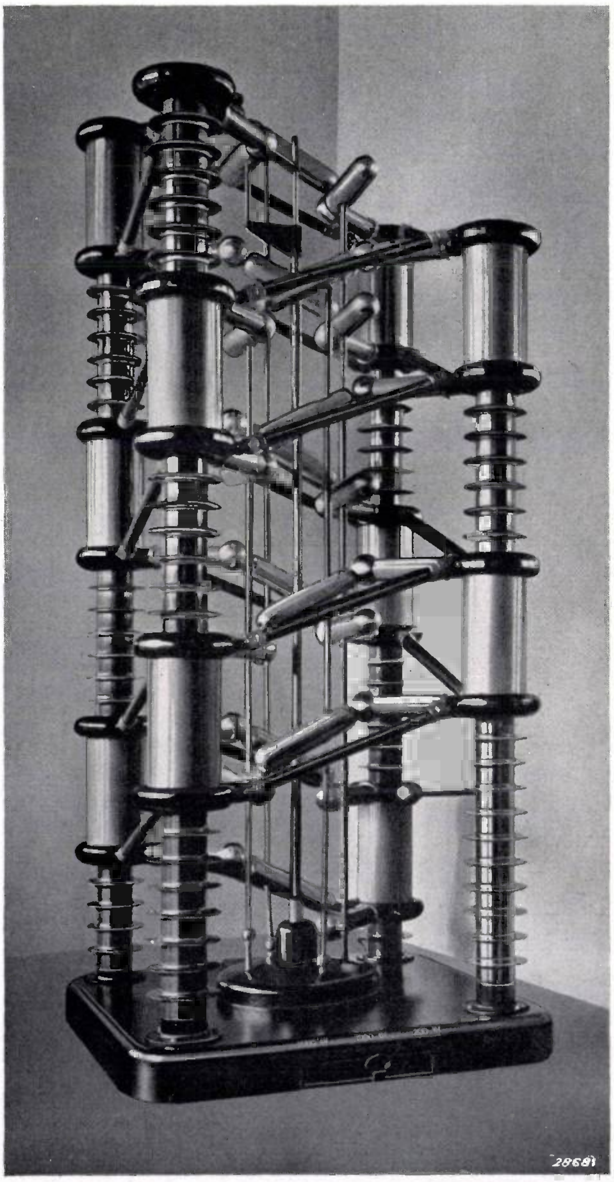


Fig. 3. Impulse generator with a impulse energy of 25 kW sec, commutable for 2000, 1000 and 200 kV (with 0.0125, 0.05 and 1.25 μF respectively). Height 4.5 m, base 2x2 m. In this type the large "Philite" cylinders are not yet used, so that the condensers may easily be distinguished from the insulators.

others added later, without increasing the necessary floor space, which is always 1x2 m. For reasons which will discussed later, the preferred number of stages will be as a rule a multiple of four, and therefore the impulse voltage will be a multiple of 800 kilovolts. In *table I* (page 316) the data are given of a series of impulse generators for different voltages, assembled with the three types of condensers shown in fig. 2.

The compact construction of the generators, which is illustrated by the main dimensions given in the table, has besides to purely spacial advantages the additional satisfactory result that, aside from the above-mentioned reduced inductance of the damping resistances, the remaining inductance in the circuit in question is also small.

The damping resistances are mounted at the top and bottom of every impulse condenser, and each has a switch arm. In this way every condenser is provided with two switch arms mounted at to top and the lower one of which serves as spark gap, while both together can connect the condenser in parallel with its predecessor. The spark gaps can be adjusted by remote controlled simultaneous motion of all the spark gap arms by means of five insulating bars (see in fig. 3 in the middle between the four columns). The middle one of these bars always serves the last, highest spark gap, the other four serve the other spark gaps and at the same time they serve to connect the impulse condensers wholly or partially in parallel by means of the second series of switch arms which are coupled in pairs with the spark gap arms.

Possibilities of commutation

In *fig. 4* the different possibilities of commutation are drawn for the case of a generator with 8 stages, *i.e.* for 1600 kilovolts. The switch arms with the same numbers, see fig. 4*a*, are controlled by a common bar (due to the screw-shaped construction they lie vertically above each other). By means of bar *I* condenser *1* is connected in parallel with *2*, and *5* with *6*, while by bar *III*, *3* is put in parallel with *4*, and *7* with *8*, see fig. 4*b*. At half voltage a fourfold capacity is now obtained. If now by means of bar *II* condenser *2* is put in parallel with *3* and *6* with *7* (fig. 4*c*) there are then two groups each of four condensers in parallel, so that 1/4 of the total voltage can be reached (400 kilovolts) with 16 times the original capacity. Finally by means of bar *IV* *4* and *5* can also be connected in parallel (fig. 4*d*) so that a surge voltage of 200 kilovolts with a capacity 64 times the original capacity is obtained.

It is clear that all the possibilities of commutation can only be utilized at nominal voltages which are a multiple of 800 kilovolts, *i.e.* when the number of stages is a multiple of four. Table I gives an ideal of the voltages and capacities which may be obtained by commutation.

Fig. 5 is a photograph of a commutable impulse generator for 16 kW sec in the present form with

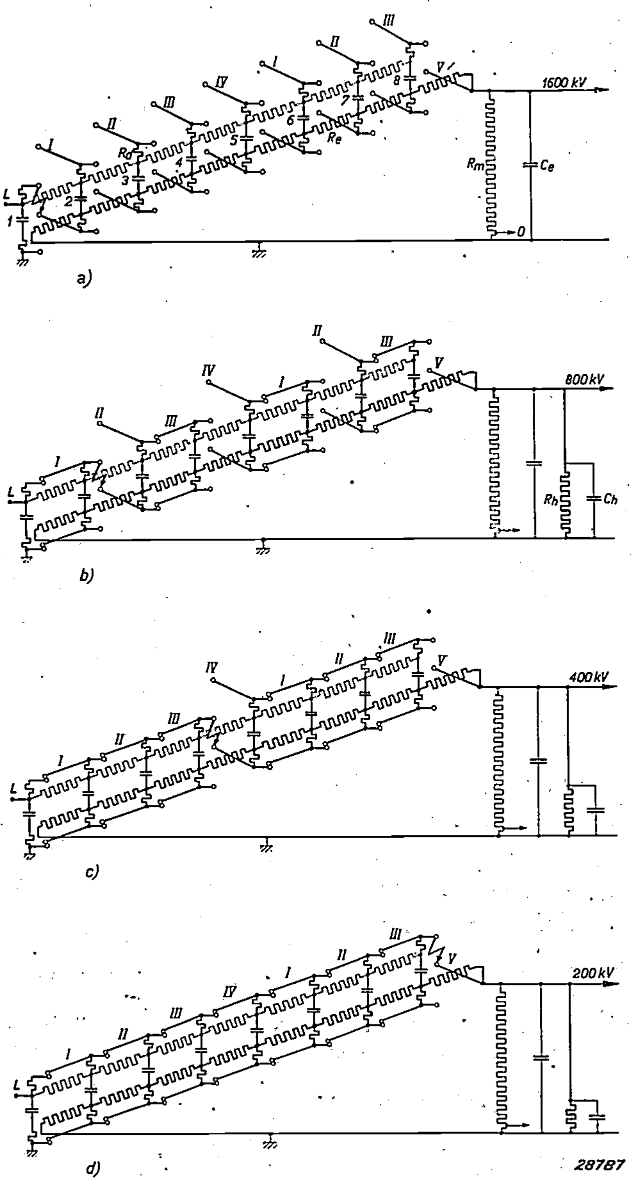


Fig. 4. Diagram of an impulse generator which can be commuted for voltages of 1600, 800 400 and 200 kV with the full surge energy in each case. The circuit which is actually built up in the form of a screw is shown here "unwound". The switch arms *I*, *II*, ... are operated by five bars (visible in fig. 3). The charging generator is connected at *L*. R_d = damping resistances, R_e = discharging resistances, R_m = measuring resistances, C_e = end condenser, C_h = auxiliary condenser, R_h = auxiliary end resistance. The cathode ray oscillograph is connected at *0*. *a*) All condensers, *1*, *2*, ... *8*, in series. *b*) Switch arms *I* and *II* closed: four groups of two condensers in parallel, 800 kV. *c*) Switch arm *II* also closed: two groups of four condensers in parallel, 400 kV. *d*) Switch arm *IV* also closed: all eight condensers in parallel, 200 kV. The ignition spark gap is indicated in each case by a zigzag arrow.

Table I Data of the Philips impulse generators

Nominal voltage	kV	800			1200			1600		2400		3200		4000	
Commutation voltages	kV	400/200			600/200			800/400/200		1200/600/200		**1600/800/200**		2000/1000/200	
Stages of 200 kV	—	4			6			8		12		16		20	
Energy per stage	kWsec	0.8	2.5	4.0	0.8	2.5	4.0	2.5	4.0	2.5	**4.0**	**2.5**	**4.0**	**2.5**	4.0
Capacity per stage	μF	0.04	0.125	0.2	0.04	0.125	0.2	0.125	0.2	0.125	**0.2**	**0.125**	**0.2**	**0.125**	0.2
Total energy	kWsec	3.2	10	16	4.8	15	24	20	32	30	48	40	64	50	80
Total capacity	in 10^3 pF	10	30	50	7	20	33	16	25	10	17	8	12.5	6.25	10
Cap. at half voltage	in 10^3 pF	40	125	200	27	85	130	62.5	100	40	70	30	50	25	40
Cap. at quarter voltage	in 10^3 pF							250	400	170	270	125	200	100	160
Cap. at 200 kV	in 10^3 pF	160	500	800	240	750	1200	1000	1600	1500	2400	2000	3200	2500	4000
Height	m	2.3	2.3	2.6	3.0	3.0	3.4	3.7	4.2	5.1	5.8	6.5	7.4	7.9	9.0
Min. height of room	m	3.3	3.3	3.6	4.5	4.5	4.9	5.7	6.2	8.1	8.8	10.5	11.4	13.0	14.0
Weight	kg	1200	1700	1800	1500	1950	2100	2200	2400	2700	3000	3200	3600	3700	4200

large "Philite" cylinders. The dimensions indicated in table I refer to this type.

The switch arms are controlled electrically from the control desk. The desired series parallel circuit is obtained after moving a handle in the base plate of the generator which can only be operated when the installation is completely earthed.

In order always to obtain the same form of impulse wave in agreement with the formulae mentioned, even with impulse condensers in parallel, the values of the discharging resistance R_e, damping resistance R_d and the load capacity C_p must be adapted to the larger impulse capacity C_s obtained.

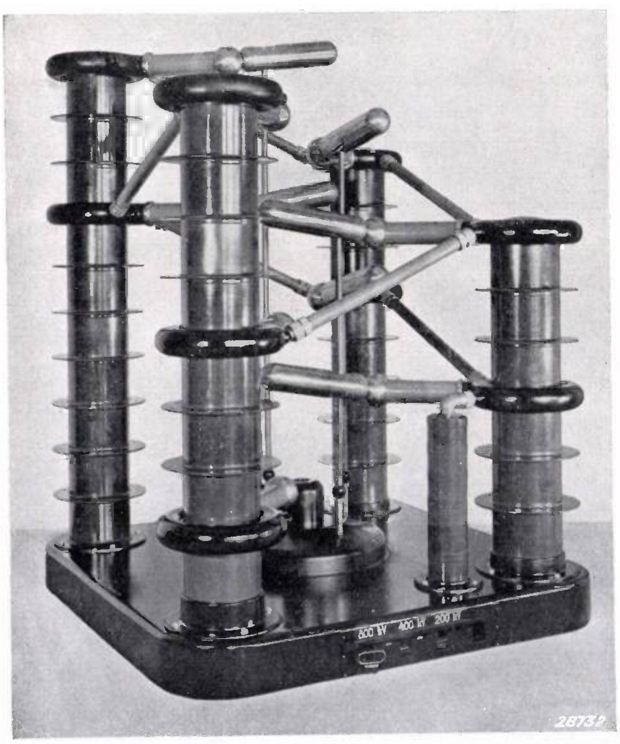


Fig. 5. Impulse generator with an impulse energy of 16 kW sec, commutable for 800, 400 and 200 kV (with 0.05, 0.2 and 0.8 μF respectively) in the present form with large "Philite" cylinders. Height 2.6 m.

In order to retain the original time of halve value of the wave tail with impulse condensers in parallel, the value of the discharging, (end) resistance must be reduced according to formula (2). From fig. 4 it is seen that when the impulse condensers are put in parallel, part of the discharging resistance (which are also the charging resistances) is short circuited. However this is not sufficient, there must also be an auxiliary end resistance (in fig. 4*b* indicated by R_h) in parallel with the test object.

Since the value of the damping resistance, as explained above, may not be more than 10 per cent of the end resistance, the damping resistance must also be reduced upon commutation. As may be seen from fig. 4 the parts of the damping resistance are automatically put in parallel, together with the corresponding impulse condensers, By this change in R_d, however, according to equation (1) the front duration t_s of the impulse wave is affected. In order to retain the normal value of t_s the end condenser might to be divided in the same way as the damping resistance and the parts connected in parallel. However it is often simpler to add an auxiliary condenser (indicated by C_h in fig. 4*b*) such that the original value of t_s is recovered. The auxiliary condenser C_h and the auxiliary resistance R_h are not combined in a single unit, in order to be able to omit the auxiliary condenser if necessary with test objects with large capacities, or to replace it by a smaller one.

The auxiliary end resistance which is used in the circuit for 200 kilovolts must be able to take up practically the entire impulse energy with the required impulse intervals.

Measurement of the voltage

For the measurement of the course of the impulse voltage the cathode ray oscillograph is now generally recognized as the only suitable instrument. It gives not only the crest value, but makes it possible to follow the entire course of the voltage wave. The indication must of cource be a faithful reproduction of the phenomenon to be measured. Since we are here concerned with a single non-periodic phenomenon of very short duration it is preferable to use a resistance potentiometer above a capacitive one, such as is sometimes used.

It is obvious that the measuring resistance may only have a minimum inductance. The ordinary methods of measuring this latter cannot be used because of the high damping. For the determination of self-inductance of the measuring resistance a model is constructed in which the resistance wire was replaced by copper wire, still retaining the geometric form (which determines the inductance in the first instance).

In how far a true reproduction is obtained with the measuring resistance used may be seen most clearly from an oscillogram of a breakdown at the wave crest of an impulse, as is given in *fig. 6* together with an oscillogram of a normal wave.

Besides the impulse voltage itself the charging voltage may also be measured. Thanks to the unipolar circuit of the charging generator it is possible by means of a very large series resistance to indicate the charging voltage of the impulse condensers directly with a voltmeter on the control desk. This gives a more complete and more accurate check on the voltage during charging, than the measurement of the length of the spark, by which only the peak value of the charging voltage can be determined.

a)

b)

c)

d)

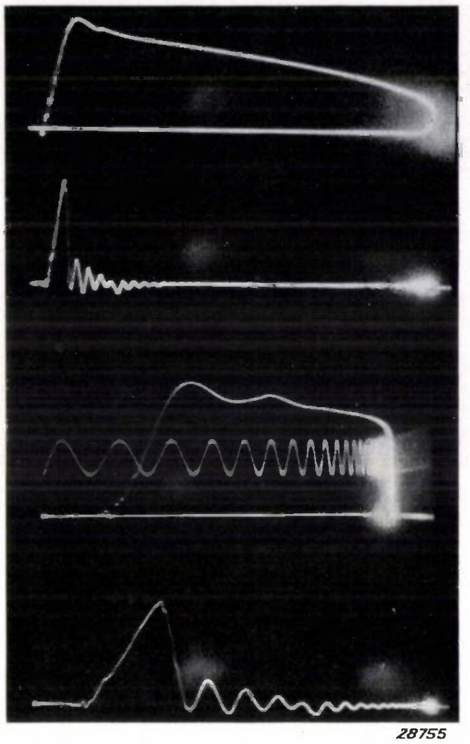


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Fig. 6. *a*) Oscillogram of a normal 1.0/50 wave (duration of the front t_s = 1.0 μ sec, time to half value of the wavetail t_h 50 μ sec). *b*) Oscillogram of the same wave with breakdown at the crest. *c*) Like *a*) with larger time base. *d*) Like *b*) with larger time base. The time base is obtained by the exponentially decreasing voltage during discharge of a condenser, in *c* the time is recorded by means of a sinusoidal oscillation of 10^6 c/s, in which therefore one period indicates 1 μ sec.

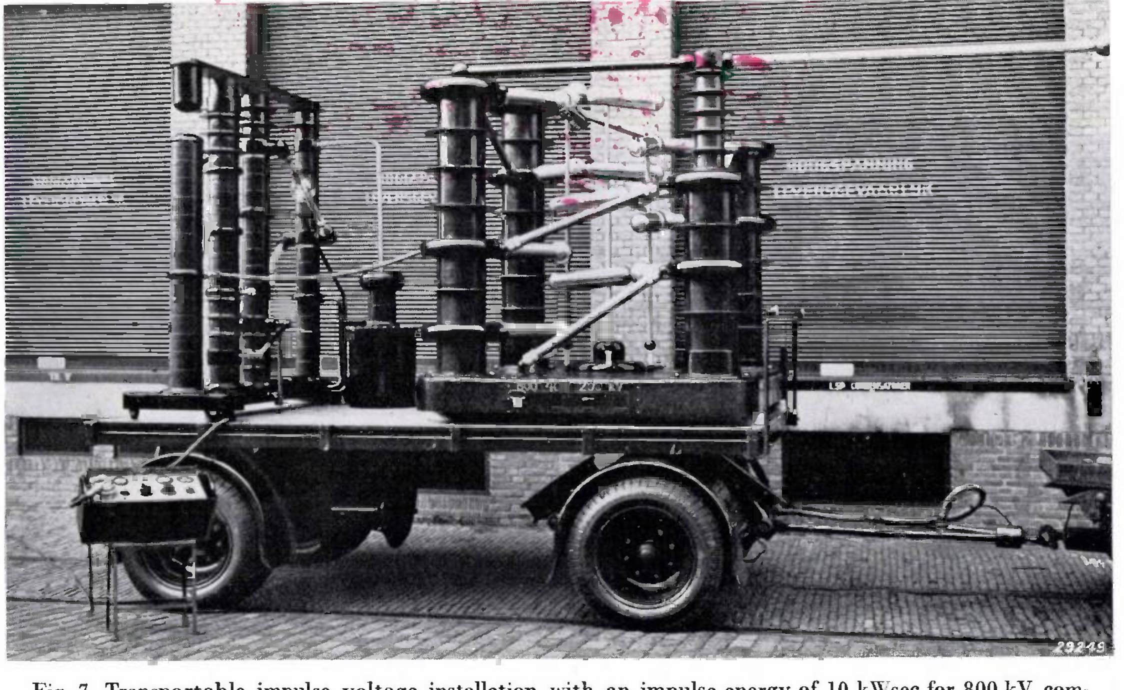

Fig. 7. Transportable impulse voltage installation with an impulse energy of 10 kWsec for 800 kV commutable for 400 and 200 kV. Surface of the trailer 2.20·4 m², total height from the earth 3.30 m.

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THE BRIGHTNESS OF ROAD SURFACES UNDER ARTIFICIAL ILLUMINATION

by J. BERGMANS.

535.241.44 : 628.971.6

In this article [1]) it is shown that the distribution of brightness of the road surface either in the wet or the dry condition has no constant relation to the distribution of the illumination. The reason for this is that the parts of the road surface at greater distances from the observer are viewed at smaller angles to the surface. A continuous layer of water on the road surface makes the designing of a good road lighting system impossible. Methods are indicated for determining the reflective properties for the dry and the damp condition (brightness coefficients). A completely worked-out diagram of brightness coefficients is given for one road surface. Several conclusions about the requirements to be made of the light distribution of the sources are discussed on the basis of the distribution of the brightness coefficients.

In a previous article in this periodical [2]) it was pointed out that the distribution of brightness over the surface of a road is very dependent upon the nature of the road surface and its condition of dryness or wetness.

The divergent characteristics of road surfaces are more pronounced the smaller the angle at which they are observed. This means chiefly that great differences between the various kinds of road surface (and the various degrees of wetness) may be expected, when parts of the road are viewed which are relatively far away from the observer. These parts of the road are of extremely great importance in visibility on the road. It has been shown by various investigations [3]) that with a fixed lighting system the objects and persons on the road are usually only seen as dark silhouettes against the light background of the distant illuminated road surface.

For fast traffic an obstacle must be observed at a distance of from 75 to 100 metres. The part of the road which can serve as background in this case is therefore 100 to 150 metres in front of the observer. At a height of the eye of 1.50 metre, and with a level road, that part of the road is therefore observed at an angle of 0.5 to 1° to the road surface (at an angle Ψ of 89° to 89.5 to the normal to the surface).

The lighting installation must give a light such that at directions of vision for which Ψ = 89 to 89.5° the brightness of the surface of the road is distributed fairly uniformly. This requirement must be satisfied in dry as well as wet weather; in the latter case the difficulty is greatest because it is just at the above-mentioned angles of observation that the reflective properties of the road surface are very variable with the degree of wetness.

In order to show that at these angles of vision we may expect very remarkable effects, we shall first deal with a degree of wetness of road surface which fortunately cannot occur in the case of a great many roads, and in other cases only occurs for a short time. Afterwards we shall draw certain conclusions about the distribution of brightness on a road.

The road surface covered with a continuous layer of water

It is well known that a continuous layer of water on the road surface acts like a mirror surface. When one looks at puddles on the road which are only a few metres away from the observer, images are seen which are nowhere near as bright as the objects whose images they form. The greater the

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1) Most of the data in this article are taken from "Light Reflection by Road Surfaces" by J. Bergmans, dissertation, Delft, June 1938, published by Waltman of Delft. This dissertation is written in the Netherlands language, while a short resumé in English is given at the end of each chapter.

2) G. B. van de Werfhorst, Philips techn. Rev. 2, 239, 1937.

3) This was first remarked upon by Millar, I.E.S. Trans. 5, 653, 1910.

distance between the observer and the puddle, however, the brighter the images.

This practical experience corresponds exactly to the theory about the reflection at non-conducting surfaces (glass, water, stone, etc.). This type of reflection takes place according to Fresnel's formulae and may be illustrated in a simple way as is done in *fig. 1* for water.

From Fresnel's formulae it may further be seen that the coefficient of specular reflection is the same for two corresponding angles Ψ and u (see sketch of path of rays in fig. 1). In other words just as large a percentage is reflected when the light passes through the air and is incident on the water surface, as when it comes from the water and is incident on the lower side of the water surface, if the angles Ψ and u are corresponding angles.

This has been expressed graphically in a simple way by plotting as abscissae the angles u corresponding with Ψ. The graph thus indicates ϱ_Ψ as well as ϱ_u.

What is now the brightness of an inundated piece of road surface seen from a distance of 75 to 150 metres?

Let us consider *fig. 2*. From all possible directions (h, g) light may fall upon the surface of water. According to Fresnel's formulae more or less light will be reflected at the surface depending on the angle at which the light is incident on it. A large percentage of this light will penetrate into the layer of water, reach the road surface, be reflected there and again emerge into the air through the surface. Because of the fact that the reflecting power of the particles of which the road surface is built up decreases sharply when they are enclosed on all sides by water, only a small amount of light will be reflected by these particles into the layer of water.

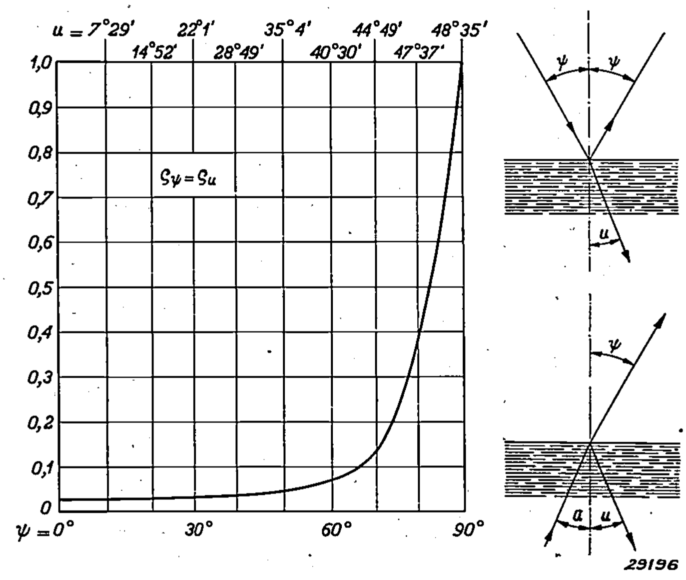


Fig. 1. Reflection coefficient ϱ_Ψ for light falling on a water surface at an angle Ψ as a function of Ψ. The same curve gives the reflection coefficient of a ray which strikes the water surface from below at an angle u, as a function of u provided the angles Ψ and u are corresponding angles (see sketch above the diagram). This is expressed by noting the angle u as abscissa at the top of the diagram.

For $\Psi \rightarrow 90°$ the reflection coefficient becomes equal to 1. If the light strikes the inner side of the water surface from below at such an angle u that the emergent ray would leave the surface in a practically horizontal direction, then it is also practically completely reflected and cannot emerge.

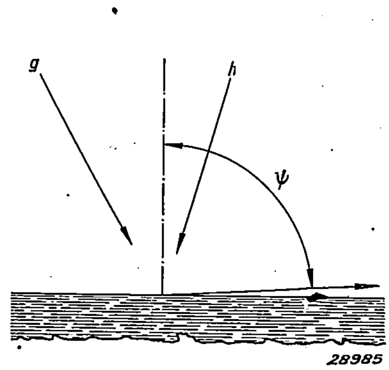


Fig. 2. In the direction of observation Ψ = 89° only that light is observed which has been reflected at the water surface, and thus satisfies the conditions of specular reflection. Light which has penetrated into the water is practically unable to emerge at an angle of 90° (see under fig. 1).

This decrease in the reflection of the particles of the road surface enclosed by water plays a much less important part than the second reflection of the light at the inner surface of the water boundary. The brightness of the surface in observation at a great distance is determined by the amount of light which is directed toward the eye of the observer and therefore which leaves the surface of the water at an angle Ψ = 89 to 89.5° (see fig. 2). It may be seen from fig. 1 that the reflection at the inner surface of the water is 90 per cent for an angle Ψ = 89°, and 95 per cent for Ψ = 89.5°.

It may therefore be seen that, of the already smaller amount of light reflected from the road surface, almost none of it can reach the eye of the observer at a great distance. Therefore no matter how strongly an inundated road surface is illuminated from all possible directions, when the angle of vision Ψ = 89 to 89.5° the road surface will always remain absolutely dark. The only way of making the surface appear bright is to have the light incident at an angle equal to the angle at which it is observed.

In this connection we should like to mention a very nice test which can be carried out with simple means.

A light-coloured slab f (see *fig. 3*) is provided with an inscription. At the same time a vertical board d with an inscription in mirror script is set up at the end of the slab. The whole surface of the slab is covered with water. We observe from the directions a, b and c which make angles

Ψ of 60°, 75° and 89°, respectively, with vertical.

Fig. 4*a*, *b* and *c* show what one sees at the directions of observation *a*, *b* and *c* respectively.

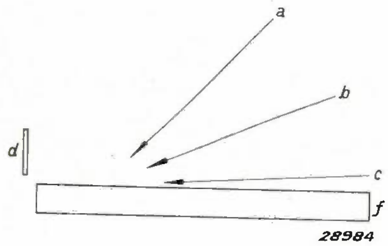


Fig. 3. Comparison of the reflective power of a wet slab upon observation from the directions *a* (60°), *b* (75°) and *c* (89°). The word STOP is written on the slab. On the board *d* there is an inscription in mirror script. Fig. 4 shows how the board is reflected better and better with increasing angle of incidence, while the inscription on the slab itself becomes vaguer and vaguer.

These photographs illustrate clearly how the road surface acts more and more as a perfect mirror upon increase of the angle Ψ.

Therefore although an inundated road surface does not at all possess the properties of a mirror for all possible directions of observation, it follows from the above explanation that when it is viewed from a great distance the road surface begins more and more to act as a perfect mirror.

Because it is just these directions of observation which are of primary importance in fast traffic, the warning contained in the twelfth conclusion, fifth question, of the 8th International Highway Congress (The Hague, June 1938) is very much to the point.

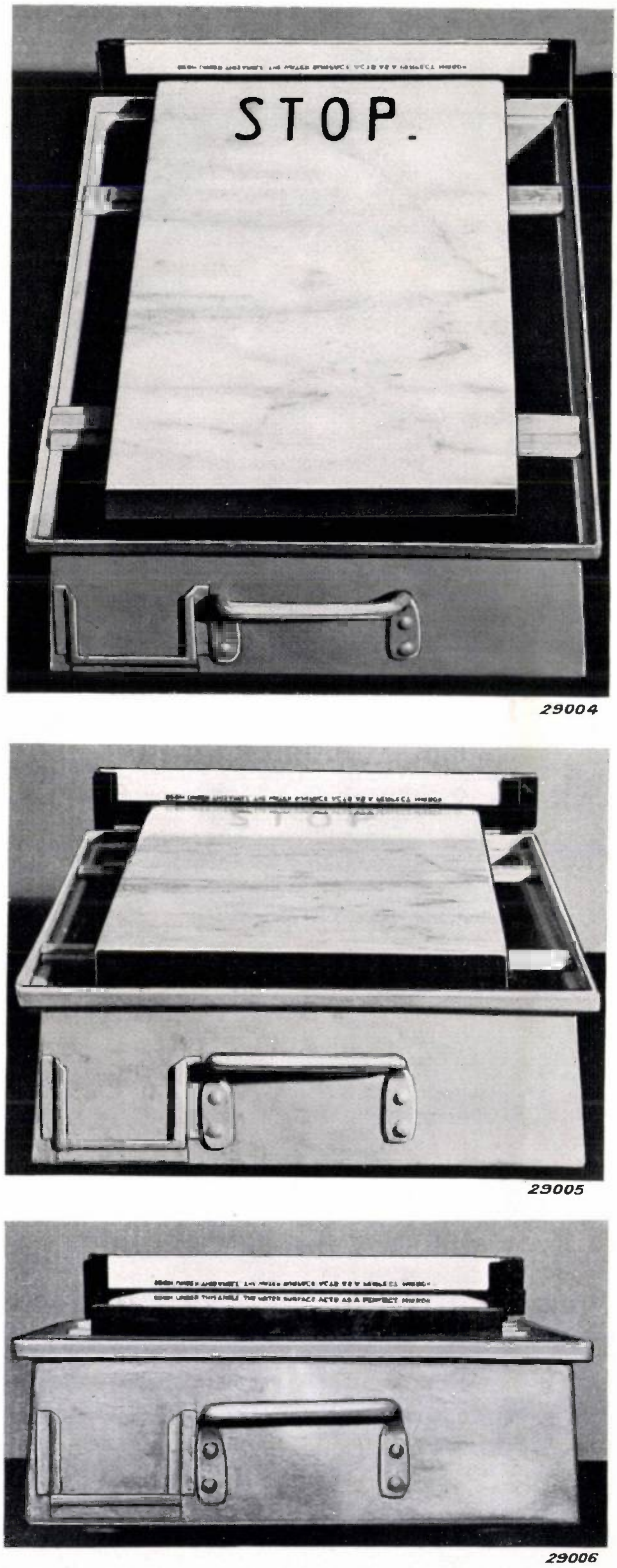
The conclusion was as follows:

12. *The road surface should be so constructed as to prevent so far as possible the formation of a water layer, which tends to make the road surface act as a mirror, producing excessively bright, narrow streaks, causing glare and thus reducing visibility.*

How does the road surface reflect when there is no layer of water?

We have seen above that an inundated road surface can never be quite satisfactorily illuminated, and we shall leave this least favourable case out of consideration in the following. Even with a dry surface, however, or with a surface which is damp but not inundated, the reflective properties which must be taken into account in planning a lighting installation are far from simple. Even with apparently diffusely reflecting road surfaces the intensity of reflection in a direction practically parallel to the road surface is found to deviate very much from Lambert's law.

In order to form some idea how the reflection takes place on a dry and a damp road surface, it is advisable to examine several kinds of road surfaces under a binocular mircoscope. Two types of structures which will be encountered are show nin *fig.* 5. From these and other types of surface which



a

b

c

Fig. 4. Reflection by a road tile in the arrangement shown in fig. 3. The three photographs refer to the directions of observation *a*, *b* and *c*.

we examined microscopically we obtained the following general impression.

Road surfaces are fairly rough. One of the main causes of the pitting of the surface is the breaking out of bits of stone. The elevations are in many cases places where a piece of hard stone lies directly under the surface or projects slightly above it.

Due to the pits and the elevations the light which falls on the road surface from a certain direction will be spread in many directions, even if it is reflected specularly by the individual particles of the road surface as is the case to some extent. Some particles themselves on the other hand exhibit a spread reflection [4]).

The latter particles usually appear grey in colour; the light penetrates into them to some extent, and its direction is so changed that the reflection is in many cases almost completely diffuse.

When such a piece of road surface becomes damp, a layer of moisture is drawn over all the component parts of the fine structure. This makes the spread reflection of many particles less, and specular reflection according to Fresnel's formulae occurs at the surface of all the particles (see fig. 1).

When the amount of water on the surface is increased there will be further changes. The water films become thicker and begin to fill up the spaces between the projecting particles. The water film thus no longer follows the outline of the particle but rounds it off, so that the radius of curvature of the water surface is greater than that of the original particle of the road surface structure. The normals which may be drawn to this rounded-off water surface deviate, however, much less from the normals to the road surface in the dry condition. In this way the well known shiny effect is obtained: more reflection with respect to normals near vertical and less with respect to the normals with a great deviation from vertical.

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Fig. 5. Photomicrographs (enlarged 20 times) of road surfaces which have been in use.
a) Highly coarse asphaltic concrete with coating layer of "sand-clinker sheet", taken from the road between the Hague and Delft. This road surface shows many small transparent crystals which are not clearly seen in the photograph.
b) Fine asphaltic concrete on binder coarse taken from the road between Amsterdam and Haarlem. The road surface is dark and fairly smooth, aside from pitting and grooves.

It is clear therefore that with regard to its reflective properties the road surface in the dry and the damp condition is intermediate between the completely diffusely and the completely specularly reflecting surface.

We shall now attempt to characterize this state of reflecting power more closely by beginning with the two limiting cases, namely, completely diffuse and completely specular reflection.

Characteristic difference between completely diffuse and completely specular reflection

In the illumination of a completely diffusely reflecting surface the brightness which is reflected by that surface in any given direction is proportional to the illumination.

Let ϱ_d be the reflection coefficient of such a surface (ratio of the total reflected to the total luminous flux). Then the formula for the quotient q_d of the brightness B' (in c.p./sq.m*) to the illumination E (in lux) may be written as follows:

[4]) We use the word "spread reflection" to indicate an intermediate condition between completely diffusing and completely specular reflection, when the intermediate state may not be described as: a certain percentage of the light completely diffused and a certain percentage completely specularly reflected.
An example for which the latter description is practically accurate is that of milk glass with a smooth surface. Most surfaces met with in practice having a matt lustre fall in the group which we here call "spread reflecting".

*) We have used "c.p./sq.m" and not the "stilb" (c.p./sq.cm) because:
1) The values, which are got for the brightness coefficients and which are already very low, increase with a factor 10^4;
2) in formula as *e. g* : $E = B \cdot \omega \cdot \cos \Psi$ the facter 10^4 disappears.

$$q_d = \frac{B'}{E} = \frac{\varrho_d}{\pi} .$$

The quotient q_d is called the **brightness coefficient for diffuse reflection.** For a perfectly diffusing surface it is a constant value with all combinations of directions of illumination and observation, and with all possible variations in the dimensions of the light source.

For a completely specular reflecting surface the quotient of the brightness to the illumination cannot be so simply given.

The specularly reflecting road surface is dark everywhere except in the direction of observation in which the mirror image of the light source is seen. The brightness B' of this mirror is determined only by the brightness B of the source, but is independent of the illumination on the road. The illumination E is not determined by the brightness of the source, but, at a given brightness of the source, it is proportional to the solid angle ω (see *fig. 6*) within which the light source is seen

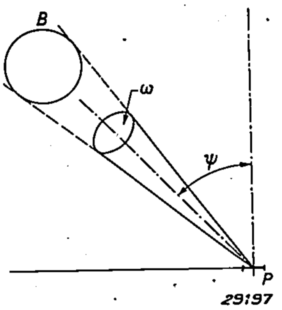


Fig. 6. The intensity of illumination of the point P on the road surface is $E = B\,\omega \cos \Psi$.

from a point on the road surface. The brightness coefficient therefore B'/E is not a property of the road surface alone but is inversely proportional to the solid angle ω for every point on the road surface, and therefore really has no significance in the characterization of the specularly reflecting road surface.

Spread reflection

Since spread reflection is an intermediate condition between diffuse and specular reflection, it may be expected that here also the brightness coefficient will not depend upon the direction of illumination and observation alone, but also upon the solid angle within which the light source is seen from a point on the road surface. It must here be assumed that the light sources continue to give mirror images — although not very sharp ones — in the spreading surface, and that the brightness of these mirror images increases when the light source (while retaining the same distribution of light) is decreased in size. When the apparent size of the light source becomes smaller, however, than the lack of sharpness of the mirror image, further decrease in size will give practically no change in the mirror image, and the brightness of the mirror image will then no longer increase.

For light sources smaller than this limit, there is therefore again a well defined brightness coefficient which depends only upon the directions of illumination and observation, and which is independent of the size of the light source.

It is clear that this limit will be the more quickly reached, *i.e.* will occur with a larger light source, the stronger the spreading of the reflected light.

A road surface which has a well-defined brightness coefficient with the ordinary dimensions of the sources of light for street lighting, is considered widely spread reflecting.

From measurements carried out by the author [5]) it was found that road surfaces in dry and moderately damp conditions may be considered widely spread reflecting surfaces. This is very important, because for that reason it has become possible to determine a value of q for every combination of directions of illumination and observation, and to discuss the properties of a road surface with the help of such a relation.

Results of measurements

In *fig. 7* are given the lines of constant brightness coefficient for the dry condition (continuous lines) and for the reproducible damp condition (dotted lines) measured by us of the road surface of the road between the Hague and Delft (see photomicrograph fig. 5*a*).

In this drawing the camber of the road has been taken into account. The camber of modern roads is no longer rounded, but has more the shape of a roof. From the middle of the road the surface slopes toward both curbs with a slope of 1 : 50. Because of this a discontinuity in the course of the iso-q

[5]) See the dissertation of the author, page 80. It was shown by experiment that the distribution of brightness of dry road surfaces does not change if the measurements are carried out first with a light source of such dimensions that the ratio: diameter of light source to height of mounting corresponds to the maximum ratio which ordinarily occurs in practice, and then with a light source which is 250 times brighter and whose solid angle ω is thus $^1/_{250}$ of the solid angle used in the first measurement. In order to carry out the same measurement on wet road surfaces a slab of the pavement was cut out and moistened in a definite reproducible way. The condition obtained in this way corresponds, according to our estimation, with the wetness which a road surface can retain for hours when there is a slight rain or heavy mist.

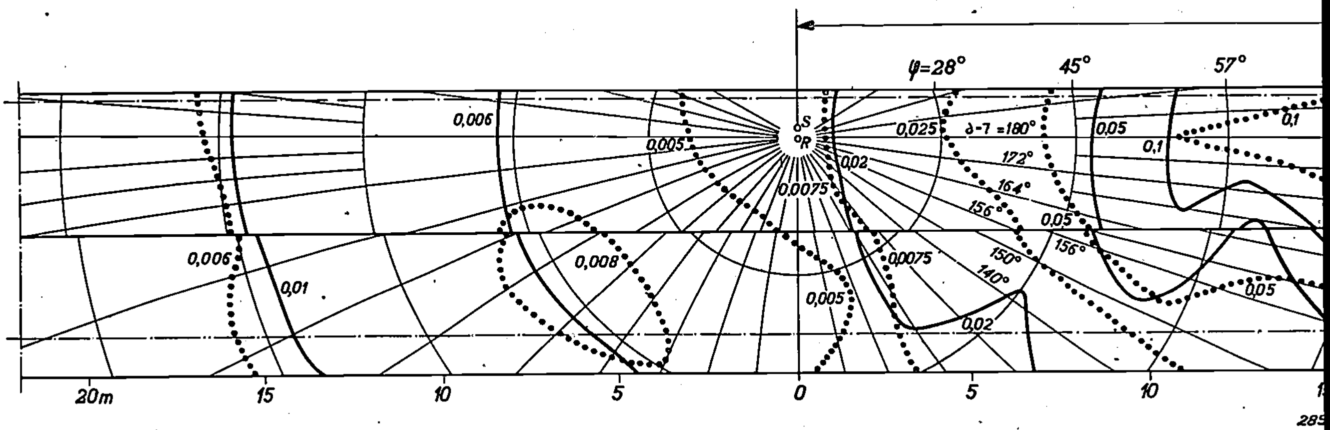


Fig. 7. Curves of equal brightness coefficient (iso-q curves) for a roof-shaped road (slope 1 : 50) 8 metres wide. Height of the light source 8 metres above the curb. The eye of the observer is at a distance of 100 metres from the lightsource and

curves may be seen at the axis of the road surface. The q values for different points of the surface, *i.e.* for different combinations of directions, are very divergent. In the diagram for the dry road iso-q curves may be seen for values of q from 0.006 to 1, while the q values for the damp road vary between 0.005 and 10.

Before we draw any conclusion from these iso-q curves about the brightness distribution with different light sources, it is first necessary to find out to what extent these iso-q curves are typical, not only for this road surface, but for all road surfaces. We were able to carry out similar measurements on brick, asphaltic concrete, granite setts and cement concrete road surfaces.

When the values from the measurements on all these different types of road surface are compared for the same combination of directions, a much smaller mutual variation is found than that for different combinations of directions measured at the same point of a single road surface.

On the basis of the figures obtained we may make the following statements.

In the dry condition the values of q for the same combination of directions for two points of different road surfaces are found to deviate from each other by not more than a factor of four.

On the other hand, also in the dry condition, the values of q for the same point on the road surface and for different combinations of directions, even with the most diffusing surface differ by a factor of more than fifty.

In the damp condition these values become greater and differences by factors of 8 occur between the different road surfaces. The different direction combinations of the same road surface then however differ very much more (by more than a factor of 2 000).

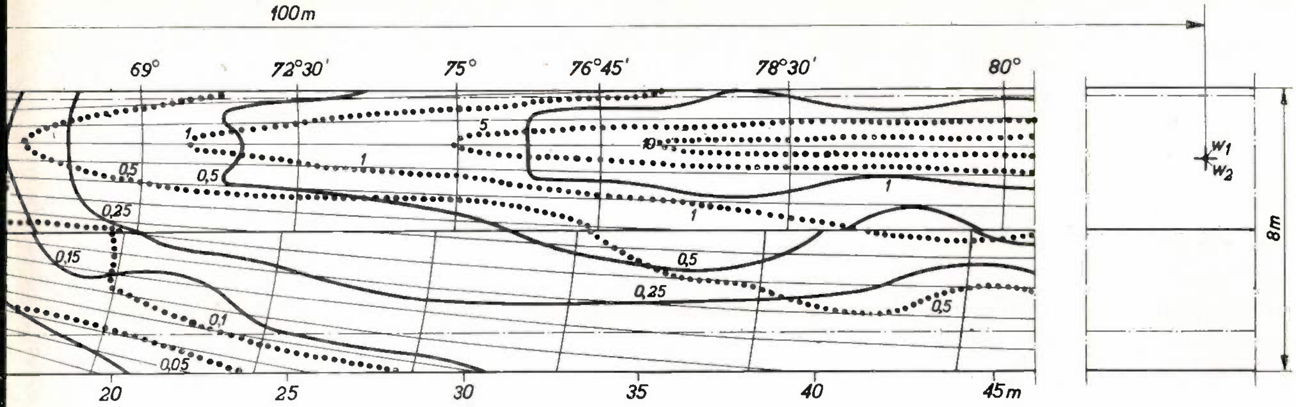
There is quite good correspondence between the iso-q curves of the different road surfaces. It is, however, not so exact that it is only necessary to provide the set of curves for one road surface with higher values in order to obtain the curves for another road surface, but it is such that the iso-q curves given in fig. 7 may be considered typical curves of a road surface in the two conditions of wetness measured by us. The surface referred to in fig. 7 was the darkest of all road surfaces which were measured. On an average the brightness coefficients have values which are higher by a factor of 2.5 [6]).

From the measurements on road surfaces in dry and damp condition which are the condition usually occurring, we may draw a general conclusion.

The isolux curves are still very often used as a basis for estimation in planning road illumination. It is then tacitly assumed that the illumination and the brightness observed by the road user have some constant relation to each other. The uniformity ratio of the strongest to the weakest illumination is considered by many persons, who must decide upon the carrying out of one system of illumination or another, as a ratio which, while perhaps not exactly corresponding to the ratio between the greatest and the least brightness, may nevertheless very well be used as a first approximation of this ratio. It is thus assumed that the value of q will be about constant.

From the fact that the value of q for dry road surfaces varies by a factor of more than 50 for

[6]) These relatively small differences between the types of road surfaces do not agree with the experience which we have of the behaviour of road surfaces during and after rain. *It must be concluded that road surfaces show the greatest relative differences when they are examined as to the amount of water which can be drained away per minute without any closed layer of water being formed.*



at a height of 1.50 metre above the surface of the road. Continuous lines indicate the dry condition, dotted lines the damp condition. The q values are measured on a section of the pavement of the road between the Hague and Delft.

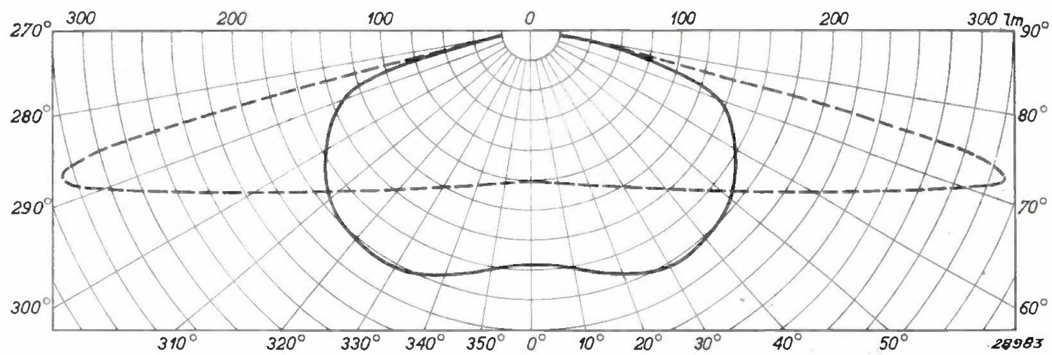
different direction combinations, and that in the damp condition this factor becomes greater than 2 000, *it is perfectly obvious that neither the isolux curves nor the uniformity ratio of the illumination have the slightest value in obtaining an impression of the distribution of brightness over the road surface.*

What practical use may be made of the brightness coefficients?

The illumination is the quantity which indicates the luminous flux consumed per square metre of road surface; the unit lux is one lumen per square metre.

Because the brightness coefficient indicates the relation between the brightness of the road surface and the luminous flux thereby consumed per unit of surface, this coefficient might be considered to be a measure of the efficiency of the road surface. This would be correct if the only purpose of road lighting were to raise the brightness of the surface as high as possible, and if the distribution of brightness were only of secondary importance.

For vision on the road, however, it is usually the other way round, and on a road where the average brightness is relatively low, but where the distribution of the brightness is good, it is usually easier to see obstacles than on a road where the average brightness is much higher and its distribution less uniform.



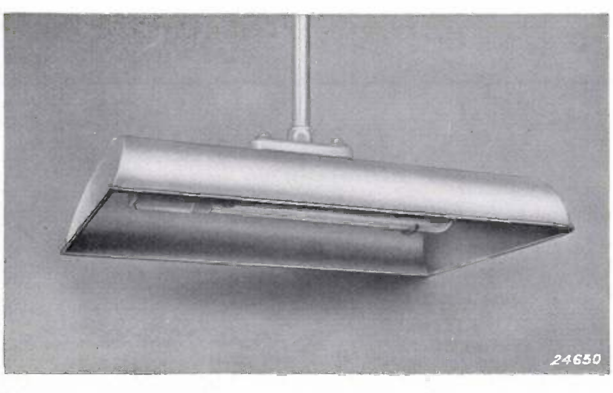


Fig. 8. Photograph (left) and curve of the light distribution (above) of the "Philora" fitting SORA 61001 for sodium lamps. This light distribution forms a compromise between the requirements of high efficiency and good uniformity of illumination. The dotted curve represents a less satisfactory light distribution.

When in order to obtain a greater brightness of the road surface more use is made of the higher brightness coefficients, it means that relatively more luminous flux is directed toward those points of the road surface for which the value of q is high. This will in most cases be the points which already possess a great brightness, so that the uniformity of brightness is thereby diminished. A relatively high luminous flux must therefore deliberately be directed toward points of the road surface with a relatively low brightness coefficient in order to avoid too great non-uniformity of brightness.

For the same reason great care must be taken in directing much light toward those points where the value of q varies very much in the damp and dry conditions. These are the points which lie between the light source and the observer at a relatively great distance from the source (see fig. 7).

Because of these requirements the distance which can be illuminated with a given height of the light source is limited. The avoidance of glare by the direct light from the light sources also necessitates a similar restriction in the length of road which can be served by each light source.

It is clear that important conclusions may be drawn from the distribution of the brightness coefficient in the dry as well as the damp condition, in connection with the height and spacing of the light sources. We shall return to this point in a later article.

The question of the degree of non-uniformity in brightness which is permissible is also a problem by itself. In dealing with this problem one must begin with the road seen in perspective, and with the dimensions, form and reflective power of the obstacles which must be able to be observed on the road surface. We intend also to make a further study of these perspective considerations.

A further study of what may or may not be permissible is also important because the reflective properties in the damp and dry conditions differ so much that one is never able to design an installation which has a satisfactory brightness and with which in the damp as well as in the dry condition the surface is seen about uniformly bright.

Discussion of the properties of a special fitting on the basis of known brightness coefficients

In *fig. 8* may be seen a photograph and the light distribution curve of the above-mentioned Philips fitting for sodium lamps "Philora"-SORA 61001. This fitting is usually mounted at the curb of the road on a bracket.

In *fig. 9* the distribution of the brightness is indicated for this fitting as observed from a point 100 metres from the pole (see fig. 6). The continuous curves (r, s and t) indicate for the dry condition respectively the brightness over the line joining the point directly under the light source with the position of the observer, the axis of the road and the boundary line of the road surface on the side opposite the light source. The dotted curves (u, v and w) indicate the brightness in the damp condition for the same lines on the road surface.

It may be seen from this figure that the fitting contributes very little to the brightness of those parts of the road which are farther away from the observer than the light source itself. This is illustrated most clearly in the damp condition. In the dry condition also, however, the brightness of points on the road which are more than 7 metres farther away from the observer than the light source itself may practically be neglected.

At a distance of about 28 metres from the light source on curve u (damp condition) is the point with the greatest brightness. At the same distance the brightness of the lines r, s and t is already considerably lower than the maxima occurring on these same three lines (dry conditions).

Various fittings have been put on the market which direct much more light (three or four times the amount given by the "Philora" fitting mentioned above) toward these points farther away from the source. Distributions of light are then obtained as

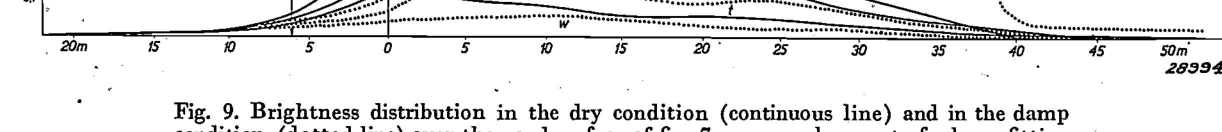


Fig. 9. Brightness distribution in the dry condition (continuous line) and in the damp condition (dotted line) over the road surface of fig. 7 upon employment of a lamp fitting according to fig. 8. The curves r and u refer to the line joining the point directly under the source and the observer; s and v refer to the axis of the road; t and w to the boundary line of the road on the side opposite to the light source.

indicated in fig. 8 by the dotted line. From fig. 9 it may be seen that this offers no difficulties for the dry condition (aside from the glare of the direct light in the eyes of the observer): the curves r, s and t may well be raised to three times their value at the cross section at about 28 metres from the light source without the brightness becoming disturbing for observation. For the damp condition, however, this is no longer the case. A brightness is then obtained which is about 40 times as great as that of any point lying on the side of the road opposite the light source. Such a great lack of uniformity is very disadvantageous to visibility on the road.

It may therefore be seen that with the fitting SORA 61001 a compromise has been made between, on the one hand, the greatest possible use of the higher values of the brightness coefficients, and, on the other hand, the avoidance of too great local brightnesses in the damp condition.

Higher illumination on the opposite side of the road

Until now we have spoken of light sources which distribute the light as well as possible along the lengthwise direction of the road. It was found that too much light must not be directed towards points on the road surface which lie on transverse cross sections at about 28 metres from the light source.

If now we consider the point in that cross section which lies on the opposite side of the road (see fig. 7) it will be seen that this point, even in the damp condition, does not have a high nor, fortunately, a very low brightness coefficient. Its value (0.1) even corresponds fairly well with the value in the dry condition.

Therefore when we conclude that for these greater distances from the light source the luminous intensity over the full width of the road must not be raised above the value which is given by the sodium lamp fitting SORA 61001, it does not mean necessarily that the luminous intensity toward points of the surface on the opposite side of the road may not be raised.

This fact offers a possibility for the improvement of the uniformity of brightness in the dry as well as the damp condition.

THE ELECTRO-ACOUSTIC INSTALLATION IN THE LEAGUE OF NATIONS PALACE IN GENEVA

by N. A. J. VOORHOEVE and J. P. BOURDREZ.

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A survey is given of the system for the distribution and recording of sound which has been installed in the League of Nations Palace in Geneva. All components as far as possible have been housed in a central amplifier station, although this necessitated a greater number of connections than would have been necessary with a less centralized system. The advantage of the system used lies in the fact that it can more easily be operated.

Introduction

Among the assembly-rooms in the new League of Nations Palace on the Lake of Geneva there are two which had to be provided with electro-acoustic installations because of their great size, and for the sake of the publicity of the meetings to be held in these halls. They are the huge general meeting hall of the League of Nations: "Salle des Assemblées", the dimensions of which are 59 by 51 by 20 metres with seats for about 1550 persons, and a smaller assembly hall which is also equiped for cinema performances: "Salle des Commissions 12/13", with dimensions of 24 by 12 by 10 metres and about 400 seats.

The following objects had to be attained in a simple and reliable way with the apparatus to be installed:

1) Amplification of the intensity of speech in both of the rooms by the use of loud speakers, since the voice of a speaker cannot be heard sufficiently well in all parts of these rooms [1]).
2) Transmission of the speeches, not only to the League of Nations transmitters, but also to all other transmitting stations. The possibility

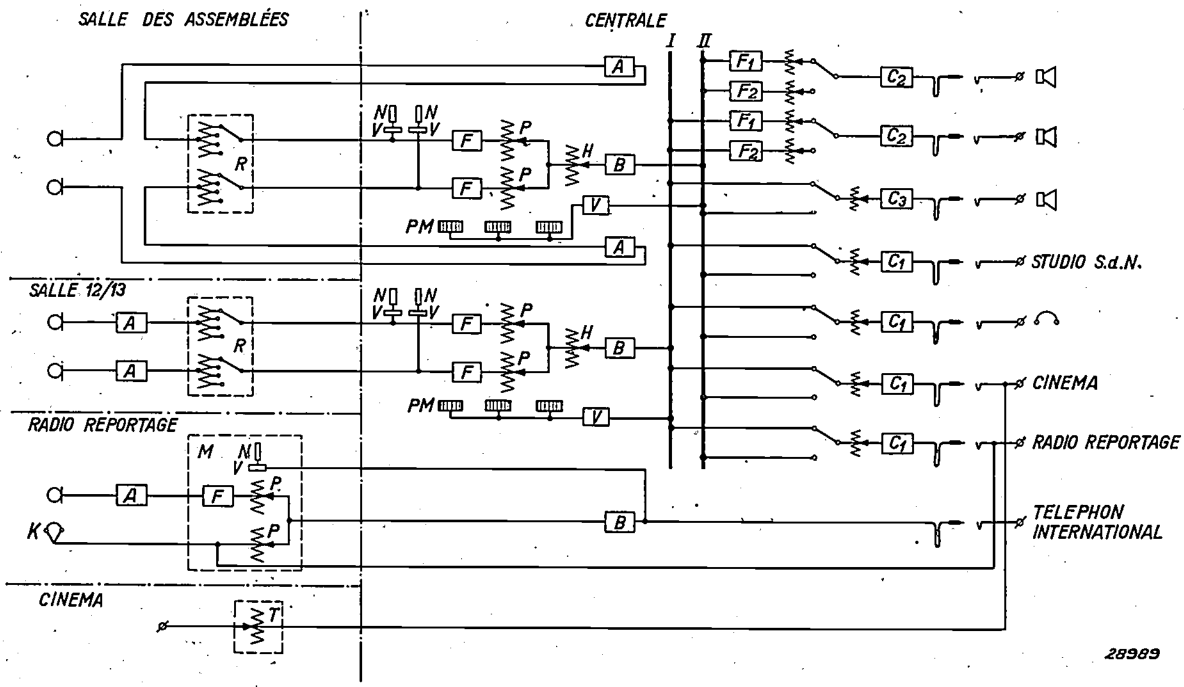

Fig. 1. Simplified diagram showing the principle of the amplifier installation (low-frequency part).

A	= microphone pre-amplifier	N	= neon indicator
B	= intermediate amplifier	PM	= programme meters
C_1	= separation amplifier	M	= mixing cabinet for radio reporting
C_2	= power amplifier	R	= poste de régie
C_3	= monitor amplifier	T	= cabinet for making sound film connections and connections with other apparatus for sound recording.
F, F_1, F_2	= filter		
H	= main volume regulator		
P	= mixing potentiometers	V	= amplifier for neon indicator and programme meter

1) For further measures taken with the aim of improving the acoustics in the general meeting hall *cf.* Philips techn. Rev. 3, 159, 1938.

had to be taken into account that different meetings might take place simultaneously in the two rooms.

3) Reproduction of the speeches by means of loud speakers in a great many other rooms of the building.
4) Radio reporting of the meetings, for which purpose a number of cabins were built in each of the rooms.
5) Recording of the speeches, either on sound film, when cinema films of the meeting are taken in the cabins for that purpose in each room, or by means of various apparatus for sound recording set up separately.

In this article several explanatory descriptions will be given, as well as particulars of general interest in installation technique about the electrical and mechanical execution of this installation.

The installation

In planning the project we were faced with a choice of two possibilities, either to built a separate system for each room or to combine the necessary apparatus for both rooms in one central control room. The latter solution was chosen in spite of the fact that a decentralized arrangement of the apparatus would have been easier to carry out technically, because of the great distance of 120 metres which separates the two rooms in the building. Centralized arrangement, however, offered, in addition to the most economical use of the necessary amplifiers, operating and control organs, the advantage that a minimum number of technically trained persons are necessary for operating the apparatus. These arguments more than make up for the necessity of certain special requirements of the cable connections due to the great distances to be bridged and the introduction of a simple signal arrangement which became necessary because the personnel responsible for the transmission is not stationed in the room itself. These two points are dealt with later on in this article.

In a very simplified form *fig. 1* gives the principle of the low frequency part of this transmission installation. It may be seen that only the indispensible parts of the installation are situated in the rooms themselves. To the left of the broken line are indicated the pieces of apparatus situated in the rooms; to the right of this line is the apparatus set up in the central station.

In both rooms a number of microphones of the table type (five in the "Salle des Assemblées", four in the "Salle 12/13") are installed on the speakers desk (*fig. 2*), and they correspond aesthetically with the decoration of the halls. In connection with the requirement that the quality of the reproduction may leave nothing to be desired ribbon microphones have been chosen.

Fig. 2. Speakers rostrum in the "Salle des Assemblées" upon which the five ribbon microphones may clearly be seen.

Since ribbon microphones are not especially sensitive, difficulties might arise from interferences, if the relatively low output voltage of the different microphones should be regulated and mixed directly by means of potentiometers. It may be seen that pre-amplifiers (A) are first placed directly behind the microphones, which raise the level so high that the operation of the mixing potentiometers (P) can no longer cause audible disturbances.

As regards the microphones set up in "Salle 12/13", another possible source of disturbance must be avoided, namely the danger that audible interference voltages might be generated in the 120 metre long connection between these microphones and the amplifier station which is next to the "Salle des Assemblées". Such voltages might be caused by interference fields from other electrical connections in the building.

This difficulty could only be met by raising the level of the microphone voltages already at the beginning of this connection to a sufficient height, and, for this purpose, the pre-amplifiers and their corresponding supply apparatus for anode voltage and heating current for the four microphones in "Salle 12/13" are set up in the room itself. These are the only low-frequency amplifiers which are not situated in the central amplifier station. The principle of central technical control was retained, however, by arranging to have these amplifiers switched on from the control room by means of press buttons and corresponding automatic switches. For the sake of necessary reliability of working a group of four additional reserve amplifiers were introduced to replace a group which might be out of working order and they can be brought into action by means of a press button in the control room.

Since the switching on and off of the microphones must take place at a spot where it is possible to follow the proceedings of the meeting, a so-called poste de régie (*fig. 3*) is set up in each room in

Fig. 3. Poste de régie in the "Salle des Assemblées". At the front of the control table may be seen the knobs for regulating the potentiometers R, and at the upper right-hand side the four switches for regulating the sound intensity of the loud speakers. Above these are the neon indicators which indicate the sound intensity at different spots in the hall (see the last section: signaling).

which are the potentiometers indicated by R in the diagram (fig. 1). These potentiometers serve solely to make the switching on and off of the microphones inaudible, and volume regulation and mixing do not take place at the poste de régie. Moreover there is a telephone as well as switches and lamps for signaling purposes at the poste de régie. These will be mentioned later on in this article.

The pre-amplified, low-frequency voltage is conducted to mixing panels in the control room, which panels form part of the central operation and control table which may be seen in *fig. 4* at the right.

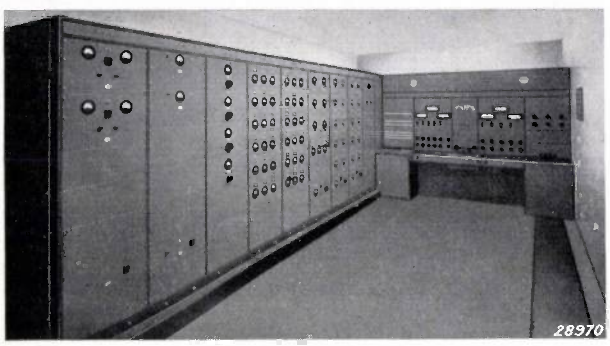

Fig. 4. View of the central amplifier station. In the background may be seen the control table on which the two mixing panels with their programme meters may clearly be distinguished.

There are two of these mixing panels, one with 4 and one with 5 channels (in fig. 1 only 2 channels of each mixing panel are drawn). In each channel there is a filter F and a potentiometer P with which respectively the low tones and the intensity of each microphone are regulated. For the sake of simple and rapid operation a neon indicator N is added for every channel. This is supplied by a special amplifier V and indicates whether or not there is low-frequency voltage on the line. In this way the time-consuming switching over of the head phones by the operator is avoided.

The modulation thus mixed is conducted from each mixing panel *via* the main volume regulator H and the intermediate amplifier B which raises the low-frequency voltage to the level of the line, to the main distribution lines, I and II respectively. To these main distribution lines all the components necessary for the realization of the objects mentioned in the introduction can be connected as desired. This is done in the case of those lines on which line level is required with the interposition of a separation amplifier C, in each line. These are the international telephone lines, the lines to the League of Nations Studio (Radio S. d. N.), to the cinema cabins, to the reporting cabins, to the apparatus for sound recording and the internal distribution lines. In this way the reliability of working is high, since the outgoing lines cannot react upon each other, and therefore in the case of disturbances in one of the lines, the others may remain in use with no difficulty. Moreover with the amplifiers C_1 extra line losses may be compensated for. The power amplifiers C_2 which supply the power loud speakers are connected directly with the main distribution lines without intermediate connection of amplifiers C_1.

Furthermore so-called monitor amplifiers C_3 are introduced, which make it possible to check the quality of the reproduction in the control room. For accurate control of the modulation, a set of so-called programme meters PM (fig. 1) is introduced on the control table for every pair of distribution lines. Each set of meters consists of three measuring instruments which indicate respectively the average voltage, the minima and the maxima, which occur on lines I and II during the transmission.

The lines to the reporters' cabins (7 in the "Salle des Assemblées" and 2 in "Salle 12/13") end in the respective cabins on mixing panels (see M in fig. 1). The reporters listen with head phones K to what is said in the hall, and they speak into a ribbon microphone (see *fig. 5* interior of reporters' cabin). The modulation given by this microphone is first amplified by a microphone pre-amplifier A, after which, *via* a filter F in the mixing panel, it can be regulated by the volume regulator P, and at the

Fig. 5. Cabin for radio reporting. The reporter can follow the proceedings of the meeting through the window. The microphone is so connected that it can be moved in all directions inside the cabin.

same time mixed with the modulation coming from the hall. In the case of "Salle 12/13" these pre-amplifiers are placed in the cabin itself, because of the great distance between microphone and control room, while for the "Salle des Assemblées" they are housed in the central amplifier station.

In the cabins for taking cinema films (7 in the "Salle des Assemblées", see *fig. 6*, and 2 in "Salle 12/13"), the lines over which the modulation is conducted end at connection cabinets T (*cf.* fig. 1). Each cabinet contains a set of contact terminals upon which the cinema operator can tap off the modulation in order to record it on the film. In order to make the adaptation as suitable as possible for the most commonly used recording apparatus, a constant attenuation of 20 dB is introduced into each cabinet, while an adjustable attenuation of 30 dB at the most can still be added with a potentiometer. The cinema operator therefore may avail himself of an adjustable level which lies 20 to 50 dB below the line level, which is adequate for all cases which may occur. The same cabinets are also used for the connections with other apparatus for sound recording.

Fig. 6. View of the gallery for taking cinema films in the "Salle des Assemblées". The seven connection cabinets (T in fig. 1) for the recording of sound film may be seen on the wall.

The internal lines for sound distribution which, as we have seen above, are also supplied *via* amplifiers C_1, include thirty contact points throughout the building, where line voltage is therefore also available. To these points, which are of the usual type with plug-in contacts, amplifier units with built-in loud speakers are connected.

In the case of the power amplifiers C_2 which supply the loud speakers for the improvement of the acoustics and several other loud speakers which are placed in neighbouring halls and rooms, several special measures have been taken to regulate the frequency characteristic. For the improvement of the acoustics it is necessary to adapt the reproduction accurately to the acoustic characteristics of the hall in question, in order to obtain the greatest intelligibility and the best quality. Therefore for each hall special permanently adjusted filters F_1 were designed, and those amplifiers C_2 which serve to improve the acoustics are then automatically connected to these filters when they are connected to the outgoing line to the corresponding loud speakers in the hall. When they have to feed other power loud speakers a filter F_2 is automatically put in series, with which filter the frequency characteristic can be regulated as desired. In this way the principle was retained which holds for all amplifiers C_1 and C_2, namely that in each group they are all mutually entirely equivalent for the various purposes.

Amplifier levels

In such a large amplifier system with which the energies generated by a number of different voltage sources must be amplified, regulated and then distributed over a number of different reproducing apparatus, it is better not to have the total amplification between source of voltage and reproducing apparatus take place in one step. Care must be taken in the amplification that the energies are amplified in the first instance to such a degree that they may later be regulated and mixed without danger of introducing disturbances. Then the energies thus collected into only a few groups (line programmes) are amplified to a level suitable for the reproduction apparatus. By constructing the installation in this way it can be built up of the necessary number of similar, and therefore interchangeable, parts quite independent of the nature of the sources of voltage and the apparatus for reproduction.

We shall now examine how this principle is applied in the amplifier installation of the League of Nations Palace. As supply level for the reproduction apparatus has been chosen the so-called line level which is present on the main distribution lines *I* and *II* (see fig. 1), and which is so chosen that the average effective value of the low-frequency voltage corresponds with an energy of 6 mW. The height of the other levels occurring in the installation will be expressed with respect to the thus defined line level.

In determining the various amplifier stages we begin with the ribbon microphones used, which at a sound pressure of 1 dyne/sq cm give an energy which is $10^{-7.4}$ times the energy of 6 mW of the line level. In other words: this ribbon microphone has a sensitivity of -74 dB with respect to the line level for a sound pressure of 1 dyne/sq cm. Most intensity maxima in normal speech, 100 cm away from the microphone, give a sound pressure of about 1 dyne/sq cm, while the average intensity of the sound is about 15 dB lower than this, and in several seldom occurring short maxima an intensity about 5 db higher is reached. Practical measurements carried out in the Philips Laboratory gave the above results, which agree very well with the data published by Fletcher [2]).

According to this author the maximum energy of the spoken word is about 1 mW. We may compare this figure with our measurements, and, starting with the sound pressure p in dynes/sq cm, we can calculate the power W of the source of sound.
W occurs in the formula:

$$W = \frac{4\pi r^2 p^2}{\varrho c} \text{ erg/sec} = \frac{4\pi}{\varrho c} r^2 p^2 \, 10^{-7} \text{ watt.}$$

where ϱ is the density of air, c the speed of propagation of sound, and r the distance from the microphone to the speaker. At a distance $r = 100$ cm from the speaker the sound pressure is 1.8 dyne/sq cm for the maxima of the spoken word. This is 5 dB higher than the average sound pressure of the spoken word (1 dyne/sq cm).
If therefore $r = 100$ cm, $\varrho = 1.8$ dyne/sq cm and $\varrho c = 42$, we find:

$$W = \frac{4\pi}{42} 100^2 \cdot 1.8^2 \cdot 10^{-7} = 10^{-3} \text{ W} = 1 \text{ mW},$$

which agrees exactly with Fletcher's result.

In the case of speech in a hall the situation is somewhat different. In order to collect experimental data on this subject, intensity measurements were carried out in 1937 with a temporary amplifier installation with different speakers in the "Salle des Assemblées". The average intensity was in most cases found to be 5 to 15 dB higher than in a personal conversation, and in one case even 20 dB higher. This also agrees perfectly with the observations of Fletcher, who states that the sound intensity of loud shouting lies 20 dB above that of an ordinary conversation (*i.e.* it is a hundred times as great), and that certain persons, even in a personal conversation, reach an intensity of as much as 9 dB above the average. With a minimum distance of 70 cm to the microphone, therefore the highest maximum which occurs in the intensity lies about 23 dB higher than the sound intensity of an ordinary conversation with a distance of 1 m between the speakers.

Let us now consider the case of a soft speaker, whose intensity falls in the course of his speech to $^1/_5$ of the normal intensity, and who moreover stands at a distance of 2 m from the microphone. The level which he produces in the amplifier has only $^1/_{16}$ of the normal intensity and thus lies 12 dB lower.

With respect to the line level of 6 mW, the commonly occurring maxima therefore lie at
$-74 + 23 = -51$ dB for the loudest speaker,
$-74 - 12 = -86$ dB for the softest speaker.

As was mentioned previously, in normal speech the average intensity lies about 15 dB lower than the commonly occurring maxima, therefore for the case of the soft speaker considered this becomes a level of 6 mW.

The average sound level of the soft speaker and the maxima of a very loud speaker thus lie about 50 dB apart, and it is required of the amplifier system that it shall deal satisfactorily with these two sound energies which differ by a factor of 100 000. The maxima in the softest speech must be brought to the line level of 6mW by the amplifier system. A maximum amplification of 86 dB is thus required. A total weakening of about 14 dB occurs in the fil-

[2]) Fletcher, Bell system techn. J. **10**, 349, 1931.

ters and sound mixers, so that the amplifiers A and B together must amplify about 100 dB. 50 dB has been chosen as the amount for each amplifier stage.

Behind the mixing panel therefore the intensity at the peaks would vary from -50 to -15 dB, and behind amplifier B from 0 to $+35$ dB, if no volume regulation took place. It is true that the amplifier B has only 1 per cent distortion at $+10$ dB, above this, however, the distortion increases quite rapidly (*cf.* fig. 8). It therefore seems desirable, and it has actually been shown by audition experiments, that care must be taken that the maximum peaks occurring do not exceed 10 dB. From this it follows that the mixing potentiometers and the main volume regulator must be able to give an attenuation of at least 25 dB.

Fig. 7 shows how the energy levels vary in

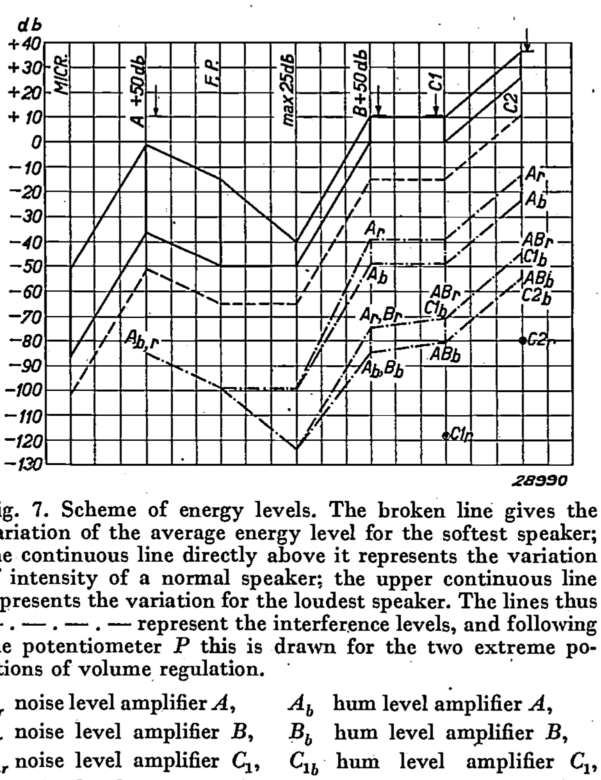

Fig. 7. Scheme of energy levels. The broken line gives the variation of the average energy level for the softest speaker; the continuous line directly above it represents the variation of intensity of a normal speaker; the upper continuous line represents the variation for the loudest speaker. The lines thus —.—.—.— represent the interference levels, and following the potentiometer P this is drawn for the two extreme positions of volume regulation.

A_r noise level amplifier A,	A_b hum level amplifier A,
B_r noise level amplifier B,	B_b hum level amplifier B,
C_{1r} noise level amplifier C_1,	C_{1b} hum level amplifier C_1,
C_{2r} noise level amplifier C_2,	C_{2b} hum level amplifier C_2,
F filter,	P mixing potentiometrs.

The amplifiers A, B, C_1 and C_2 have a distortion of less than 1% at the levels indicated by an arrow.

the different stages of the whole amplifier system, while the position of hum and noise levels is also indicated. For amplifiers A these latter levels are both -85 dB, they are reduced to -99 dB by the filters F. Following potentiometer P in fig. 7 the variation of the hum and noise levels are drawn for the *two extreme limits* of volume regulation. Depending upon the position of the potentiometers the hum and noise levels then have a value which lies between -99 and -124 dB. The noise level of amplifier B lies at -75 dB, this is 10 dB higher than that of A. This rise of 10 dB is due to the fact that the frequency characteristic of amplifier B from 2 000 c/s increases to $+14$ dB at 10 000 c/s, which is necessary for compensating a corresponding fall in the frequency characteristic of the ribbon microphone.

The hum level of amplifier B is -85 dB like that of A; because of this the hum and noise levels behind amplifier B differ by 10 dB as may be seen in fig. 7. In the minimum position of the volume regulators the noise contributions as well as the hum contributions of A and B are the same following amplifiers B, so that the total interference level is about doubled, which means an increase of these levels by 3 dB to AB_r and AB_b respectively (*cf.* fig. 7). In the maximum position of the volume regulators, on the other hand, the hum and noise contributions of amplifiers B are small with respect to those of A following amplifiers B, so that the hum and noise levels A_r and A_b in that case increase scarcely at all.

The hum and noise levels behind amplifiers C_1 and C_2 are calculated in the same way. It may be noted that with amplifiers C_1 and C_2 the hum is stronger than the noise, in contrast to amplifiers A and B. This must be ascribed to the fact that the heating elements of amplifiers C_1 and C_2 are supplied with alternating current and those of A and B with direct current. In the most unfavourable case with regard to the occurrence of disturbance, namely when in the case of a soft speaker the potentiometers are wide open, the hum levels of amplifiers C_1 and C_2 still, however, lie so low that the final hum level behind these amplifiers, which is chiefly due to amplifier A, is practically not increased.

Finally it may be seen that the average intensity of speech of the softest speaker at the end of the installation lies at $+24$ dB with respect to the maximum noise level which is chiefly due to amplifiers A. As soon as the speech into the microphone becomes louder, the ratio between noise and sound intensity becomes more favourable.

We shall now deal with the construction of the different kinds of amplifiers which were used in this installation.

These amplifiers may be divided into two groups:

1) The pre-amplifiers: A and B.
2) The final amplifiers, to which belong:
   the separation amplifiers C_1,
   the power amplifiers C_2 and
   the monitor amplifiers C_3.

The pre-amplifiers A and B

Of very great importance is the quality of the first amplifier *A*, whose noise level must be as low as possible, while all variations of intensity must be reproduced without distortion. In *fig. 8* it

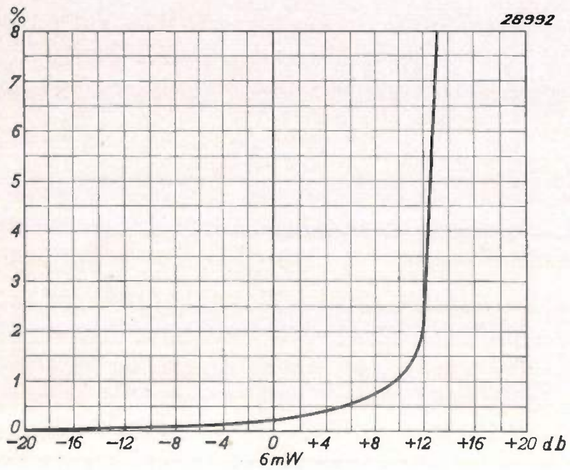

Fig. 8. Distortion characteristic of the amplifiers *A* and *B*.

may be seen that amplifier *A* satisfies these conditions. With an output of + 10 dB above 6 mW a distortion of only 1 per cent occurs, while it may be expected that the intensity at the peaks will not rise higher than to about 6 mW. The noise level of − 85 dB with respect to 6 mW at the output lies only slightly higher than the noise level which may be expected theoretically due to the presence of the resistance in the grid circuit of the first amplifier valve.

Amplifiers *A* and *B* consist of two amplifier stages with inverse feed-back. The inverse feed-back is introduced in order to keep the distortion very low over a wide range and moreover in amplifier *B* in order to be able to change the characteristic of the amplifier easily in such a way that the de-

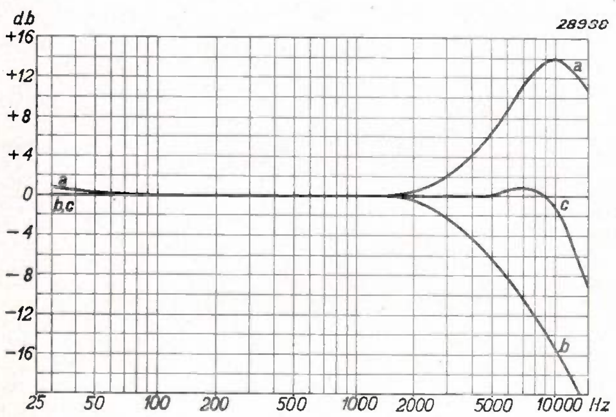

Fig. 9. Frequency characteristic of amplifier *B* combined with the frequency characteristic of the ribbon microphone.
a = frequency characteristic of amplifier *B*,
b = ,, ,, ,, ribbon microphone,
c = ,, ,, ,, both together.

crease in the sensitivity of the microphone for high tones is thereby compensated.

In *fig. 9* line *a* gives the frequency characteristic of amplifier *B*. Line *b* is the characteristic of the microphone. The combination of *a* and *b* gives line *c*, which thus represents the total frequency characteristic of the microphone with the amplifiers *A* and *B*, since the frequency characteristic of amplifier *A* is straight.

The amplifiers C_2, C_1 and C_3

The power amplifier C_2 consists of two resistance-coupled push-pull stages with inverse feed-back, and the frequency characteristic is straight from 30 to 10 000 c/s. The non-linear distortion is also very slight (see *fig. 10*). Up to the maximum output

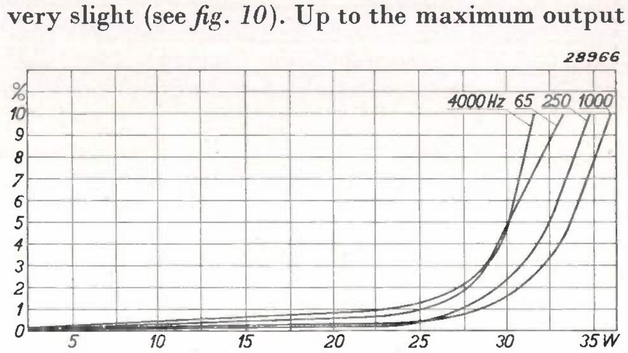

Fig. 10. Distortion characteristic of amplifier C_2 at different frequencies.

power of 25 W the distortion for all frequencies from 65 to 4 000 c/s is less than 1.3 per cent. These amplifiers like *A* and *B* are constructed in panel form (*fig. 11*).

Fig. 11. Amplifier C_2.

The separation amplifier C_1 consists of a simple one-stage amplifier with volume regulation, and can give on the line a level of 10 dB above 6 mW with less than 1 per cent distortion. Three of these are mounted on one panel. The monitor amplifier C_3 is of the same type as the power amplifier C_2 but of smaller energy.

The anode voltage apparatus of all amplifiers except the monitor amplifier C_3 are kept entirely separate from the amplifier part, in order to prevent the power transformers and smoothing chokes from causing hum in the amplifiers.

Signaling

It will be clear that considering the extensiveness of the installation a practical signaling system is indispensable in order to avoid incorrect switching manipulations at the various operating spots which are separated from each other. The places where working, the operator at the poste de régie must throw the key k. It may be seen in fig. 12 that then the white lamps in the control room and at the poste de régie are lighted. This is carried out per channel, so that it can immediately be ascertained in the control room which microphones are to be used. At the same time this is the signal for the personnel in the control room to begin with the preparations for the transmission.

The desired connections to the outgoing lines are made by means of the flexible conductors with jacks and sockets on the central control table. All

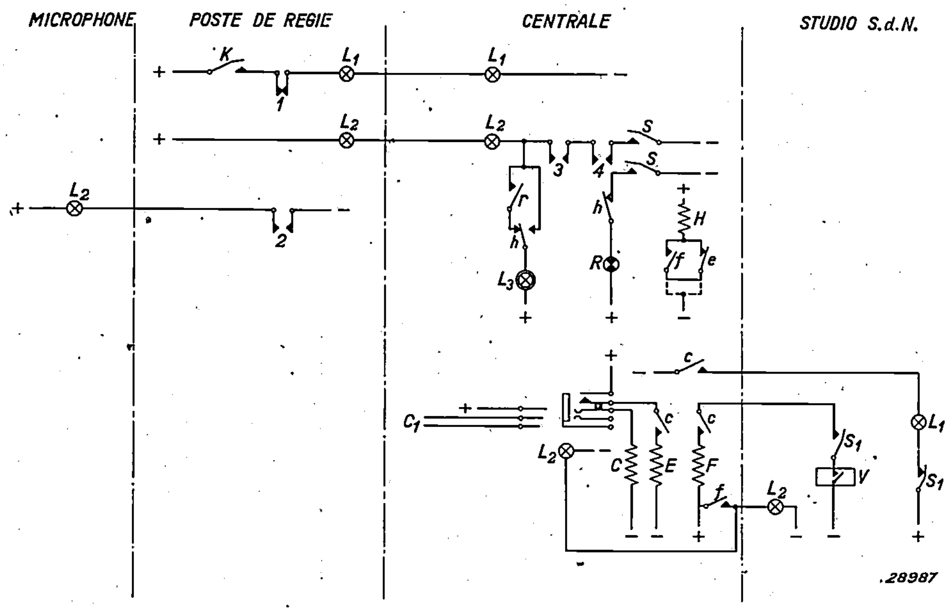

Fig. 12. Simplified diagram of the signaling system in which all relay contacts are shown in the currentless state.
L_1 white lamps, L_2 red lamps, L_3 red control lamp,
K key at the poste de régie, S key on the mixing panel in the control room,
S key on the connection panel in the studio,
1 and *2* are contacts on potentiometers of the poste de régie,
3 and *4* are contacts on potentiometers of the mixing panel in the control room,
V contact on connection plug in the studio, R relay interruptor with accompanying contact,
C, E, F and H are relays with accompanying contacts c, e, f and h.

switching processes are carried out by various persons, except in the central amplifier station, are: postes de régie, the reporters' cabins and the studio of the League of Nations transmitters.

The simplified diagram of the signaling system (*fig. 12*) makes clear the guiding principle according to which this system is arranged. Only two signals are used, one white (L_1) and one red (L_2). The white signal light is the preparatory signal, the red one indicates that all the manipulations which are necessary for carrying out the transmission are complete. Incorrect switching manipulations are furthermore betrayed by intermittent red light (L_3).

When for example a microphone must be set such conductors which are connected with the outputs of the amplifiers C_1 have a third strand (+), while the socket of the outgoing lines for which signaling is necessary are provided with auxiliary contacts over which the auxiliary signal currents are supplied. If for example we consider a line to the studio, we see in fig. 12 that when the jack of a cord is stuck in the socket a white lamp is lighted in the studios, because the relay C attracts its armature and the contacts c are closed. This lamp indicates the room from which the transmission is taking place. When the personnel in the studio, having noted the white signal, has connected the line with the transmitter (whereby a contact V_L

is closed) and has opened the volume regulator, the key S_1 is thrown and the relay F is thereby brought into action, so that the contacts f are closed and the red lamp L_2 near the socket in the control room begins to burn. It may be seen clearly that this signal cannot be given before all the manipulations just mentioned (closing of V and S_1) have been completed in the studio. When in the control room the state of the outgoing lines has been ascertained, the channel potentiometers are opened on the mixing panel. These potentiometers possess contacts (*3* and *4* in fig. 12) which are mechanically coupled with the axle, and which are closed as soon as the potentiometer is turned away from the zero position. If the key S is then thrown, the red lamps L_2 corresponding to channels in use at the poste de régie (and those in parallel on the mixing panel) begin to burn. The operator of the poste de régie now knows that they are ready in the control room for dealing with the transmission, and he may therefore set the microphones in action. This takes place, as was mentioned above, by the wide open setting of the microphone potentiometers. These potentiometers also have auxiliary contacts (*1* and *2* in fig. 12) coupled with the axle, which, however, only come into action when the potentiometer is turned to its utmost position. At that moment *1* is opened and *2* is closed. The white lamps L_1 are hereby extinguished, so that in the control room it may be ascertained immediately which microphone is switched on, while at the same time a red lamp L_2 is lighted near the microphone, so that the speaker can now finally see that his microphone is working.

It may also be seen from the figure how it is known in the control room that for instance a mistake has been made in the studio. In such a case the red lamp is not lighted at the corresponding socket, because the relay F has not worked and the contacts f are therefore not closed. If nevertheless the key S is reversed, an intermittent light from the red control lamp L_3 serves to remind those in the control room that there is a disturbance on one of the lines. In these cases the relay H falls out, since the relay E has functioned and the contact e for non-working is opened, while the contact f in parallel with it is also opened. A contact h for non-working switches in a relay interruptor R by which the contact r regularly interrupts the current circuit of the control lamp L_3. A similar connection is used for the lines to the reporter's mixing cabins.

A special signaling system is introduced in the "Salle des Assemblées" for the regulation of the sound intensity of the loud speakers for improving the acoustics. The sound intensity of each of the loudspeakers set up in the hall can be regulated separately with switches which are placed at the poste de régie (*cf.* fig. 3). Since the desired sound intensity is dependent on the number of persons seated in the hall, and since this sound intensity naturally cannot be judged for the whole hall at the poste de régie, control points are equipped at four places in the huge hall from which the operator at the poste de régie can be informed whether the sound intensity must be made greater or smaller. At the poste de régie, above every intensity regulator, a neon indicator lamp has been introduced. (In fig. 3 these neon indicators may be seen in a poste de régie of the "Salle des Assemblées" in the backward slanting metal holders which are placed above the upper row of knobs). At each control point in the room a switch button has been fastened to the wall, which has three positions. With this button resistances are put in series with the neon tube, in such a way that the length of the light streak corresponding with the three positions of the switch, can be made short, medium and long, to indicate that the sound in the room is too soft, satisfactory or too loud respectively.

THE PRODUCTION AND USE OF NEUTRONS

by F. A. HEYN.

539.185

The uncharged particles of matter with almost the same mass as a hydrogen atom, discovered by Chadwick in 1932 and called neutrons, are of great importance, not only for science, but also for various practical applications, for example in chemical, biological and even medical problems. For the production of neutrons so-called nuclear reactions may be used, in which certain elements are bombarded with α-particles from radium, for instance, or with deutons, which are ions of heavy hydrogen generated in a canal ray tube and accelerated by means of high voltages. The apparatus necessary for obtaining the fast deutons is described in detail in this article. With such a neutron generator for 600 kV, $6 \cdot 10^9$ neutrons per second are obtained; upon the use of α-particles from a natural source instead of deutons, 300 g. of radium would be necessary for the same yield of neutrons.

It is not so long ago (1932) that Chadwick discovered the elementary particles of matter which he called "neutrons", and which have been added as new components of matter to our picture of the universe along with the already known electrons and protons. In the few years which have elapsed since their discovery, numerous experiments have been carried out which have enlarged our knowledge of the properties of these particles. At the same time the technique of producing neutrons has made great progress, especially since it soon appeared that neutrons could be put to practical uses.

The discovery of the neutron was considered to be of such great importance that Chadwick was awarded the Nobel Prize. Actually he was not the first to observe the existence of neutrons: German physicists had already noticed that when beryllium is bombarded with α-particles (helium nuclei emitted by radio-active substances), it sends out a very penetrating radiation which can be shown by the ionization which it causes under certain circumstances. They believed themselves to be concerned with γ-rays (extremely hard X-rays).

This radiation was also investigated by others, but Chadwick was the first to be able to prove that there was no question of γ-rays here, but that the beryllium, when bombarded with α-particles, sends out uncharged particles of matter with a mass about equal to that of the proton, *i.e.* the nucleus of the hydrogen atom. These neutral particles were called neutrons. It soon appeared that they play a very important rôle in the structure of atomic nuclei [1]).

Principles of the production of neutrons

According to present conceptions atomic nuclei are composed of neutrons and protons. If we wish to obtain free neutrons, we must get them out of the atomic nuclei. This can be done by bombarding the nuclei with fast particles, for instance with the already mentioned α-particles. nuclear reactions then occur, in which, in some cases, neutrons are freed from the nucleus. For example in the experiment which led Chadwick to make his discovery, the nuclear reaction proceeded as follows:

$$\mathrm{Be} + \alpha \rightarrow \mathrm{C} + \mathrm{n}, \quad \cdots\cdots \quad (1)$$

which means in so many words that a beryllium nucleus, struck by an α-particle (α), is transformed into a carbon nucleus [2]) with the emission of one neutron (n). In practice this reaction is brought about simply by mixing beryllium powder with a substance which emits α-particles, such as radium or radium emanation. The mixture then emits neutrons in all directions.

The yield of reaction (1) is not high. If we bombard beryllium with the α-particles produced by 1 mg. of radium, *i.e.* $1.5 \cdot 10^8$ α-particles per second, we obtain about $2 \cdot 10^4$ neutrons per second. Thus only one in about seven thousand α-particles produces a neutron. With other substances bombarded with α-particles we find a still lower yield. If for example 100 mg. of radium is available it is possible in this way to produce $2 \cdot 10^6$ neutrons per second.

This quantity of neutrons is usually insufficient for the applications to be mentioned later. More radium (*i.e.* more α-particles per sec.) would be necessary. Aside from the very great expense, however, the world's supply of purified radium is limited, so that the investigator will not in general have a larger quantity than the above-mentioned (100 mg) at his command.

1) *Cf.* for instance: W. de Groot. Nuclear Physica, Philips techn. Rev. 2, 97, 1937.

2) The reactions are here given roughly. If the atomic weights are taken into account, it will be seen that the newly formed atoms are usually isotopes which rarely occur in nature.

Fortunately other nuclear reactions are now known in which neutrons are freed. In the first place there are the reactions in which "deutons", nuclei of heavy hydrogen (the isotope with atomic weight 2 of normal hydrogen, indicated by D as abbreviation for "deuterium"; d is the abbreviation for deuton), are used as projectiles. While the fast α-particles are supplied by nature in the form of radium rays, the fast deutons necessary for bombardement must be produced artificially in the following way.

Ions are produced in a discharge tube filled with heavy hydrogen; these ions, which are the desired deutons, are then accelerated by means of high voltages in some way or other. This technique is much more complicated than that used in reaction (1), but the results are also much better. While the yield of the reactions with deutons is indeed lower than of those with α-particles, the number of projectiles can be made much larger, so that many more neutrons are finally obtained. If for example lithium is bombarded with deutons with the following reaction as a result:

$$\mathrm{Li} + \mathrm{d} \rightarrow 2\ \mathrm{He} + \mathrm{n}, \quad \cdots\cdots (2)$$

the yield (with an accelerating voltage of 600 kV) amounts to $3 \cdot 10^{-6}$, *i.e.* only one in $3 \cdot 10^5$ deutons gives a neutron. If, however, our apparatus produces a number of deutons corresponding with a current of 1 mA, *i.e.* $6.3 \cdot 10^{15}$ deutons per second we obtain more than $2 \cdot 10^{10}$ neutrons per second, thus many more than with the α-particles from 100 mg of radium. The voltages and current hereby assumed may easily be realized in practice.

Other possible nuclear reactions are the following:

$$\mathrm{D} + \mathrm{d} \rightarrow \mathrm{He} + \mathrm{n}, \quad \cdots\cdots (3)$$

and

$$\mathrm{Be} + \mathrm{d} \rightarrow \mathrm{B} + \mathrm{n}, \quad \cdots\cdots (4)$$

in which heavy hydrogen itself and beryllium, respectively, are bombarded with deutons. It is not only possible to produce considerable quantities of neutrons per second with these reactions, but the neutrons here produced in general also possess a higher energy than those produced in reaction (1). This is very important in many applications.

The bombardement with fast deutons is therefore used in the Philips Laboratory for the production of neutrons and the previously described high voltage generators [3]) give excellent results in this process. We shall describe here several of the neutron generators developed.

[3]) See for example: A. Bouwers and A. Kuntke, Z. techn. Phys. **18**, 209, 1937. *cf.* also Philips techn. Rev. **1**, 6; 1936 and **2**, 161, 1937.

Neutron generator for 300 kV

The chief element of the neutron generator, namely the tube in which the deutons are produced and accelerated, is in principle similar to a canal ray tube. *Fig. 1* gives a diagram of such

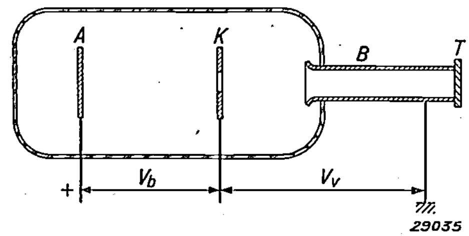

Fig. 1. Diagram of a canal ray tube. By a discharge between K and A ions are produced which pass through a canal in the cathode K and are accelerated by a voltage V_v between K and B. The ions may be allowed to strike a target at T.

a tube. The tube is filled with a gas (in our case heavy hydrogen) at low pressure. Between the cathode K and the anode A a discharge takes place, and ions are formed which move toward the cathode. Part of these ions pass through a hole in the cathode K in the right hand part of the tube, and are accelerated by the voltage V_v between K and B. The accelerated ions pass into the cylinder B. If neutrons are to be produced, the ions are allowed to fall on a target T of suitable material at the end of B. For the convenience of the investigator the tube is earthed at this end. The discharge in the space between K and A, which we shall call the ion source, can be maintained by thermionic emission from K, or it may be an ordinary glow discharge for which a sufficiently high voltage V_b between A and K is necessary and not too low a gas pressure [4]). This last method has been used in our case.

In order to obtain a satisfactory yield of neutrons with such a canal ray tube, a very high accelerating voltage V_v is necessary, as well as a high current of the ion beam. With these objects in view the tube shown in *fig. 2* was constructed [5]). A value of 300 kV was chosen for the accelerating voltage. A certain insulation length of the tube is necessary in order to avoid breakdown through the air along the outside wall of the tube. On the other hand, however, breakdown and the occurrence of a discharge in the accelerating tube must also be avoided, since that would mean a breaking down of the high

[4]) At least if no special measures are taken, such as the use of a magnetic field. F. M. Penning and J. H. A. Moubis: Physica **4**, 1190, 1937; *cf.* also Ned. T. Natuurk. **5**, 102, 1938.

[5]) The tube is a modified type of that first proposed by Oliphant and Rutherfond.

voltage. This may be done by making the gas pressure very low in the space between K and B, and the distance very short between K and B. A low gas pressure corresponds to a long free path λ of the electrons which are freed by collisions of ions with the electrode B. These electrons, when their path from B to K is much shorter than λ, will cause no appreciable ionization, *i.e.* no breakdown will occur. The two requirements: great insulation length in air, small separation in the vacuum, led to the construction which is shown in fig. 2. The source of ions S is placed in the accelerating tube G.

Source of ions

It is important for the discharge to take place at the source of ions in such a way that as many ions as possible pass through the canal k, *i.e.* the discharge must be concentrated in the neighbourhood of k. With the shape indicated of the discharge space S such a focussing action is automatically obtained. The discharge follows approximately the directions of the dotted lines between the partition L soldered on to the anode A and the neighbourhood of the canal in electrode K. In order to prevent the occurrence of a discharge at other spots between A and K, as was explained for the accelerating space W, the distance between the electrode cylinders is made so small (see fig. 2) that, with the gas pressure used, it is still small compared with the free path of the electrons. The insulation between the electrodes A and K is formed by the glass cylinder J, which is connected by airtight chrome-iron glass joints with the flanges C and E.

At a voltage of 50 kV and a current of 40 mA a power of 2 kW is transformed at the source of ions. The heat hereby developed makes good cooling necessary. The electrode K dissipates its heat by radiation; the partition L in the electrode A is cooled with running water which enters and leaves at O. Since A is at high voltage, a closed cooling water circuit with an insulated pump is necessary. If it is desired to avoid the insulated pump installation, oil may be used for cooling, and this is conducted to the apparatus through glass tubes (visible in the photograph fig. 7 at the extreme left).

The electrode K must be able to withstand a vigorous bombardement with ions, and is therefore made of aluminium which erodes relatively little and also possesses a good heat conductivity. The electrode A and the wall L are made of chromium-plated copper, since a copper surface easily leads to an arc discharge.

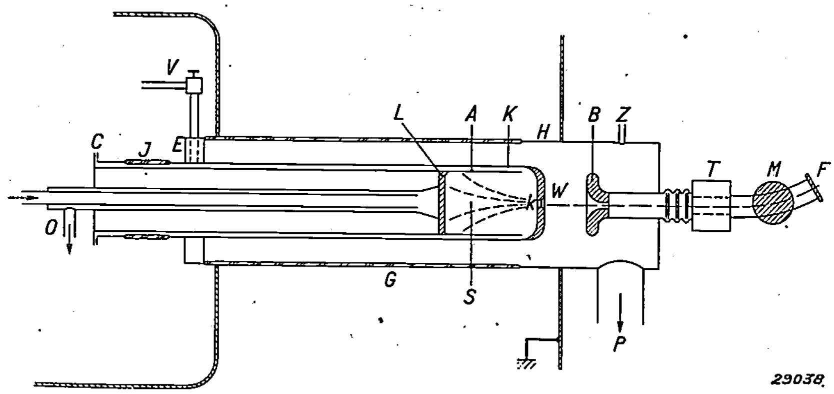

Fig. 2. Cross section of the tube constructed. S discharge space of the source of ions. A and K electrodes. L cooled partition: at O the cooling water or oil enters and leaves. J glass cylinder for insulation fused to the flanges C and E of the electrodes A and K. V needle valve. G accelerating tube. The ions pass through the canal k and are accelerated in the space W between the electrodes K and B. H earthed metal cylinder. The pump is connected at P, a manometer at Z. T stopcock. M magnetic field. A Faraday cage is connected at F, or, after removal of the magnetic field, a target is placed at the end of the tube.

The canal k has a length of 5 mm and a diameter of 3 mm. The greater the diameter the greater the percentage of ions which can enter the accelerating tube. In determining the dimensions of the canal, however, the gas pressure must be taken into account. As already stated, the gas pressure must be very low in the accelerating tube (of the order of $5 \cdot 10^{-4}$ mm of mercury). The pressure at the ion source may however not be too low, since otherwise the voltage necessary to obtain a discharge would be too high. It is therefore necessary to maintain a certain difference in pressure between ion source and accelerating tube, which is done by rapid pumping *via* P. The pumping speed of the pump used (an oil-diffusion pump with a speed of 20 litres per second) therefore determines the permissible dimensions of the canal.

In order to replace the gas which is continually pumped away, fresh hydrogen is continually added through an opening in the flange E. The pressure in S can be regulated while the tube is under tension by means of a needle valve V, which is operated *via* an axle of insulating material. This regulation is very important, since the number of ions which enter the accelerating tube from the source depends strictly upon the pressure.

Acceleration of the ions

The acceleration of the ions takes place in the acceleration space W between K and B (length 5 cm) in tube G, which is 80 cm long and 20 cm wide, and may be made of glass or of artificial resin ("Philite"). With the latter material the chance of breakage or breakdown is much less, while it is very satisfactory as to vacuum and high voltage. The electrode B is also of chromium-plated copper, and has such a shape that the field strength at its surface is as low as possible, in order to avoid cold emission of electrons. The wide edge of B provides for a variation of the field between K and B such that the electrons, which are formed by the bombardment of ions in the neighbourhood of the opening in B, are unable to leave the acceleration space W. This point is important for the satisfactory functioning of the tube, since otherwise the freed electrons wandering about outside W can cover great distances and may also cause secondary emission. The high ionization caused by all this might finally cause breakdown. For the same reason, *i.e.* in order to prevent the stray electrons from doing any harm, the metal cylinder H extends slightly beyond the ends of the actual acceleration chamber W. The part of K around the exit of the canal is made concave in order to obtain a concentration of the ion beam.

After the acceleration the ion beam passes a stopcock T, a magnetic field M, and falls upon F. If the current of the ion beam is to be measured a Faraday cage is connected at F (it must be cooled with water). For the production of neutrons the magnetic field is first removed, and the already mentioned target is introduced, for example a plate of lithium which, upon bombardment with ions of heavy hydrogen, sends out neutrons in all directions (reaction (2)). The neutrons pass through the walls of the tube to the outside since they are practically unabsorbed. By means of the stopcock T, the main part of the tube is shut off when the target is to be changed.

The magnetic field M serves to determine the yield of active projectiles of the tube, *i.e.* to separate the ions of heavy hydrogen from the other ions present (from impurities in the gas), and also to separate the ions of the heavy hydrogen molecule (D_2^+) and the heavy hydrogen atom (D_1^+; in the reaction equations we have used d following ordinary usage).

These two kinds of ions are deflected by different amounts in the magnetic field because of their different masses. *Fig. 3* gives the part of the magnetic

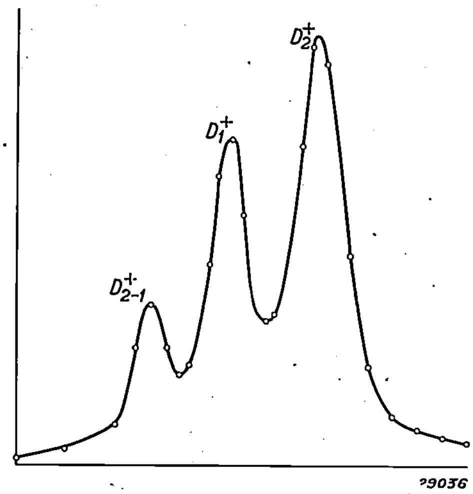

Fig. 3. Part of the magnetic "spectrum" of the ions in the ion beam. The deflection in the magnetic field is plotted horizontally, the ion intensity vertically. The ions D_1^+ (deutons d) of the middle maximum are chiefly effective in the production of neutrons, since they possess the greatest velocity.

"spectrum" of the ion beam which is important for our purposes. The second and third maxima correspond to the atomic ions and the molecular ions respectively, while the first (D_{2-1}^+) may be ascribed to particles which are dissociated in the space between B and M, and which are therefore accelerated as molecular ions, but are then deflected as atomic ions. The second maximum is lower than the third, *i.e.* the number of atomic ions is smaller than the number of molecular ions, which was to be expected since at the ion source the formation of molecular ions (D_2^+) is the primary process in the production of deutons (D_1^+). The ions of the second maximum are the most effective in the production of neutrons, since they have the greatest velocity.

The total current of the ion beam reached a maximum of 1.5 mA with this tube. The source voltage was 50 kV and the source current about 40 mA, while the acceleration voltage lay between 250 and 300 kV.

The accelerating voltage is supplied by a high tension generator of a type which has been described previously in this periodical [3]). The excitation of the source voltage of 50 kV is a problem by itself, since this voltage must be supplied to the electrodes A and K which are already at a potential of 300 kV with respect to earth. Use is usually made of a highly insulated transformer or of an insulated generator which is driven by a motor with an insulating belt or axle. In the Philips neutron generators, however, a circuit is used which corresponds in principle with the well-known cascade circuit of two stages. The right half of fig. 4 shows the scheme. The condenser of the first stage of the cascade circuit is replaced by two condensers C_1 and C_2 in series. This does not alter the principle, and C_3 will receive double the transformer voltage as in the ordinary cascade circuit; C_1 and C_2, however, are insulated for 300 kV, so that the appropriate electrodes of the tube can be connected at K and A. This solution has various advantages: it is very economical, the apparatus occupies little space, it causes no noise or vibrations and makes possible an easy regulation of the source voltage by regulation of the primary voltage of the transformer. In order to keep the ripple on the source voltage small, even at currents up to 40 mA, without C_1 and C_2 being made very large, a frequency of 500 c/s is used. For this a converter from 50 to 500 c/s is necessary, but this solution still remains more economical than the use of large condensers for C_1 and C_2.

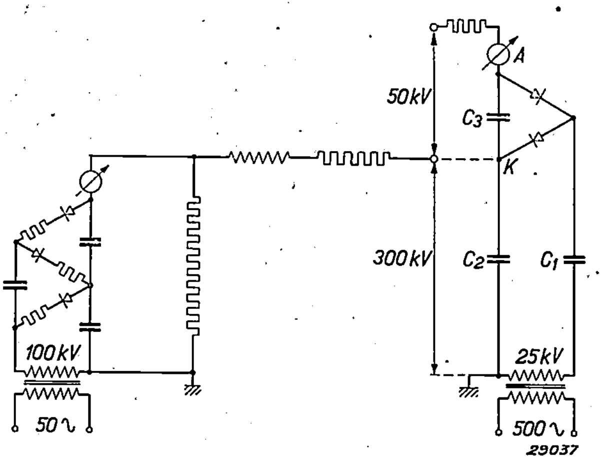

Fig. 4. Sources of high voltage for the neutron generator. In the left-hand part the accelerating voltage of 300 kV is generated with the well known cascade connection; in the right-hand part the source voltage of 50 kV which is supplied to the electrodes A and K, the latter of which is raised to a potential of 300 kV with respect to earth.

Fig. 5 is a photograph of a complete neutron generator constructed for the Laboratoire des Rayons X in Paris. The partition from which the sliding door is removed forms the separation between two rooms, at the left the workroom, at the right the high tension room. The tube passes through the partition, so that the neutrons are produced on the earthed target in the workroom. The pump installation is also placed in that room, and is mounted on the left-hand side of the partition, while the high voltage apparatus is also operated from there.

When the neutrons are produced according to reaction (2), the yield from this apparatus is $2 \cdot 10^8$ neutrons per second. In order to obtain the same quantity of neutrons with α-particles from radium, according to reaction (1) 10 g of radium would be necessary with a value of about £ 55,000.

Neutron generator for 600 kV

While the intensity of neutrons obtained with the apparatus for 300 kV is sufficient for many physical experiments, it is not adequate for biological and medical applications. An installation for 600 kV was therefore designed. For this purpose a second tube is connected with the earthed end of the tube in fig. 2, in which the ions are further accelerated by a negative voltage of 300 kV with respect to earth. In this way the target situated at the extremity of the second tube is under high tension. This is, however, no objection for most applications, while a construction with an earthed target, in which therefore a voltage of 600 kV with respect to earth had to be generated, would be many times as expensive. The arrangement is drawn schematically in *fig. 6*. The left-hand part is in the main similar to the 300 kV tube described. The two electrodes Q and R of the second accelerating tube, between which the ions are accelerated the second time, are again of chromium-plated copper. Special provisions are necessary for the concentration of the ion beam on the target. The shape and separation of Q and R have little effect on this, since in this region the heavy fast particles can only be slightly deflected by any irregularities in the variation of the field. Very important, however, is the separation and shape of the electrodes K and B of the first accelerating tube where the particles still have only low velocities. Since the beam never leaves the first accelerating tube exactly in the direction of the axis, the second tube is connected with the first by a flexible joint (a piece of tombak tubing) and suspended (at D) in such a way that it can be set in the direction of the beam, and all the ions can impinge on the target.

The current in the second tube is 0.5 mA at the most, so that at 600 kV an energy of 300 W is given off at the target which makes good cooling necessary. This is done by means of oil.

Due to erosion under the bombardement with ions a cavity is formed in the target after some time; impurities (decomposition products of vapours of grease) are also deposited upon it. By means of the second flexible connection N stopcock and target can be shifted so that the ion beam, which has a cross section area of several sq.mm, strikes the target at a different spot.

The yield of the apparatus for 600 kV is $6 \cdot 10^9$ neutrons per second, according to reaction (2). In order to obtain the same quantity of neutrons by reaction (1) with α-particles from radium, 300 g of radium would be needed!

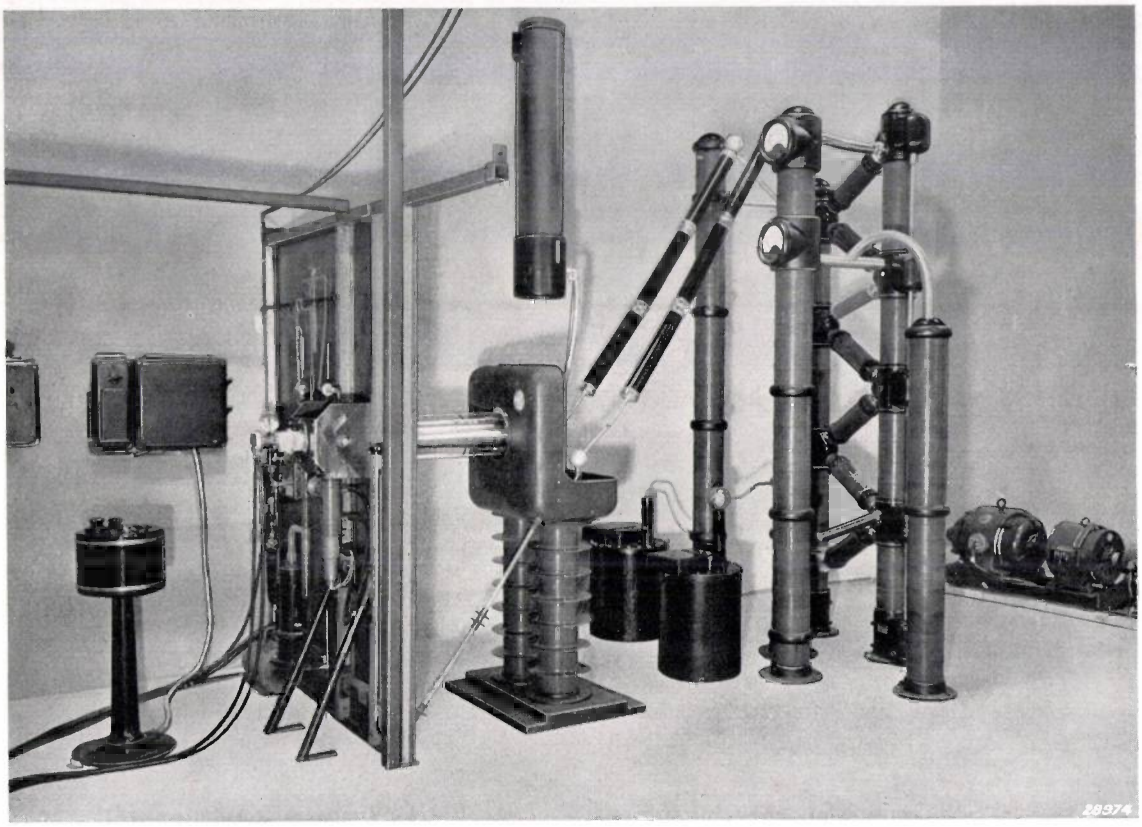

Fig. 5. Photograph of a complete neutron generator for 300 kV, constructed for the Laboratoire des Rayons X in Paris. In the high voltage chamber to the right of the partition (from which the sliding door has been removed) is the canal ray tube with a high voltage apparatus. At the extreme right stands the converter from 50 to 500 c/s. for the source voltage. Above the tube may be seen the insulated cooling pump. The extremity of the tube projects through the partition into the workroom on the left, where the neutrons are produced on an earthed target, and where all the control apparatus necessary for working are set up.

Fig. 7 is a photograph of the complete apparatus for 600 kV.

Several properties of neutrons

The use to which neutrons may be put is of course determined entirely by their unusual properties, several of which we shall mention here.

Since neutrons are neutral they cannot react electrically with the electrons and atomic nuclei of matter. Only by direct collision with atomic nuclei can neutrons lose energy and change their direction. It is because of this fact that their power of penetrating into matter is very great: those forces which have the greatest stopping power for electrons and ions are lacking here.

There are, however, several special substances which absorb neutrons strongly, chief among these are all hydrogen compounds, especially water and paraffins with a relatively high hydrogen content. If a neutron collides with a hydrogen nucleus, the collision is one between two almost equally heavy particles, and one particle, the neutron, can give off a large part of its energy to the other, the hydrogen nucleus. The neutron is thus strongly retarded. In addition, the hydrogen nucleus, like many other atomic nuclei, has the property of combining with slow neutrons. The result is a strong absorption. Of the other neutron absorbing materials we mention particularly cadmium, samarium, europium and gadolinium. These substances absorb neutrons, especially when the neutrons have first been retarded, by means of water for example. The uncharged neutrons penetrate particularly easily into the nuclei of these elements and are held permanently by the nuclei.

Another special property of neutrons, and for practical purposes perhaps the most important one, is that practically all the elements of the periodic system may be made radioactive by bombarding them with neutrons. The activation usually takes place by a "process of capture" as described above. The atomic nucleus which has captured a neutron is now, however, unstable and breaks up after some time with the emission of β-particles, as do many natural radioactive elements.

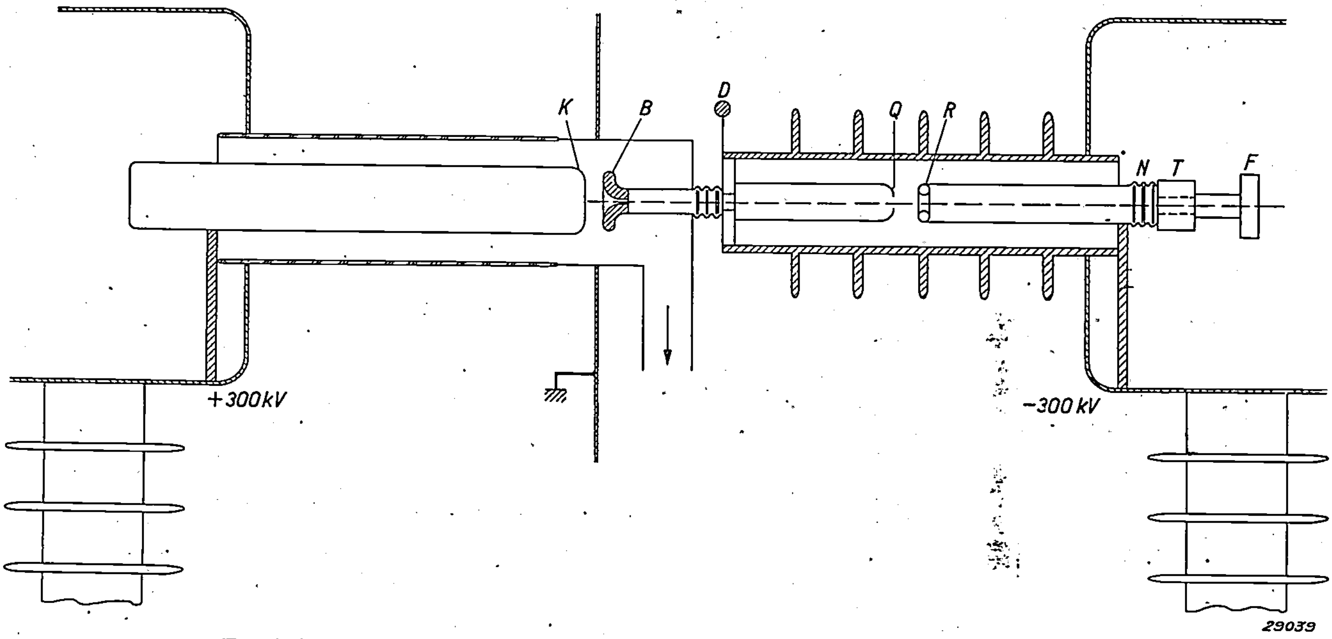

Fig. 6. Canal ray tube for 600 kV. Left-hand part as in fig. 2. To the right a second accelerating tube, where the already accelerated ions, which have passed through the earthed middle section B-Q, are again accelerated between Q and R with a voltage of 300 kV. The target is at a high potential (—300 kV) with respect to earth. The second tube is connected with the first by a flexible joint, and suspended at D in such a way that the target may be adjusted with respect to the ion beam. By means of a second flexible joint N different points of the target may be exposed to the ion bombardment.

Applications

In the applications of which we shall only mention those which have already been tested practically, use is made in the first place of the artificial radioactivity induced by neutrons. The amount of radioactive material which can be made with neutrons is considerable, and is for example enough for medical purposes. Compared to the natural radioactive substances the artificial ones have several advantages from the medical point of view. For example they can be introduced into the body in the form of all kinds of compounds, which is not generally possible with natural ones. The natural radioactive substances form series of equally radioactive decomposition products, some of which have long life times, so that once introduced into the body, they cause a permanent irradiation, which is not at all desired. The artificial radioactive substances, however, do not form such series, their decomposition products are no longer radioactive. If for example a patient is allowed to drink a solution of common salt which has been made radioactive, the radiation in his body will have disappeared entirely after some time.

Furthermore elements which have been rendered radioactive may be used as indicators in chemical and biological experiments. They behave chemically just like ordinary non-radioactive elements, while their presence can be detected by their radioactivity, a method which is much more sensitive and universal than chemical methods. In the realm of biology important results have already been obtained with this method. For example the distribution of phosphorus compounds throughout the body has been investigated, by feeding animals radioactive phosphorus compounds, and then finding out what parts of the body had become radioactive.

Neutrons themselves are important especially in medical applications. In experiments carried out in the Philips Laboratory by Dr. den Hoed of the cancer institute in Amsterdam it was found that neutrons have an action on organic tissues which is analogous to that of X-rays [6]). There is,

[6]) The results of this investigation were communicated at the 5th International Congress of Radiology, Chicago, 13 - 17 Sept. 1937.

however, a certain difference. While X-rays exert their influence by the ionization which they cause in the tissue by their reaction with the electrons present in it, neutrons cause an ionization due to the fact that they transfer energy to hydrogen atom nuclei present in the tissue by collision, and the latter in turn cause ionization in the tissue. In the last case the ionization is much more local, while the ionization by X-rays is more homogeneous. It may also be expected that it will be easier to irradiate more deeply lying parts with neutrons.

In experiments carried out in the Philips Laboratory in collaboration with the Kaiser Wilhelm Institut in Berlin [7]), in which insects were irradiated with neutrons, it was found possible to provoke mutations, as may be done with X-rays. Here especially several differences were observed as a result of the different manner of ionization.

As may be expected after the above, it was also found from the animal tests carried out that workers must be well protected against neutrons just as against X-rays and γ-rays. On the basis of the properties mentioned above it follows that this cannot be accomplished with lead as in the case of X-rays. The best method is that of surrounding the source of neutrons with water of paraffin, about which a layer of cadmium can be added. In fig. 7 the large number of blocks of paraffin may be seen which surround the source of neutrons.

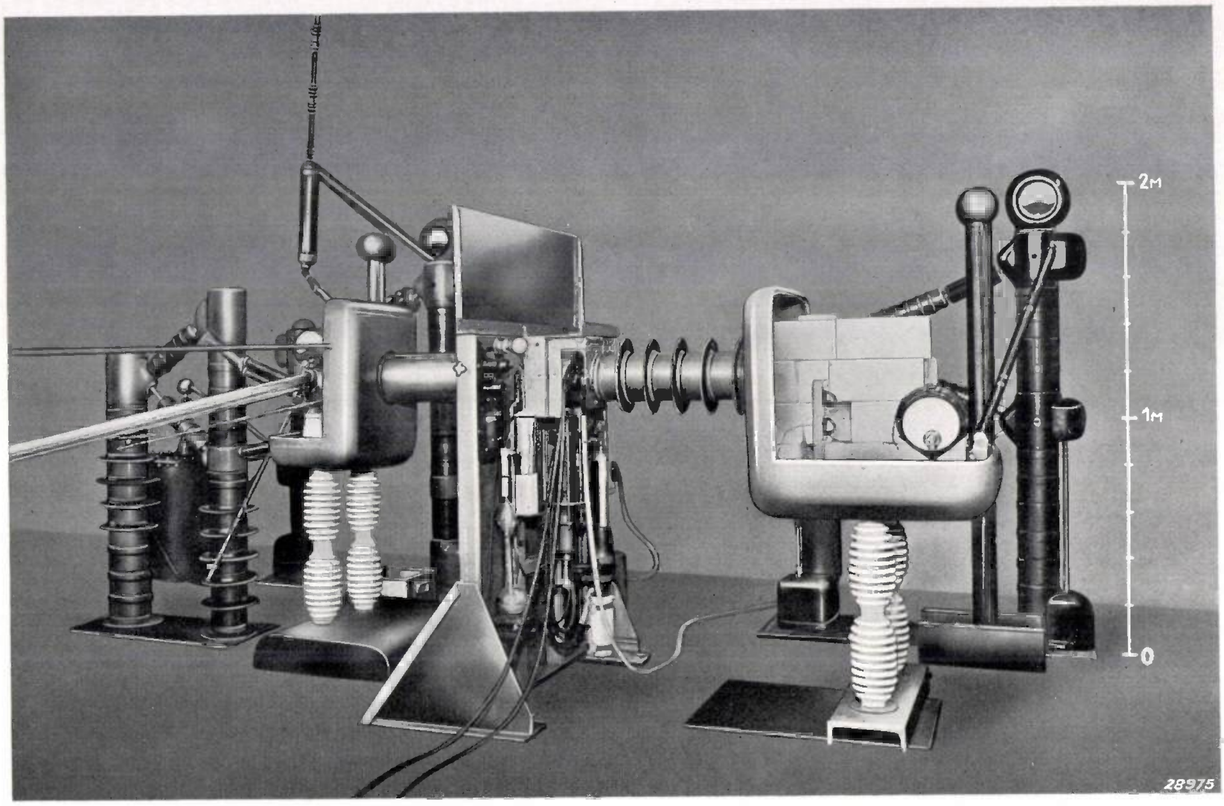

Fig. 7. Photograph of the complete neutron generator for 600 kV. The partition in the middle on which the pumping installation is mounted is earthed. To the left of it the first accelerating tube (this one was made of glass), to the right the second (of "Philite"). The neutron source proper at the right-hand end is surrounded by a number of blocks of paraffin for the protection of the worker against the neutrons. On the extreme left may be seen the source voltage generator for 50 kV, while behind the apparatus the two high-voltage generators each for 300 kV, are visible.

[7]) N. W. Timoféeff-Ressovsky, K.G. Zimmer and F.A. Heyn, Naturwiss. **26**, 108, 1938.

APPLICATIONS OF CATHODE RAY TUBES IV.

by H. VAN SUCHTELEN.

621.317.755 : 621.385.832

The recording of diagrams

While the cathode ray tube is usually considered to be an oscillograph, it is actually in the first place an instrument which can record simultaneously and in mutually perpendicular directions two voltages (or currents). It is typically an instrument which records diagrams directly, and the diagrams only become oscillograms when one of the voltages to be recorded is connected with the time by a simple relation, and therefore provides a time base.

If is therefore quite natural that a great many applications of the cathode ray tube consist in the recording of diagrams or characteristics, and in most cases the ordinary cathode ray oscillograph is particularly adapted for these applications.

The recording of a diagram on the cathode ray tube offers several advantages over the point-by-point construction of the diagram by static measurements. The taking of meter indications point by point is undoubtedly the most accurate method in most cases. This static method can, however, not always be used.

In the first place it is impossible when overloading and consequent damage to the object to be measured might occur during static measurement. The cathode ray oscillograph makes possible much shorter but continually repeated measurements. Such rapid measurements would not be possible with a slow pointer instrument.

Very rapid measurements may also be necessary when it is a question of momentary values in a typical periodic phonomenon. The relation between current and voltage in a self-induction, for example, is dependent upon a definite speed, and can therefore never be measured statically.

A further very obvious advantage is the fact that the complete diagram is obtained so quickly. The diagram can usually be traced at a relatively arbitrary speed by the light spot, and if it is repeated regularly an impression of a stationary figure is obtained as in the case of oscillograms. This last advantage is particularly important for measurements in series production, where accuracy may to a certain extent be sacrificed to speed.

In the examples to be discussed the advantages mentioned will play a part individually and collectively. In the foregoing article of this series one case of the recording of a diagram was mentioned incidentally, namely the recording of the modulation characteristic of a transmitter. This is a typical example of a control measurement during the process of manufacture which is striking because of its simplicity and its rapidity.

Measurement of valve characteristics

In the development and manufacture of radio valves measurements of longer or shorter series of valves of the same type are often very important. It may in one case be a question of purely quantitative measurements, as in the determination of deviations in the products, or in other cases it may be more a question of qualitative determinations, or the detection of certain deviations which might occur in a few cases. In the last case the cathode ray tube is certainly the most suitable instrument.

An important characteristic of a valve is the one which gives the relation between anode current and anode voltage with the grid voltage as a parameter. In some cases a single value of this parameter is enough, in other cases a series of characteristics is required. The circuit with which such characteristics are recorded for receiver valves in the Philips Laboratory is shown in principle in *fig. 1.*

The anode circuit of the valve to be examined X is fed with a voltage which varies periodically between zero and its maximum value. This voltage

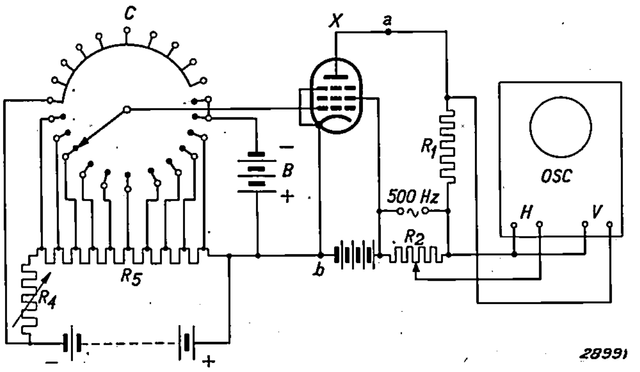

Fig. 1. Circuit for the recording of diagrams. The oscillogram shows the anode current of the valve X as a function of the anode voltage. By means of the rotating commutator C a number of different control grid voltages are allowed to act during each revolution ($^1/_{25}$ sec), so that a series of curves appears on the screen. One of the grid biases (-6 volts) is not taken from the resistance R_4, but it is supplied by the battery B. This is connected with two of the contacts. Because of this the line for -6 volts appears stronger than the other curves in fig. 2.

is obtained by the connection in series of a direct voltage and an alternating voltage of 500 cycles and of the same amplitude. In the cathode ray tube a horizontal deviation must be obtained which is proportional to the variation in anode voltage. For this purpose a potentiometer R_2 is connected over the alternating voltage, which reduces the value used to a figure suitable for the tube. The voltage tapped off is led to the terminals of the plates for horizontal deflection of an ordinary oscillograph (terminals H in fig. 1). The Philips cathode ray oscillograph G.M. 3152 is also provided with such a connection. The anode current flows through the measuring resistance R_1 and the fairly low voltage which occurs over this resistance is fed to the built-in oscillograph amplifier V. In this way a vertical deviation is obtained which is proportional to the current.

In this simple manner a single diagram can already be recorded with a given screen grid and control grid voltage.

The way in which a series of characteristics is recorded simultaneously is also indicated in fig. 1. By means of a rapidly rotating commutator the control grid is momentarily connected to the successive taps of a potentiometer R_5. The difference in potential between successive contacts can be set at 1, 2, 3 volts, etc. by means of a series resistance. One would be inclined to suppose that the commutator should be exactly synchronized with the frequency of the anode alternating voltage. This is, however, unnecessary. Without the slighest difficulty it is possible to pass from one characteristic to the following one, even before the first one has been completed. The part lacking is then traced during the next rotation of the commutator Since the frequency at which the commutator rotates, and therefore at which the whole series of characteristics is repeated, is of the order of 25 c/s an impression is given of a single picture with no interruptions.

The commutator is composed of 22 contact-lamina. 10 of these are connected to the successive taps of R_5. Series of 10 characteristics are therefore recorded. For the first characteristic $V_g = 0$; all the other curves are recorded with negative grid bias. Two lamina are connected to a calibration voltage of, for instance, —6 volts. Because of this two characteristics are always traced one over the other for $V_g = -6$ volts, independent of the setting of R_4. This line is noticeable because of its greater brilliance. It is now possible to adjust R_4 in such a way that the second, third, fourth, etc. characteristic of the series coincides with that for $V_g = -6$ volts. Then it is known that V_g changes by the amounts 6, 3 or 2 volts, respectively.

Finally the other 10 laminae of the commutator are connected to a fairly high negative voltage, so that the anode current is completely suppressed during half of every revolution. This is necessary for the following reason. The anode voltage reaches a very low value for some time during the measurement (it falls as far as zero.) In pentodes an abnormally high screen grid current occurs at these low anode voltages. In order to protect the screen grid from excessive heating the cathode current is suppressed for some time during each measurement. This is an example of the first advantage mentioned above to be derived from the use of the cathode ray tube.

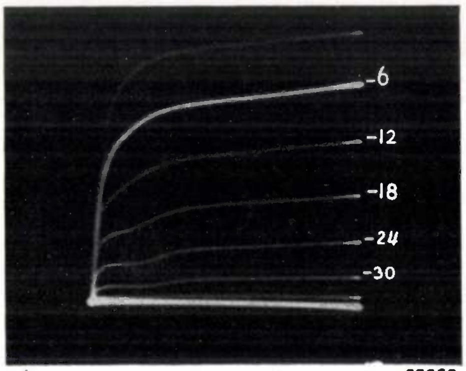

Fig. 2. Diagram of the anode current as a function of the anode voltage recorded with the help of the circuit shown in fig. 1. The abscissae run from 0 to 500 volts, the ordinates from 0 to 200 mA.

Fig. 2 shows a series of characteristics recorded on a test model of a pentode end valve. The second curve from the top corresponding to $V_g = -6$ volts is noticeable because of its intensity. The grid voltage was apparently varied by steps of 6 volts. Furthermore in the bends of the third, fourth and fifth characteristics an irregularity may be seen which indicates secondary emission of the anode. This irregularity is the weak counterpart of the much greater dip which occurs with tetrodes (see Philips techn. Rev. 3, 215, July 1938).

When the resistance R_1 in fig. 1 is added to the screen grid circuit of the valve to be examined, the characteristics are obtained of the screen grid current as a function of the anode voltage (*fig. 3*). In this figure also the guiding line for —6 volts may again be recognized. Furthermore irregularities in the curve may also be seen which are due to the secondary emission of the anode; for every fall in the anode current due to secondary emission there is a corresponding rise in the screen grid current.

Finally it may here be seen that relatively high currents flow when the anode voltage is low. These currents would overload the screen grid in static measurement.

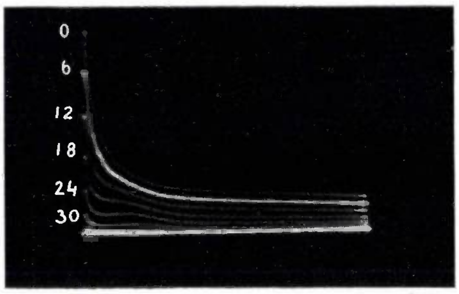

Fig. 3. Diagram of the screen grid current as a function of anode voltage, recorded with the same valve as in fig. 2, and on the same scale.

The measurement of a hysteresis loop

Another example of a measurement in which a dynamic method gives a certain simplification compared with static measurement is the determination of the induction B in a magnetic material, *i.e.* the measurement of a hysteresis loop. A ballistic galvanometer is usually used for this, and the induction impulse occurring in a secondary winding is measured when the magnetizing field is interrupted by switching off the current in the primary winding (see J. J. Went, Philips techn. Rev. 2, 84, 1937). A secondary voltage occurs with momentary values proportional to $\mathrm{d}B/\mathrm{d}t$. When the magnetization is not changed suddenly, but periodically, the induced voltages can be integrated very simply electrically, instead of mechanically as in the ballistic galvanometer, so that voltages proportional to B occur.

The principle of the circuit with which this is done is shown in *fig. 4*. A closed ring is made of the test material in the usual way and two windings are laid around it. The primary winding S_1 is supplied with alternating voltage over a resistance R_1. A voltage therefore occurs over this resistance which is proportional to the number of ampere turns, and this voltage can be led directly to the terminals H for horizontal deflection.

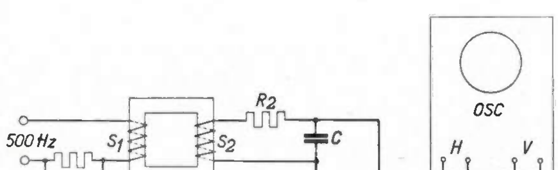

Fig. 4. Circuit for recording hysteresis loops. The voltage at the coil S_2 is proportional to $\mathrm{d}B/\mathrm{d}t$. If the impedance of C is small with respect to R_2 the voltage at C is proportional to the momentary value of B.

As mentioned above the following voltage occurs over the secondary winding S_2:

$$v_2 = \text{const.}\ \frac{\mathrm{d}B}{\mathrm{d}t}.$$

The combination of resistance R_2 and capacity C is connected with this winding. When care is taken that the voltage at C is very small compared with v_2, it may be assumed that a current flows through R_2 and C which has a momentary value of:

$$i_2 = \frac{v_2}{R_2} = \frac{\text{const}}{R_2}\frac{\mathrm{d}B}{\mathrm{d}t}.$$

The voltage at C thereby becomes:

$$v_C = \frac{1}{C}\int i_2\ \mathrm{d}t = \frac{\text{const}}{R_2\ C}\ B.$$

The voltage at C is therefore actually a measure of B. This voltage must be small with respect to that at S_2, while the latter will not for practical reasons be much higher than for instance 100 volts. From this it follows that the oscillograph amplifier must be connected between C and the plates for vertical deflection.

Fig. 5. Hysteresis loop recorded by means of a circuit like that of fig. 4

Fairly heavy demands as to phase fidelity are made of the oscillograph amplifier in this and in the previous case. In general a constant amplification over a very wide frequency range is the primary requirement in the case of an oscillograph. Over the greater part of this range the phase fidelity is also practically perfect. At the boundaries, however, inevitable small phase shifts occur. Since very slight shifts are very disturbing in the reproduction of diagrams (double lines), in the two cases described the measurements were carried out at a frequency of 500 c/s instead of at the usual 50 c/s since upon amplification of these low frequencies phase shifts would already occur.

A hysteresis loop recorded according to fig. 4 is reproduced in *fig. 5*.

Measurement of frequency with the cathode ray tube

A type of measurement which is connected with the two preceding ones to only a certain extent is the comparison of unknown frequencies with a standard frequency, *i.e.* for example a method for the measuring generators which are used in acoustic measurements.

The comparison of the frequency of a generator to be calibrated with a standard frequency is often done by ear, when the frequency to be determined differs from the standard by a whole number of octaves. This method may be very accurate, because of the fact that slight deviations can be detected due to the occurrence of beats. When, however, one of the two oscillations contains overtones, it becomes difficult to determine the number of octaves difference. Similarly the comparison by ear is not possible with great accuracy when the frequency to be calibrated does not differ from the standard frequency by a whole number of octaves.

By supplying two alternating voltages of different frequency to the two sets of plates of a cathode ray tube a very accurate comparison is possible. This is also true in the case of quite arbitrary relations between the frequencies, as long as they may be expressed in the form of a ratio of two whole numbers which are not too large.

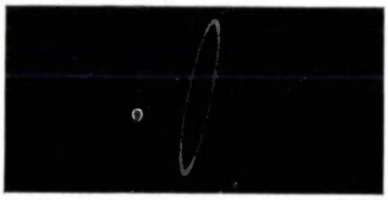

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Fig. 6. If sinusoidal alternating voltages of the same frequency but of different phase are supplied to the horizontal and vertical deflection plates of a cathode ray tube, the resulting figure is an ellipse.

When voltages whose frequencies are related as whole numbers, are supplied to the plates, a stationary figure results. If the frequencies are the same, a straight line or an ellipse is obtained according to the phase relation of the voltages (*fig. 6*). With other ratios of frequencies the so-called Lissajous figures are obtained. *Fig. 7* gives an example of one of these.

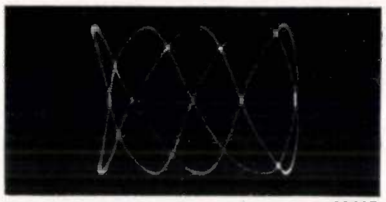

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Fig. 7. Lissajous figure which occurs when the frequencies of the vertical and horizontal deflecting voltages are in the ratio 5/2.

The ratio of the frequencies may be deduced from the last figure in the following way. In the time during which the figure is completely traced five upper peaks have been reached in the vertical direction, therefore there have been five periods. In the horizontal direction the left (or right) side of the figure has been reached twice, therefore two periods have elapsed. The frequency of the vertical oscillation was therefore the higher, and was apparently $^5/_2$ of that of the horizontal oscillation. As soon as the ratio of the frequencies deviates slightly from $^5/_2$, the successive figures no longer exactly coincide, and the figure appears to move. Because of this it is possible to measure frequencies very accurately.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS GLOEILAMPENFABRIEKEN

1309: K. S. Knol, M. J. O. Strutt and A. van der Ziel: On the motion of electrons in an alternating electric field (Physica **5**, 325 - 334, May 1938).

With a high vacuum electron tube which consists of a cathode, two grids and an anode, by applying the correct alternating voltages of high-frequency and direct voltages, it is possible to cause a part of the electrons to attain a higher speed than that which corresponds to the momentary value of the electric potential at the point where the electron is situated. From the equations of motion an expression is derived for the kinetic energy of the electrons. This expression is discussed and applied to special cases. Upon comparison with the results of measurements it is found to give satisfactory agreement. As an application a new method is described of rectifying an alternating voltage of high frequency.

1310: A. van Kreveld and J. A. M. van Liempt: Measurements on dark-room illumination (Physica **5**, 345 - 373, May 1938).

The conditions are examined which must be satisfied by dark-room lamps for the successful development of negatives. The sensitivity of the eye to contracts was determined for different colours from green to dark red under the conditions prevailing in a dark-room. The sensitivity of different photographic emulsions was measured for the whole visible spectrum. From these measurements the most satisfactory kind of light for orthochromatic and panchromatic emulsions was determined. Finally the energy was calculated of a dark-room lamp which caused an inadmissible decrease in photographic contrast, from which the quality of any given dark-room lamp follows.

1311: R. Houwink: The yield value (Second report on viscosity and plasticity; Verh. kon. Ned. Akad. Wet. A'dam, afd. Natuurk. 1e Sect. **16**, No. 4, 185 - 240, 1938).

This report, which was given before the Kon. Ned. Akademie van Wetenschappen, deals with the yield value, by which is meant the lowest shearing stress at which a substance undergoes permanent deformation and therefore begins to behave non-elastically. For several simple cases diagrams are given of the way in which the deformation depends upon the shearing stress, and the experimental observation of the yield value is discussed as well as the conditions under which "flow" may occur. A model is described to illustrate the phenomena of plastic deformation exhibited by crystalline substances. The dependence of the yield value on temperature and on the degree and rate of deformation is discussed. The significance of migrations of individual atoms, ions and molecules is dealt with, and conclusions are drawn about the plastic behaviour and the yield value of crystalline substances. The yield value of amorphous substances and suspensions is then dealt with, while finally the relation between yield value and coefficient of viscosity is discussed.

1312: M. J. O. Strutt and A. van der Ziel: Messungen der komplexen Steilheit moderner Mehrgitterröhren im Kurzwellengebiet (El. Nachr. Techn. **15**, 103 - 111, April 1938).

A brief explanation is given of the influence of a phase shift between grid voltage and anode current in receiving valves on the limit of stability of an amplifier stage and in receiver valves on the excitation of alternating voltages. The measuring arrangement and the sources of error are discussed in the measurement of the complex slope of valves with several grids in the short wave region. The method of measurement resembles that of Llewellyn. The results of the measurement for high frequency pentodes are given and compared with calculations, with which they are found to give satisfactory agreement. Measuring results on hexodes are also briefly discussed. Measurements of the phase shift of the slope were also carried out on amplifier valves based on secondary emission, and the results agree satisfactorily with the calculations. The effect of a super secondary emission according to Malter is dealt with briefly and illustrated by measurements on the decrease of slope in the high-frequency region. Finally the influence is discussed of the phase shift in the slope on the frequency dependence of the characteristic admittances of the tubes, whereby it is found that the formulae previously given by the writers remain correct in the approximation in question.

1313: J. Bergmans: Light reflection by road surfaces (Diss. Delft 1938; 111 p.).

The coefficient of brightness is an important quantity in the characterization of the nature of light reflection on road surfaces. By this term is meant the quotient of the brightness in the direction of vision divided by the intensity of illumination of the surface viewed. An arrangement is discussed for the measurement of coefficients of brightness of road surfaces and the results obtained with it are given. Conclusions are drawn as to the construction of road surfaces with a profile such that water is drained away so rapidly that even during a shower they never or scarcely ever become inundated. In order to obtain a more uniform brightness of road surface it is finally proposed to place the sources of light higher, which causes only a slight decrease in the average brightness.

1314: J. H. de Boer: The electrons in non-metals in their relation to thermo-electric properties (Ingenieur **53,** E26 - 31, June 1938).

The conduction of heat and electricity by non-conductors is discussed in relation to the variation of potential in the crystal lattices concerned, in which there may be electrons with different energy contents while other energy contents are „forbidden".

1315: J. E. de Graaf: X-ray interpretation in the electric welding of steel (Weld. Industr. **6,** 41 - 45 and 93 - 96, March and April 1938).

A survey is given of defects in welding which may be detected by X-ray examination. It is found that the pictures obtained are often sufficiently characteristic to make possible the detection of the cause of the defect as well (*cf.* Philips techn. Rev. **3,** 93, 1938).

1316: R. Vermeulen: The Philips-Miller method of recording sound (J. Soc. Motion Pict. **30,** 680 - 693, June 1938).

In this article a survey is given of the Philips-Miller system of sound recording, for which we may refer to the series of articles on this subject in the first volume of this periodical (Philips techn. Rev. **1,** 107, 135, 211, and 230, 1936).

1317: M. J. O. Strutt: Electron transit time effects in multigrid valves. Measurements on short waves. (Wirel. Eng. **15,** 315 - 321, June 1938).

The electron coupling in octodes for frequency transformation can be represented by a unilateral negative capacitance of the oscillator grid (1st grid) to the input grid (4th grid), at least when the phase shift caused by the electron transit time may be neglected. If this phase shift does play a part, a unilateral negative resistance in parallel with this negative capacitance must also be taken into account. The transit time can be calculated from measurements of this negative resistance and capacitance. These measurements were carried out at a wave length of 31 metres; their results are found to be in satisfactory agreement with the roughly calculated transit time. Moreover the slope was measured between the oscillator grid and the oscillator anode of an octode, and it was found that a phase shift of about 60° occurs in it at a wave length of 9 m due to the transit time. The calculated transit times agree quite well with the values measured. In addition there is a discussion of the transit time current which, in the form of a low direct current, flows to the first grid of a tetrode with several volts negative bias, while the anode is connected in series with a negative bias and a high-frequency generator, and the screen grid is positive. Finally it is indicated, how this transit time current can be used for a new kind of detection.

1318: Balth. van der Pol and H. Bremmer: Ergebnisse einer Theorie über die Fortpflanzung elektro-magnetischer Wellen über eine Kugel endlicher Leitfähigkeit (Hochfrequenztechn. u. Elektroakust. **51,** 181 - 188, June 1938).

This is a summarizing review of the theory of the diffraction of electromagnetic waves around a sphere which was dealt with in detail in Phil. Mag. **24,** 141 and 825, 1937 (*cf.* **1264**) and in Phil. Mag. **25,** 817, 1938 (*cf.* **1338**). Formulae are given from which the field of a transmitter can be determined in a simple way. To these are added a complete set of practical curves for the field at the surface of the earth when the transmitter is on the ground, not only for propagation over the sea but also over land of average conductivity. These curves relate to the whole wave length range from 1 m to 20 km.

In Sept. 1938 appeared:

Philips Transm. News **5,** no. **3**:

C. G. A. von Lindern and G. de Vries: Decimetre wave radiolink between Eindhoven and Nimeguen.

A new Philips Studio equipment.

Tj. Douma and P. Zijlstra: The use of the cathode ray oscillograph for the determination of transmitting valve characteristics that cannot be measured in the ordinary way.

VOL. 3 No. 12 DECEMBER 1938

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RELATING TO THE PRODUCTS, PROCESSES AND INVESTIGATIONS OF
N.V. PHILIPS' GLOEILAMPENFABRIEKEN

EDITED BY THE RESEARCH LABORATORY OF N.V. PHILIPS' GLOEILAMPENFABRIEKEN, EINDHOVEN, HOLLAND

ZIRCONIUM AND ITS COMPOUNDS WITH A HIGH MELTING POINT

by J. D. FAST.

In this article the preparation of zirconium and its mechanical, chemical and electrical properties are dealt with. The most important applications of zirconium and its compounds with high melting points are then discussed. Zirconium is often used in discharge tubes where its high melting point, small value of secondary emission and, when used as a getter, its property of being able to take up large quantities of oxygen, nitrogen and hydrogen are of significance. Zirconium oxide and zirconium carbide are important as refractory materials.

Introduction

Even since the beginning of the development of the manufacture of electric incandescent lamps, much research has been carried out on metals with high melting points in the laboratories of such factories. The metals in the following groups of the periodic system drew particular attention:

1) The platinum group: platinum, iridium and osmium, the last of which was used for some time as filament in the sources of electric light.
2) The fourth main group: titanium, zirconium, hafnium and thorium.
3) The fifth main group: vanadium, niobium and tantalum, the last of which was used for some time in so-called tantalum lamps.
4) The sixth main group: chromium, molybdenum and tungsten.

At the beginning of the development of the electric incandescent lamp industry (around 1880) practically nothing was known about the mechanical and physical properties of the last three groups mentioned, and very little about their chemical properties. The interest of this industry, however, contributed very much to the research which has led to the fairly extensive knowledge of these metals which we now possess. This interest was at first chiefly directed toward the melting point and the speed of evaporation, but when the electric lamp industry also began to manufacture X-ray tubes, transmitting and receiving valves for radio purposes, rectifier valves etc., it began to include all the properties of the metals under consideration.

The technical result was the present general use of tungsten as filament in incandescent lamps, and the use of this element in the place of all the other substances which were used for that purpose. Several others of the metals mentioned, such as zirconium, tantalum and molybdenum, have also found uses, although less important ones.

The aim of this article is to make several statements about the properties and applications of the element zirconium and several of its compounds. We shall discuss successively the metal, the oxide (ZrO_2) and the carbide (ZrC).

The metal in the pure state is used chiefly in transmitting valves and other discharge tubes, where use is made of its great chemical affinity for various other elements such as oxygen, nitrogen, carbon and hydrogen. The oxide is chiefly used as a refractory and as an opacifier for glazes and enamels. The carbide is still little used, but it shares the interest which has existed during the last ten or fifteen years for the metallic carbides with a very high melting point and very great hardness.

Brittle and ductile zirconium

Twenty-four years ago the first scientific publication from the Philips Laboratory appeared [1]). It contained statements about the preparation of metallic thorium, uranium, zirconium and titanium. Zirconium was prepared by reduction of zirconium tetrachloride ($ZrCl_3$) with sodium. The metal was

[1]) D. Lely and L. Hamburger, Herstellung der Elemente Thorium, Uran, Zirkon und Titan, Z. anorg. allg. Chem. 87, 209, 1914.

obtained in the form of a powder and the size of the grain depended upon the reduction temperature, the nature of the chloride, etc. Later the method was improved upon by previously subliming the chloride in hydrogen and thus obtaining it in large compact pieces. If this sublimed chloride is used, and if it is reduced in large portions at a fairly high temperature, the metal is obtained in the form of chunks with dimensions of several centimetres, which are, however, porous. This method of preparation produces metal of great purity and is still one of the best methods of preparing zirconium in the form of a powder or of porous chunks. The best samples were found upon analysis to have a purity of 100 per cent. The separate grains showed a certain ductility (mechanical deformability). It was, however, very remarkable that when a rod was pressed from the pure metal powder, according to the method used in working tungsten, and when this rod was sintered at a high temperature in a very high vacuum, the rod showed practically no sign of ductility. Working to sheet or wire by rolling, hammering and drawing was therefore impossible. Since however chemical analysis indicated a very high degree of purity, as was mentioned above, it was for many years generally assumed that zirconium is a very brittle metal and must be considered to belong to the so-called half-metals.

However, thirteen years ago an entirely new method of preparing zirconium in the form of rods was discovered in this laboratory [2]), namely by thermal decomposition of zirconium tetraiodide on a glowing filament (see below) and the rods as prepared were found to possess a high degree of ductility. The deformability is so great that it is now even possible to cold-work rods of 7 mm thickness to very thin wire and sheet. The crystal structure of ductile zirconium was found to be the same as that of brittle zirconium, so that it was not a question of two allotropic modifications. It has therefore to be assumed that the brittle zirconium contained one or more impurities of which chemical analysis gave no indication.

Only in recent years has it become possible to arrive at a reasonably satisfactory explanation of this phenomenon. It has been found that zirconium is able to take up quantities of oxygen and nitrogen in solid solution. When these elements are present in the dissolved state they have a very detrimental influence on the mechanical properties of metals, so that very small amounts are already enough to make the metal brittle. Titanium which can also contain considerable quantities of oxygen and nitrogen in solid solution, exhibits the same phenomena as zirconium. Thorium, however, which belongs to the same group of the periodic system as titanium and zirconium, is found to have no appreciable dissolving power for oxygen and nitrogen. The result is that this last metal does not exhibit the phenomenon mentioned: thorium rods can also be obtained in the ductile state by compression and sintering of the powder.

We must assume that grains of metal obtained by the reduction of the chlorides of the three metals in question are covered with a film of oxide and (or) nitride. In the sintering of rods pressed from the powders, the oxygen and nitrogen dissolve in the metal in the case of zirconium and titanium, and the metal thus becomes brittle, while in the case of thorium, the oxide and nitride remain present as such in the rod. It is true that in the sintering of pressed thorium rods a change in the structure also appears, but this change in the structure has a favourable influence on the mechanical properties. It consists namely in the fact that the oxide or nitride is collected into separate grains, which are then present after sintering surrounded by a coherent basic mass of ductile metal. This makes it understandable that much greater quantities of oxygen and nitrogen are required to render thorium brittle than in the case of zirconium and titanium.

The grains of zirconium and titanium obtained by reduction usually contain small traces of oxygen and nitrogen already dissolved, so that each individual grain has a ductility considerably lower than that of the absolutely pure metals, but considerably higher than that of the pressed and sintered rods.

Preparation of ductile zirconium in rod form

Ductile zirconium is prepared by thermal decomposition of zirconium tetraiodide. At a temperature of 431°C the vapour pressure of this compound reaches a value of one atmosphere. If the gaseous iodide comes into contact with a surface whose temperature is higher than 1100°C, it decomposes partially into zirconium and iodine. When the process is carried out in practice the decomposition takes place on a thin wire which is heated by the passage of current to a suitable temperature, for instance 1300°C. The reaction takes place in an apparatus made of pyrex glass into which thick tungsten terminals are fused. The required amount of crude

[2]) J. H. de Boer und J. D. Fast, Über die Darstellung der reinen Metalle der Titangruppe durch thermische Zersetzung ihrer Jodide. I. Zirkonium, Z. anorg. allg. Chem. **153**, 1, 1926.

zirconium is introduced into this apparatus in the form of porous chunks (obtained by the reduction of zirconium tetrachloride or of sodium zirconium fluoride with sodium) together with a relatively small amount of iodine. The apparatus is then evacuated to a pressure of about 10^{-6}mm. During evacuation the crude zirconium is heated to drive off adsorbed gases, while the iodine is cooled. After outgassing the apparatus is sealed off and heated in an electric oven. At a temperature of about 100°C (depending on the nature of the crude zirconium) the iodine combines with part of the zirconium to give ZrI_4, and the substance evaporates and is decomposed on the wire into zirconium and iodine when the temperature of the wire is sufficiently high. In this way crystals of zirconium begin to grow on the wire.

The vapour pressure of the iodide at 100°C is so small that it may be neglected, so that upon heating the core wire to 1300°C, for instance, no appreciable growth takes place. At an oven temperature of 150°C the vapour pressure is already so high (about 0.001 mm Hg) that zirconium slowly begins to be deposited on the glowing wire. With increasing temperature of the oven the speed of growth increases and reaches a high value at 250°C. The vapour pressure of the tetraiodide then amounts to about 0.5 mm Hg. From about 250°C, however, the ZrI_4 pressure rises only slightly, because of the fact that above this temperature the tetraiodide begins to react with the excess crude zirconium giving the tri-iodide (ZrI_3), which has no appreciable vapour pressure in this temperature range. At still higher temperatures the di-iodide (ZI_2) is also formed, a compound which also has no appreciable vapour pressure at the temperatures attainable in pyrex glass. Above 310°C the tri-iodide decomposes according to the equation

$$2\ ZrI_3 \rightleftarrows ZrI_2 + ZrI_4.$$

while above 430°C the di-iodide decomposes according to the equation:

$$2\ ZrI_2 \rightrightarrows Zr + ZrI_4$$

At higher temperatures therefore the pressure of ZrI_4 is determined by these equilibria.

The result of these phenomena is that the speed of growth does not increase appreciably above 250°C, and that above 350°C it finally begins to decrease again.

Fig. 1 gives a graphic representation of the speed of formation of ductile zirconium as a function of the oven temperature. The temperature is plotted as abscissa, and as ordinate the weight of the zirconium rods which were obtained after a time of growth of 30 hours. In each separate experiment the oven temperature was kept constant during this time of growth. All the apparatus which were used in these experiments were filled with 250 g. of crude zirconium from the same preparation (obtained by reduction of the chloride) and with 12 g. of iodine. The wire temperature was the same, 1300°C, in all the experiments.

For practical uses zirconium is made in rods about 7 mm thick which weigh about 200 gr. The oven temperature is kept at 250° to 350°C during the preparation. A tungsten wire 40 microns thick is often used as core wire. The tungsten content of the zirconium rods then amounts to 0.01 per cent. For applications in which this tungsten content would be detrimental a core wire of zirconium itself is used.

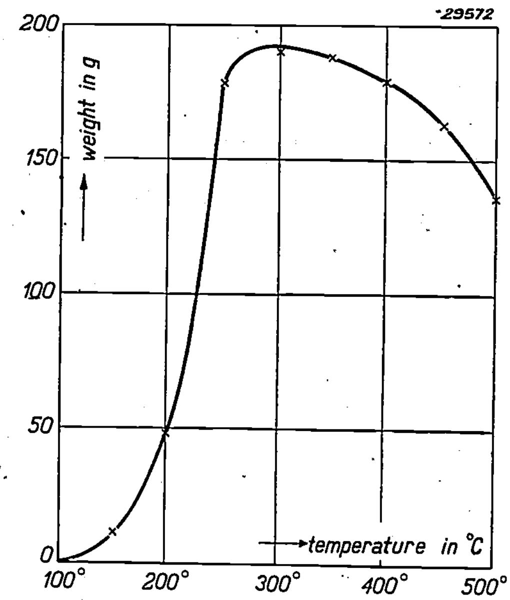

Fig. 1. Speed of formation of zirconium obtained by decomposition of zirconium tetraiodide, as a function of the oven temperature. The weight obtained at the corresponding oven temperature and at a wire temperature of 1300°C after a time of growth of 30 hours is plotted as ordinate.

As stated above only a relatively small amount of iodine is used in the preparation of ductile zirconium. It serves only as means of transporting the metal from the crude zirconium to the growing rod. It is continually freed at the growing rod, combines again with part of the excess crude metal and "loaded with zirconium" it enters the gas atmosphere again. Care must be taken that the crude zirconium does not contain any great amount of substances like iron, aluminium and silicon, since these substances also form relatively volatile iodides and their presence might lead to contamination of the rods. The oxygen and nitrogen content of the crude zirconium on the other hand is not harmful, since these substances remain behind in the crude metal.

The rods which are prepared according to the process described have a completely compact structure. The specific weight (6.54) determined, with the aid of a hydrostatic balance, agrees accurately with the specific weight deduced theoretically from röntgenographic data. The surface of the rods is composed of well formed crystal planes. *Fig. 2* shows a photograph of zirconium rods 7 mm thick.

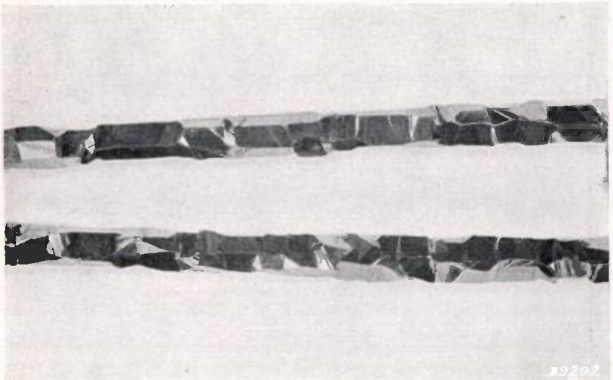

Fig. 2. Zirconium rods 7 mm thick.

Mechanical working of zirconium

We have already mentioned that the zirconium rods prepared by decomposition of the iodide are so ductile that they may be deformed into thin wire and sheet without it being necessary to heat them. In the preparation of zirconium wire the rods are first swaged with swagging machines to wire of about 0.5 mm thickness. This treatment can be carried out entirely without heating. From the diameter of 0.5 mm the wire is drawn through diamond dies to smaller diameters. If an attempt is made to carry out this process also entirely cold, great difficulties arise. A grating noise is heard during the drawing, while the wire repeatedly breaks. If the wire is examined after drawing under a microscope, grooves are seen in the surface along the whole length. The cause of this phenomenon is that even with careful lubrication particles are torn loose from the surface due to the adhesion between the metal and the surface of the die. These particles remain stuck to the die and are the cause not only of the seizing but also of the continuous rise in the required drawing force to a value which exceeds the tensile strength [3]). If the wire is hot drawn (temperature about 500°C) an oxide film is formed on the wire which is indeed very hard, but which makes drawing possible by preventing seizing. Since we are concerned with a surface phenomenon we can surround the wire with a metal covering instead of the oxide covering. The entirely cold hammering and drawing of rods about 7 mm thick to wire of 30 microns for instance is possible on a technical scale only when a covering of iron, nickel or copper is given to the rods. This covering is then removed chemically (by solution in acid) after the working. Zirconium wire prepared according to this method is entirely free of oxygen and nitrogen, while hot drawn zirconium wire always contains very small quantities of these elements.

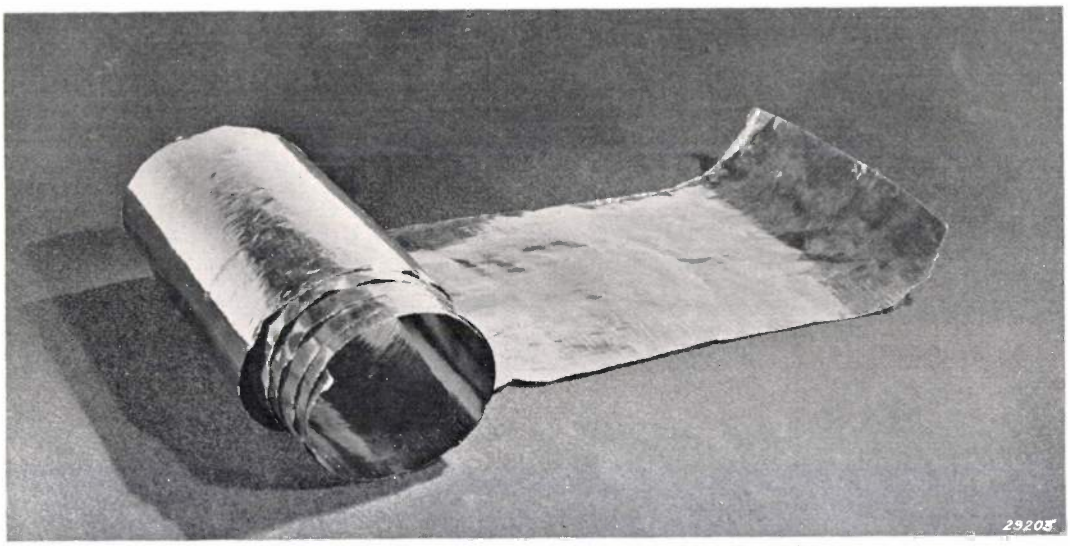

Fig. 3. Zirconium sheet 20 microns thick, obtained by cold rolling of the material from the thickness of 0.6 mm.

The rolling of the rods to sheet can also be done cold. Usually, however, the rods are hot rolled to a thickness of 1 to 0.5 mm. The surface is then cleaned and the rest of the rolling is done cold. It was found possible in this way to make strips 20 microns thick, 30 cm wide and 1 to 2 m long.

[3]) It might be expected that the cause of the difficulties in drawing lays in a mechanical hardening occurring during the process. This was, however, found not to be the case, since the same difficulties are also encountered when a zirconium wire is drawn immediately after being annealed in a high vacuum.

If the rolling is done on polished rollers, fine specular sheet is obtained. *Fig. 3* shows a photograph of zirconium sheet 20 microns thick which was rolled from the thickness of 0.6 mm.

It has been found possible to make cold rolled shell bodies of zirconium by the deep stamping process. No particular difficulties arise during drawing when the zirconium sheet which serves as raw material has been rolled in the correct way. *Fig. 4* shows a photograph of three different types of drawn products made of zirconium. *Fig. 5* shows the various stages in making such an article. The processes were carried out cold. Just as in the drawing of wire it was found that bare metal is very difficult to draw, since seizing takes place not only with cold-rolled sheet but also with sheet annealed in a high vacuum. If, however, the material is heated previous to the stamping for a very short time to 650°C in air, a film of oxide is formed which facilitates drawing. The best results are obtained upon heating the metal for only a few seconds in the colourless flame of a Bunsen burner, and then quenching it in water. The metal is then in the form of very fine crystals, since the cold-worked sheet recrystallizes upon heating, while due to the extremely short time of heating there is no appreciable growing together of the very fine crystals formed. This fine grained structure is an advantage for the stamping process.

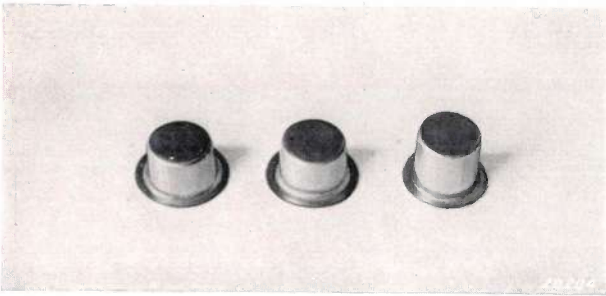

Fig. 4. Shell bodies made of zirconium sheet by application of a deep drawing process.

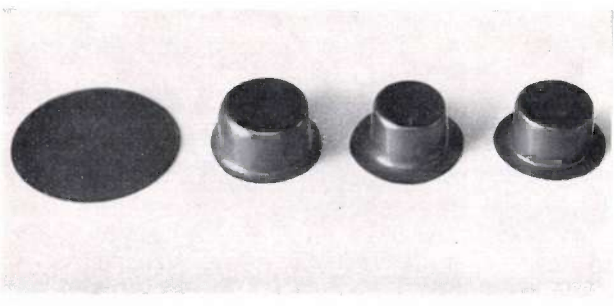

Fig. 5. Different stages in the drawing of shell bodies. The manufacture of the cap (right) from the zirconium sheet (left) requires three processes which can be carried out cold.

Chemical properties of zirconium

The most characteristic chemical property of metallic zirconium is its great affinity for various other elements. Finely-divided zirconium powder ignites when heated above about 200°C in the air, and burns with a blinding white light to give zirconium oxide (ZrO_2). When zirconium is heated in nitrogen a very stable nitride is formed having the formula ZrN. With hydrogen the compound ZrH_2 is formed, which however decomposes again into its elements at relatively low temperatures (above 600°C). *Table I* gives the heats of formation of the three compounds mentioned.

Table I

Compound	Heat of formation k. cals. per mol.
ZrO_2	258
ZrN	82
ZrH_2	40.5

If the metal is not allowed to react with an excess of gas, but with quantities in a closed space, it is found that large quantities of gas can be taken up without a new phase being formed. In the reaction with nitrogen the cubic phase of the nitride is formed in addition to the hexagonal phase of the metal only after the nitrogen content has exceeded about 20 atom per cent. The first 20 per cent are thus taken up in solid solution. For oxygen the limit of solid solubility even lies at about 40 atom per cent, while for hydrogen it is already reached at 5 atom per cent.

In the investigation of the systems zirconium-oxygen, zirconium-nitrogen and zirconium-hydrogen it is best to begin with ductile zirconium rods or bare wires which are made by cold working of the rods. In the case of the systems with oxygen or nitrogen the wires must be heated for some time at a high temperature after taking up the different amounts of gas in order to reach complete homogeneity. The simplest method of heating is by the passage of a current. When the content of gas is not too high the solutions formed cannot be distinguished in appearance from the pure metal. When unworked rods are used the solutions still exhibit the same crystal planes and the same high metallic lustre. *Fig. 6* shows two pieces of the same zirconium rod (thickness about 3 mm). The shorter piece has been kept in the original state and is therefore quite pure and very ductile. The longer contains 13 atom per cent of nitrogen and is very brittle. Superficially, however, the two pieces cannot be distinguished from each other.

If wires containing several atom per cent of oxygen, nitrogen or hydrogen are heated in a high vacuum, it is found that the hydrogen can be driven out quite easily. (The removal of the last traces, however, takes a long time even at a high temperature). Oxygen and nitrogen, however, cannot be driven out. Upon increase of temperature the gas pressure remains zero, and at temperatures near the melting point metal atoms begin to evaporate, while the oxygen or nitrogen remains in the wire. These gases are therefore also bound by strong chemical forces when they are present in the metal in the dissolved state.

The specific weight of zirconium is increased by the introduction of oxygen or nitrogen atoms into the metal lattice. From this it may be concluded that these atoms lie in the open spaces of the metal lattice. They are probably quite irregularly distributed throughout the lattice. In *table II* are the results of several determinations of volume and specific weight (of rods 3 mm thick) before and after the taking up of known amounts of nitrogen or oxygen.

Table II

Composition in atom per cent	Volume increase	Specific weight
100% Zr		6.540
98 % Zr + 2 % O	0.15%	6.555
98 % Zr + 2 % N	0.23%	6.546
90 % Zr + 10 % O	1.05%	6.600
87 % Zr + 13 % N	1.72%	6.577
78.5% Zr + 21.5% O	1.51%	6.755

The temperature coefficient of the electrical resistance decreases upon the solution of oxygen or nitrogen in the metal. The values given in *table III* were measured.

Table III

Composition in atom per cent	Temperature coefficient of the electrical resistance
100% Zr	0.00445
98% Zr + 2% O	0.00264
98% Zr + 2% N	0.00261

For oxygen and nitrogen therefore the same lowering of the temperature coefficient was found.

Applications of metallic zirconium

Mixed with one or more oxidation agents finely-divided zirconium is used as a flash light powder.

Several other applications of zirconium are also based upon the property of this element of reacting easily and rapidly with various gases, with the result that the gas pressure falls to zero if there is an excess of metal. For this reason it is used as a getter in discharge tubes, especially in transmitting valves. The zirconium must be placed in the tube at such a spot that it assumes the required temperature at the moment when its function of combining with the gas must be exercised, or it must be at such an electric potential that the gas ions formed strike the zirconium and are then held bound. The favourable action of the zirconium is manifested in two ways. In the first place the times of evacuation during manufacture are considerably shortened because the zirconium also acts as a pump. In the second place the small amounts of gas which may be freed when the discharge tube is in use are immediately bound. The temperature which the zirconium must assume depends upon the nature of the gas to be absorbed and in addition on the nature of the zirconium itself. If ductile (and therefore compact) zirconium is used, and if the freeing of gases like oxygen,

Fig. 6. Two pieces of the same zirconium rod. The shorter piece is chemically pure and very ductile. The longer piece was heated in an atmosphere of nitrogen, it contains 13 atom per cent of nitrogen and is very brittle. In appearance the two pieces cannot be distinguished from each other.

nitrogen, carbon monoxide and carbon dioxide must be taken into account, it is desirable to have the temperature as high as possible (1400°C for example). Hydrogen, however, is given off again for the most part at this high temperature (an equilibrium pressure is established). The optimum temperature for taking up hydrogen is 300 to 400°C, in which temperature range a solid solution of hydrogen in zirconium has no appreciable gas pressure, while the absorption velocity is already very high. Therefore if provision must be made not only against the freeing of oxygen, etc., but also against the freeing of hydrogen and water vapour; it is desirable to have present in the tube, in addition to the zirconium heated to a high temperature, a portion of zirconium with a temperature of 300 to 400°C. It is often possible to combine these two requirements by introducing the zirconium at such a spot that different parts of the metal are raised to different temperatures so that the temperature range extends from 300° to for instance 1400°C.

In many cases metallic zirconium can also be used in the form of a fine powder which may for instance be applied to the anode of a transmitting valve. The finely-divided zirconium reacts with the different gases at a considerably lower temperature than compact zirconium.

The fact that zirconium may contain large quantities of oxygen and nitrogen in solid solution is a great advantage when the metal is used as a getter. At sufficiently high temperature there is no layer of oxide or nitride formed which hinders further taking up of gas, but the oxygen or nitrogen diffuses to the inside and the metal surface always remains clean.

A quite different property of zirconium which is important for its use in transmitting valves is the small value of its secondary emission. By making the control grid of a transmitting valve of zirconium, or covering it with zirconium, the secondary emission of this grid can be diminished to a negligible value. The zirconium used for this purpose also retains its property of combining with gases. By covering a grid of molybdenum or tungsten with zirconium oxide, the secondary emission of the grid may also be sufficiently reduced. For further particulars about this action of zirconium the reader is referred to an article by H. G. Boumeester in this periodical (Philips techn. Rev: **2**, 115. 1937).

Although zirconium, as may be seen from the above, combines readily with all kinds of elements or reacts with all kinds of compounds at higher temperature (depending on the form of the metal), the compact metal is remarkably inert at ordinary temperatures. This inertness must be ascribed to a film of oxide which is formed on the surface and cannot dissolve in the metal at ordinary temperatures. This oxide film, which is quite invisible, protects the metal to such an extent, that zirconium keeps its lustre under all circumstances and is not attacked by acids, bases or other corrosive agents (except hydrogen fluoride). Thanks to this property zirconium is a very suitable metal for use in chemical processes where other metals would quickly corrode. For example in the etching or cleaning of metal articles which must be immersed successively in alkali and acid baths, zirconium wire is used to hang up the articles as it does not corrode and does not contaminate the baths, while in electrolysis it may also serve as supply line for the current.

The refractory compounds of zirconium

The oxide, the nitride and the carbide of zirconium melt at considerably higher temperatures than the element itself, as may be seen from *table IV*.

Table IV

Substance	Melting point, °C
Zirconium, Zr	1860°
Zirconium oxide, ZrO_2	2700°
Zirconium nitride, ZrN	2980°
Zirconium carbide, ZrC	3530°

The oxide dissolves poorly in glass and is therefore used to render glaze and enamel opaque. It is also used in the manufacture of very refractory articles in the form of crucibles, tubes, boats, fire bricks, insulators, etc. These articles can be used at temperatures considerably higher than 2000°C. In addition to its high melting point the small coefficient of expansion and the chemical stability are important in this connection. If the products are made of pure ZrO_2, they show a tendency to develop cracks upon heating due to the transition at 1000°C from a monoclinic to a tetragonal modification, which is accompanied by a contraction of 7 per cent in volume. If the zirconium oxide is mixed with a few per cent of magnesium oxide, a homogeneous cubic solid solution is formed upon heating above 1700°C, which is subject to no changes in modification, and which therefore can be used for the manufacture of very highly resistant refractory articles. Instead of magnesium oxide other substances may also be used, calcium oxide or thorium oxide for example. Beryllium oxide with zirconium oxide forms an eutectic system without the formation of solid solutions. With the help of this compound it is therefore impossible to suppress the reversible change in modification of ZrO_2, since we have to do with mechanical mixtures in the whole system. However, when BeO is used in combination with MgO (for instance 2 per cent MgO + 0.5 per cent BeO), it seems to give a more satisfactory result than MgO alone.

Because of the increasingly high requirements being made of many ceramic products with respect to their resistance to heat, it is probable that many possibilities for the application of ZrO_2 will still be found.

Zirconium carbide, ZrC, is one of the few substances which have a higher melting point than tungsten. In the history of the electric light bulb industry this compound has played a part. In addition to a high melting point it has metallic electrical conductivity so that it could be used for filaments of electric lamps. Upon the invention of the tungsten lamp, however, the compound passed out of use. It was discovered later that a mixture of 4 parts of tantalum carbide and 1 part of zirconium carbide forms a material with the highest melting point of any known substance. The two carbides form an uninterrupted series of solid solutions with a maximum melting point at the composition mentioned. The melting point of this mixture is almost 4000°C (At the same high temperature a mixture of 4 parts tantalum carbide and 1 part hafnium carbide melts). Zirconium carbide is also one of the metallic carbides with a very great hardness, for which compounds there has been great interest in recent years in connection with the important applications which the carbides of tungsten, tantalum and titanium have found in the working of metals.

QUANTITATIVE CONSIDERATIONS OF ELECTRIC WELDING

by J. ter BERG.

In ordinary welding with a rod provided with an oxidizing coating, it could be shown that 82 per cent of the fused metal core is used in the weld, 12.5 per cent is lost by spattering and about 4.5 per cent is lost due to oxidation by the coating, while evaporation plays only a minor part. If instead of the ordinary atmosphere, nitrogen or compressed air is used as gas environment, little change is observed: more or less iron is then also oxidized by the flux. When a mixture of hydrogen and nitrogen is used, this reaction does not take place at all. It has been shown clearly by these experiments, that in the attempt to increase efficiency care must be taken not only to have a low loss from spattering, but also that the coating of the rod should possess little or no oxidizing power, or that this coating should give off gases during welding which form a reducing atmosphere around the arc.

Introduction

It is a well known fact that in electric welding a certain percentage of the metal used is lost in some way or other. What may be the causes of this loss of material, and to what extent each cause may contribute to the total loss, has not, as far as is known to the author, been dealt with in the literature in the case of coated welding rods. The experiments and calculations described below were carried out for the purpose of procuring data on this problem.

The term efficiency is used here, and by the efficiency of a welding rod we mean the ratio between the weight of the metal found in the weld and the weight of the core metal consumed in making the weld.

The rods specially made for our experiments contained a malleable iron core with a diameter of 5 mm, while the coating consisted of ferrous and ferric oxide, quartz and aluminium silicate with potassium water glass as a binder. The average efficiency of several tests carried out under ordinary conditions was low and amounts for these rods about 82 per cent.

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The question is, what has happened to the remaining 18 per cent? The following causes of loss of metal may be listed:

a) evaporation
b) spattering
c) reactions of molten metal with liquid slag.

We shall deal with these three factors in succession.

Evaporation

Quantitative data on this factor were obtained by sliding a copper jacket over the piece of work during welding. The vapours developped are deposited on the copper. Although the amount of evaporation of core and coating were determined together in this way, it was found from the weight of material deposited - which consisted chiefly of iron oxides - that the evaporation of the metal plays a very small part, and cannot in any case account for a loss of core metal of more than 0.5 per cent.

Spattering

The loss of metal due to spattering was determined by making a weld 2 cm wide on a strip 2.5 cm wide, with the strip laid on a large thick copper plate. In this way practically all of the spattered drops fell on the copper plate to which they do not adhere, so that they may easily be removed and weighed. The diagrammatic representation in *fig. 1* illustrates the arrangement employed.

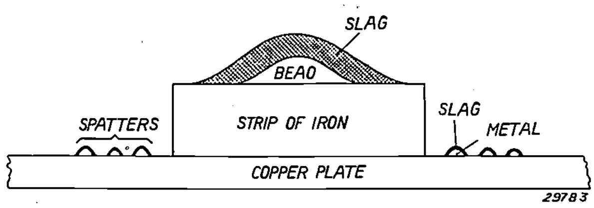


Fig. 1. Arrangement for the determination of the amount of metal spattered away during welding.

Before welding is started, the weight is determined of the strip, of the iron core and of the coating. The weight of the latter could be determined by weighing the iron core before and after it was coated. The following weights were determined after welding:

1) The weight of the remainder of the iron core. This determined the weight of metal fused down;
2) The weight of the remainder of the coating. This gave the weight of the coating fused down;
3) The weight of the slag on the weld (*fig. 1*) which could be separated quantitatively from the metal;
4) The weight of the strip with the weld, from which the weight of the weld and the efficiency could be calculated;
5) The weight of the spatters which consisted of drops of metal surrounded by a layer of slag. Under the entirely reasonable assumption that the ratio of weights of metal and slag is the same for the spatters as for the weld, the weight of the spattered metal could be calculated.

This is presented in a clearer form in *table 1*, in which are given the results obtained in welding with three rods of the type described above. It also appears from the table that the agreement among the three columns is so good that the average values may certainly be assigned real significance.

Table I

	I	II	III	average
Fused metal	52.5 g	53.4 g	55.5 g	53.8 g
Fused coating	18.0	19.2	19.5	18.9
Metal in the weld	43.7	44.1	44.8	44.2
Slag on the weld	18.4	18.8	18.0	18.4
Total weight (metal + slag) of the spatters	8.7	9.8	9.0	9.2
Metal in the spatters *)	6.1	6.9	7.0	6.7
Slag in the spatters	2.6	2.9	2.0	2.5
Total fused slag	21.1	21.7	20.0	20.9
Effic. $= \frac{\text{Metal in weld}}{\text{Metal fused}}$	83.2%	82.6%	80.7%	82.1%

*) Calculated as follows: $\frac{43.7}{43.7 + 18.5} \cdot 18.7 = 6.1$

It may immediately be deduced from this table that the loss by spattering amounts to 12.5 per cent of the weight of metal fused from the rod.

Reactions of molten metal with liquid slag

The average efficiency, as may be seen from table I, was 82.1 per cent. Therefore about 18 per cent of the original core material is lost. As given above, evaporation and spattering cause a loss of $0.5 + 12.5 = 13$ per cent. The disappearance of the remaining 5 per cent of the core material may be explained by assuming that the liquid iron reacts with the Fe_2O_3 of the molten slag. This assumption may immediately be tested qualitatively by the fact that the total weight of the slag is greater than the weight of coating fused, which indicates a slag formation by a part of the molten iron.

Quantitatively also the above assumption corresponds to the values given in table I. The following consideration will make this clear.

By a ferro-ferri analysis it was shown that the 18.9 g of coating melted contained 2.6 g FeO and 8.5 g Fe_2O_3, while there was 12.1 g FeO and 1.2 g

Fe_2O_3 in the slag. Therefore $8.5—1.2 = 7.3$ g Fe_2O_3 disappeared during the welding process. If we now assume that it reacted as follows with the iron:

$$\begin{array}{l} Fe_2O_3 \rightarrow 2\ FeO + O, \\ Fe + O \rightarrow FeO\quad, \\ \hline Fe_2O_3 + Fe \rightarrow 3\ FeO, \end{array}$$

then the 7.3 g Fe_2O_3 has used 2.5 g Fe from the core to form 9.8 g FeO which is taken up in the slag. The fact that the reaction occurred in this way is confirmed by the consideration that according to this calculation the slag should contain $2.6 + 9.8 = 12.4$ g FeO, while analysis shows very good agreement with 12.1 g FeO. Moreover, the whole "welding account" is now balanced, since, from the foregoing calculations, of the 53.8 g of iron which was melted down the following amounts were used:

in the weld:	44.2 g =	82.1%,
evaporated (maximum): . . .	0.3 g =	0.5%,
spattered:	6.7 g =	12.5%,
slag formation with the coating:	2.5 g =	4.8%,
Total . . .	53.7 g =	99.9%.

Now that we have obtained a quantitative idea about the distribution of the metal fused down over the various forms which it takes on during welding, the following point must be noted. One of the great advantages of a coated welding rod over a bare rod is usually considered to lie in the fact that the oxygen and nitrogen of the air are so well kept away during welding that they have no opportunity of contaminating the metal, and thus unfavourably affecting its mechanical properties. It will, however, immediately be seen that in the case here studied, where the coating contains ferric oxide [1]), a primary oxidation of the metal must occur, since oxygen gas is freed during the welding process and comes into direct contact with the liquid iron. Even if the slag prevents the entrance of oxygen from the air, it cannot prevent the oxidation of the iron. The advantages of a coated welding rod over an uncoated one must be sought in other directions. Much more attention must be given to the solution of the oxides present and those newly formed in the liquid slag, while under certain conditions a stream of strongly reducing gas is generated in the coating, which surrounds the arc during welding and may exert a favourable influence on the oxygen content of the metal. In any case the idea that the slag protects the molten metal from the action of oxygen is in its general form incorrect.

It is also a question whether the slight nitrogen content of a weld made with a coated rod can be adequately explained by the mechanical shielding action of the liquid slag and the mantle of gas formed; it is possible that processes similar to the one discussed above also play an important part.

Welding in different atmospheres

The results given in the foregoing were of such a nature that a decision was made to carry out the welding process in other atmospheres than air. For this purpose use was made of a very simple

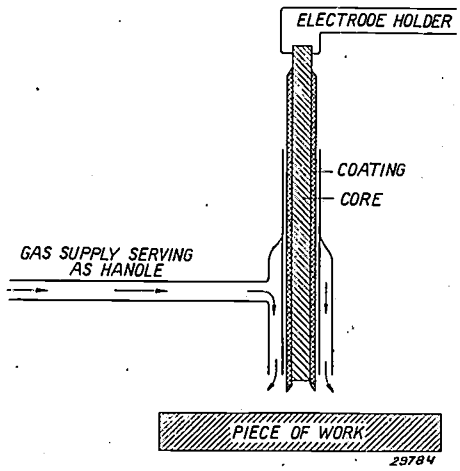


Fig. 2. Apparatus for welding in different atmospheres.

piece of apparatus shown in fig. 2. During welding the apparatus was gradually pushed up, so that as the rod melted off, the arc was surrounded by

Table II

	A ordinary	B nitrogen	C compressed air	D 75% N_2 + + 25% H_2
Metal fused	100 g	100 g	100 g	100 g
Coating fused	35.3	33.2	31.9	31.0
Metal in the weld. . .	82.1	81.7	75.7	84.1
Slag on the weld . . .	34.2	30.2	36.7	24.6
Total weight (metal + slag) of spatters . .	17.1	17.7	17.8	19.0
Metal in spatters . . .	12.5	13.0	11.9	14.6
Slag in spatters	4.6	4.7	5.9	4.4
Total slag melted . . .	38.8	34.9	42.6	29.0
FeO of coating	4.8	4.5	4.3	4.2
Fe_2O_3 of coating . . .	15.8	14.9	14.5	13.9
FeO of slag	22.3	20.5	23.9	15.8
Fe_2O_3 of slag	2.3	1.7	5.5	1.4

[1]) Besides this oxide there are many other substances which are often used in the coating, and which give off free oxygen during the welding process. There are for example the carbonates, which certainly show oxidizing properties at the high temperature of the arc.

the gas in which the test was being made, and the gas was supplied at a definite rate. Compressed air, nitrogen and a mixture of 25 per cent hydrogen and 75 per cent nitrogen were used.

Table II gives the results of these experiments together with those obtained in ordinary welding; the different quantities are here again averages of several observations, and for ease in comparison they are all re-calculated for 100 g of metal fused, so that the amount of spattering and the efficiency for example may be read directly in per cent of metal fused.

We shall deal separately with the four columns of this table.

A. Ordinary atmosphere. During the process of welding 15.8—2.3 = 13.5 g Fe_2O_3 disappeared. This reacted with the liquid iron according to the following equation:

13.5 g Fe_2O_3 + 4.3 g Fe →	17.8 g FeO
There was present in the coating	4.8 g FeO
The slag must therefore contain	22.6 g FeO

Analysis of the slag gave 22.3 FeO, thus good agreement. Therefore 4.3 per cent of the iron went into slag formation, and this is manifested in the increase in weight of slag with respect to weight of the coating melted.

B. Nitrogen atmosphere

Reaction . . 13.2 g Fe_2O_3 + 4.2 g Fe →	17.4 g FeO
There was present in the coating	4.5 g FeO
Total calculated	21.9 g FeO

Analysis showed 20.5 g FeO. This slightly too low result may, however, be explained by the fact that some of the oxygen of the decomposed Fe_2O_3 does not react with molten iron but under the conditions of the experiment escapes as free oxygen or in the form of oxides of nitrogen.

C. Compressed air

From the values found it may immediately be deduced that in addition to the normal oxidation of the liquid iron by Fe_2O_3, the oxygen of the surrounding air also exerts an oxidizing action. A large increase in weight of the slag may be observed with the resulting lowered efficiency.

D. Reducing atmosphere

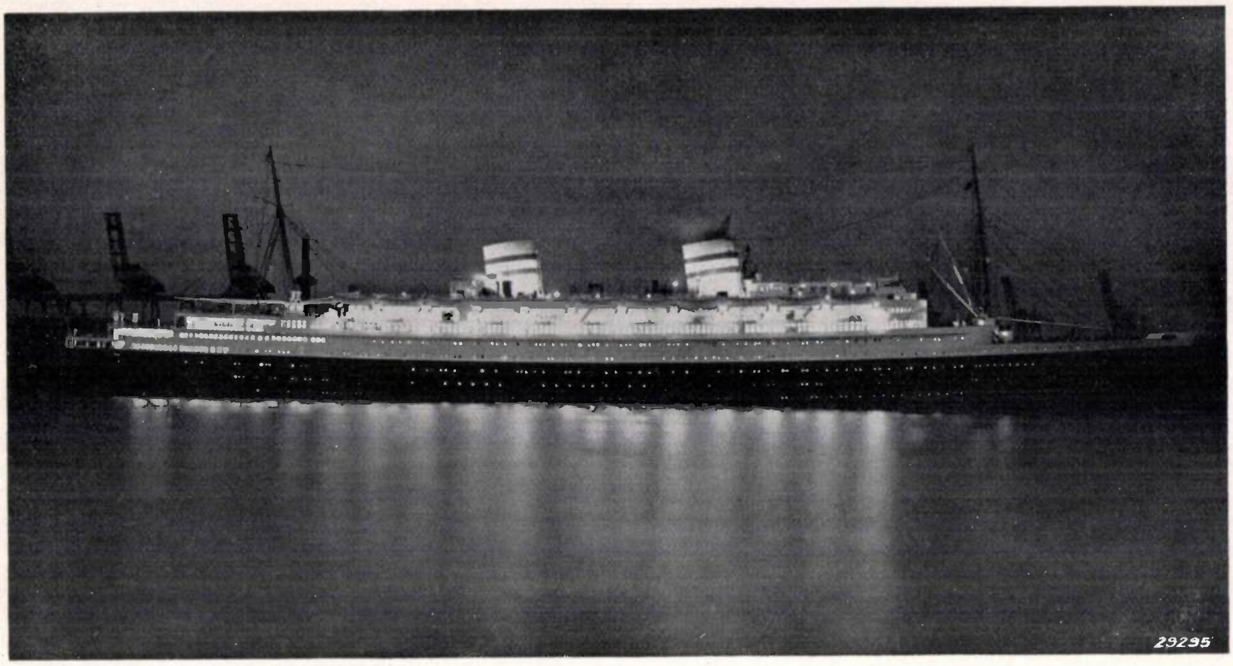
This case is particularly interesting in connection with welding with a so-called "shielded arc", where an attempt is made to surround the arc with a reducing gas. It may be seen from the values given in the table that there was a decrease in weight of the slag, in contrast to the previous cases where an increase in weight of slag with respect to original coating was observed.

This difference can now be explained simply by assuming that the oxygen freed from the Fe_2O_3 no longer reacts with the iron, but with the surrounding hydrogen with the formation of water which escapes, in the following way:

$$Fe_2O_3 + H_2 \rightarrow 2\,FeO + H_2O\nearrow$$

Since the molten iron is protected against oxidation in this way, there will be increased efficiency, and this will be manifested as a decrease in weight of the slag with respect to the weight of coating melted.

Quantitatively also the calculation and experiment may be shown to agree. During the welding process 13.9—1.4 = 12.5 g Fe_2O_3 disappeared and were decomposed into 11.2 g FeO and oxygen. According to the calculation therefore the slag should contain 11.2 + 4.2 = 15.4 g FeO, while analysis gave good agreement with 15.8 g FeO.



THE ILLUMINATION OF PASSENGER SHIPS

by L. C. KALFF. 628.972:629.123.3

In connection with the installation of the lighting system of the s.s. "Nieuw Amsterdam" the most important problems are discussed which are encountered in such a case. A series of photographs give some impression of the results achieved.

It is not so long ago, that the problem of illumination was almost entirely neglected during the designing and building of passenger ships. And when it was finally considered, economy was held to be the main essential, economy in the number of lamps, in the power installed and in the method of illumination. An attempt to achieve technically satisfactory illumination with the avoidance of too great differences between light and dark was not valued, and no attention was paid to uniformity and harmony in the illumination.

Nowadays ideas have been changed, at least with respect to the illumination of the luxurious modern ships. In the case of such ships the main aim is now that the passengers should be offered the same comforts as the guests of a first class hotel. A pleasing, adequate and efficient illumination certainly belongs among such comforts. Fortunately modern lamps and fittings are more economical than earlier ones, and therefore with the same wattage the number of lumens available is larger. Moreover with the present fast liners the power consumed by the lighting system, even when it is very generously designed, constitutes only a small percentage of the energy necessitated for propulsion.

For reasons of safety 65 volts was usually chosen as working voltage for the lighting system. In the present huge ships the diameter of the connections would have to be very great with this working voltage, because of the much greater length of cable. A higher voltage, 125 volts for instance, is therefore chosen. An advantage of this voltage over a still higher one lies in the fact that with it the lamps are much better able to withstand shocks and vibrations.

In designing the illumination of a ship the main interest naturally lies in the various parts of the ship which are used by passengers, *i.e.* cabins, corridors, dining rooms and recreation rooms. In the first place the desired level of illumination must be decided upon, as well as the system for each individual type of room or corridor. Moreover the mutual relation of the intensities of illumination in these different places must so be chosen that no disturbing transitions are encountered in passing from cabin to corridor or from vestibule to hall. Such transitions might lead to dazzle, or, if the transition is from a better lighted portion to a less well-lighted one, it would make a somber and unpleasant impression.

In the cabins (*figs. 1* and *2*), where the passengers must be able to see everything with ease, and where they dress and sometimes even read, a general diffuse lighting with no dark corners is desired. Light-coloured walls may contribute to this effect. An intensity of illumination of 85 lux (8 foot-candles) is certainly necessary.

If a cabin is compared with a good hotel room, then, within certain limits which make the fixing of this value somewhat arbitrary, this level of illumination of 85 lux is arrived at. For ships which sail in tropical waters an exception must be made. In this case the ventilators succeed in reducing the temperature a few degrees only with the greatest difficulties. The heat development of many electric lamps would make the work of ventilation more difficult, and the direct radiation would be particularly disagreeable. In such cases, therefore, small light sources are now used which give sufficient local illumination, above the toilet mirror, on the writing desk and the side of the bed in the form of a reading lamp. A fundamentally different solution would be the use of discharge lamps which, in contrast to ordinary electric lamps, may be said to give "cold light".

Fig. 1

Figs. 1 and 2. For the cabins on board the "Nieuw Amsterdam" an average intensity of illumination of 85 lux was assumed to be necessary. For practical reasons (the lowness of the ceilings) use had to be made of many small units. The cabins have double ceilings so that special built-in plafonniers could be used. These can hold two or three horizontally mounted lamps of 65 Dlm. The bowl of opal glass projects 7 cm from the ceiling. This prevents strong contrasts, while the chance of damage during the carrying in of trunks is still very small. Where it seemed desirable opal glass "Philinea" tubular lamps of 25 and 40 W (30 and 50 cm long) were introduced, for instance above the toilet mirror, the writing desk, etc. These lamps are so mounted as to be fairly free of vibrations, and thus last as long on a ship as under ordinary conditions. There are also lamps in the closets which are switched on and off automatically by the opening and closing of the doors.

In ordinary cases, however, the level mentioned of 85 lux (8 ft. cdl.) may be retained as a standard. If this level is used in the cabins, the other rooms may be compared with them.

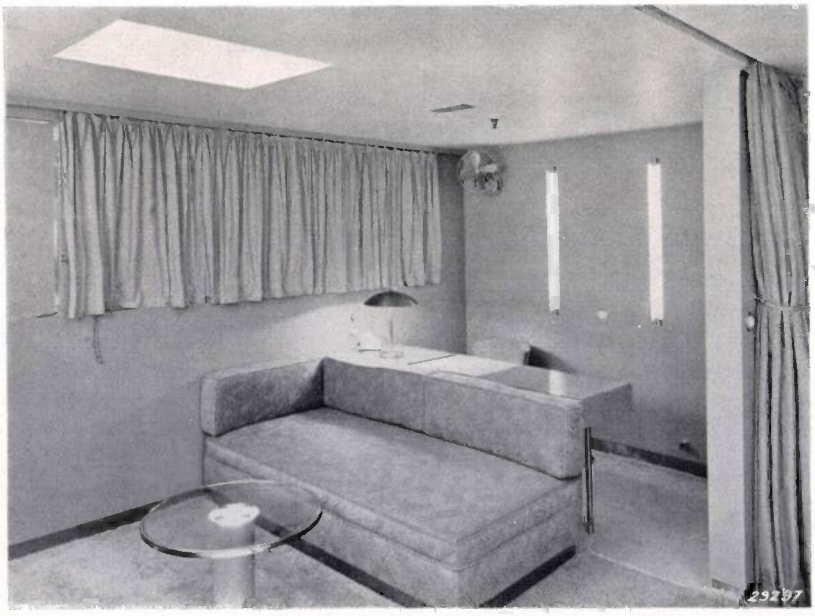


Fig. 2

A certain degree of economy may very well be observed in the case of the corridors (*fig. 3*), where light is burning day and night, and where the total length is so great. The level may not of course be so low that upon coming out of brightly lighted rooms such as cabins, one can no longer see clearly. Taking the level of 85 lux (8 ft. cdl.) for the cabins into account, a level of 40 lux (3.7 ft. cdl.) for the corridors may be said to be satisfactory.

The vestibules (*fig. 4*) and the dining rooms (*fig. 5*) must be festively and decoratively lighted. A high level of illumination of for

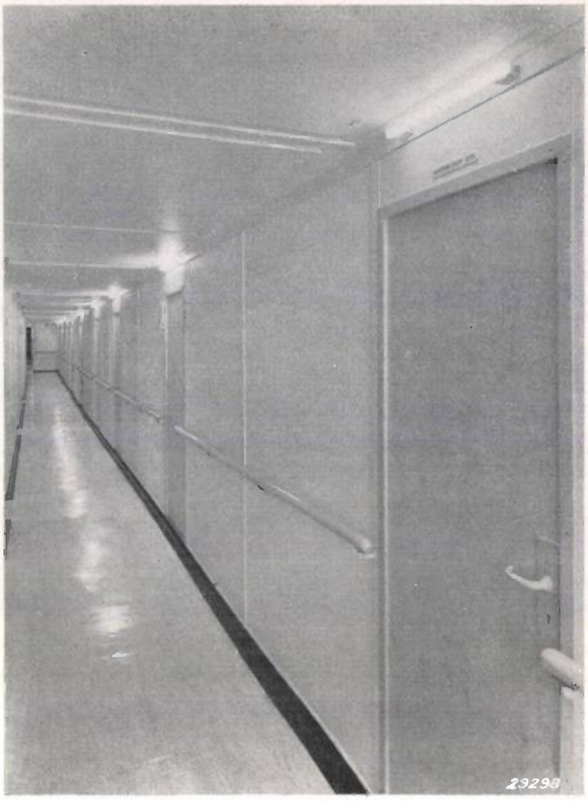


Fig. 3. A corridor on board the "Nieuw Amsterdam". There was here little choice possible in the type of illumination. The ceilings consist of large trap doors on hinges, behind which are the connection lines. The walls must be kept as smooth as possible because upon embarcation there are always many porters passing along the fairly narrow corridors with heavy trunks on their backs. For this reason "Philinea" lamps were used which occupy very little space; they are installed in the angle between ceiling and wall in a horizontal position above every door. In this way a fairly uniform illumination of 40 lux on an average is obtained.

Fig. 4. Large vestibule of the "Nieuw Amsterdam". This vestibule often serves as entrance hall for those embarking from the dock. In this case the height of two decks was available, which made it easier to reach a satisfactory uniformity in the illumination. The light-coloured ceiling and the almost white floor covering also contribute to this result. At the height of each deck there is a continuous line of "Philinea" tube lamps around the vestibule, while on the stair side there is an additional row of "Philinea" lamps behind matt glass plates. The intensity of illumination is 100 lux on an average. The relief in dull gilded metal on a background of green patiné bronze is lighted from the edges by a row of lamps in a light cove (75 lamps of 25 W).

Fig. 5. Dining room of the tourist class of the "Nieuw Amsterdam". The main illumination is semi indirect, and consists of a light-cove of Venetian glass, which runs around the higher middle portion of the ceiling. The glass is composed of glass rods containing air bubbles which are fused together and bent, thus forming elements 12 cm wide. These elements are mounted in a row side by side and are alternately very pale green and pink in colour. Since the points of light are faintly visible through the glass the lamps must be placed at regular intervals behind the glass. There are in all 172 lamps of 40 W installed in the recess. The surrounding lower-ceilinged portion of the room is lighted by 24 plafonniers and 16 wall appliqués of the same Italian glass each containing 3 lamps of 40 W. The intensity of illumination in the middle of the room is 110 to 150 lux, in the lower portion 45 lux. The illumination also serves to bring out the two large bas-reliefs at the ends of the room.

example 120 lux (11.1 ft. cdl.) is desired in the middle of the rooms at table height. A higher level of 120 to 200 lux (11.1 to 18.6 ft. cdl.) is suitable for the ball room as well as for writing and reading rooms and card rooms. A smoking room, on the other hand, were the main object is restfulness, requires no more than about 60 lux (5.6 ft. cdl).

When the main level of illumination has been fixed upon, attention must be given to the systems of illumination. A system of direct lighting is suitable in the cabins as well as in the corridors. In the larger halls a combination of direct and indirect lighting will be chosen in order to obtain uniformity. The occurrence of reflections then improves the appearance of table silver, ladies' toilets and jewels. In other cases (*fig.* 5) recessed lighting was chosen.

In carrying out the details of the plan thus established, a choice must be made of types of lamps and fixtures in order to obtain the desired results. In doing this the lowness of the ceilings presents a problem which is typical of ship illumination in general. It is therefore necessary to place the sources of light against the ceiling or even partially in the ceiling, so that the ceiling itself is partially unilluminated. Since in addition the rooms often have a relatively large floor space, the angles at which the light is reflected from the walls upon the usually horizontal surfaces to be illuminated are generally unsatisfactory, so that the contribution from the walls is only slight.

Under these circumstances it is not easy to achieve a uniform illumination without disturbing local contrasts. Where there are double ceilings, specially designed built-in plafonniers are used. The fitting constructed for this purpose can hold two or three lamps of 65 Dlm in a horizontal posi-

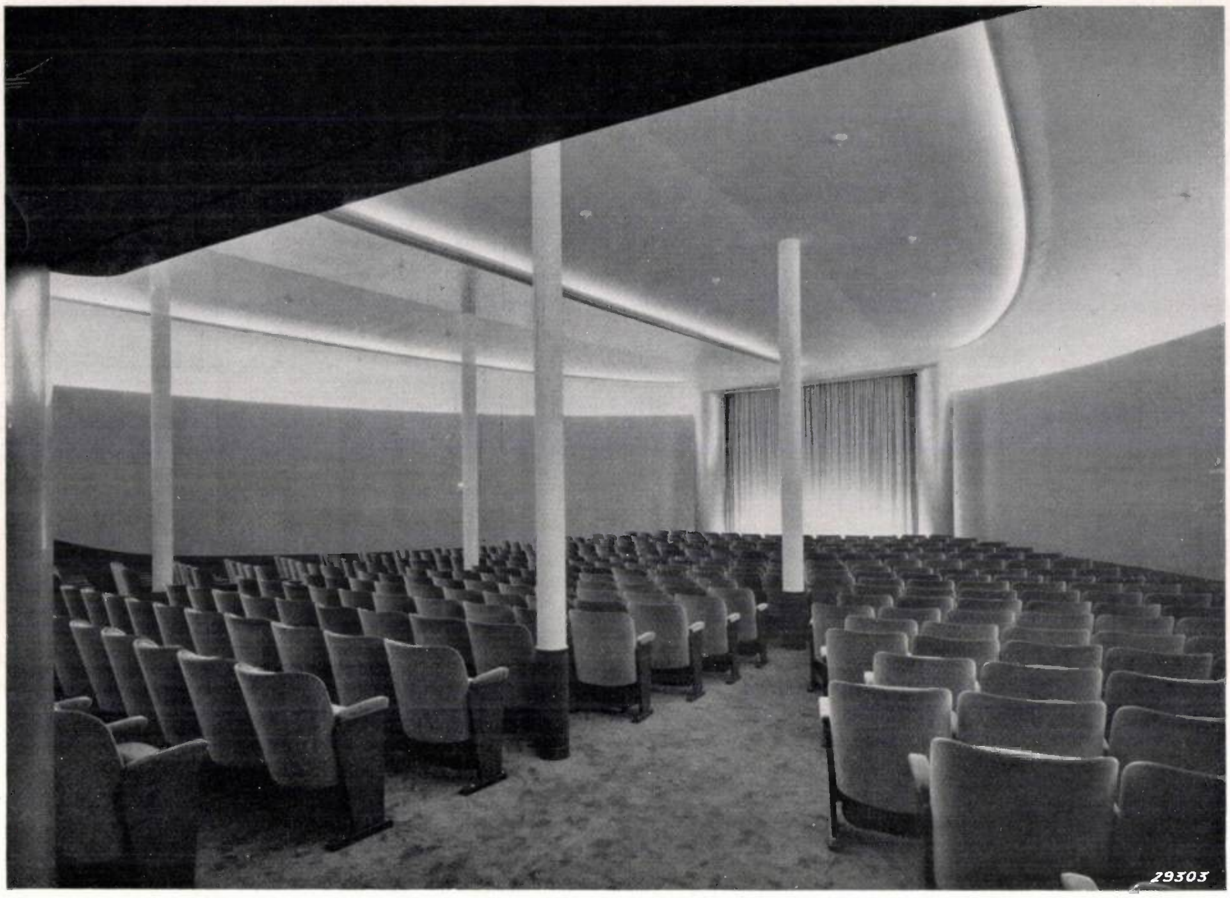


Fig. 6. Illumination of the theatre and cinema hall of the "Nieuw Amsterdam", which seats 350 persons. In order to lose nothing of the height the illumination had to be built into the walls and ceiling. A continuous line of tube lamps was installed along the three narrow bands where the slightly curved surfaces of the ceiling meet. In addition a light recess was made above the wine-red wainscoting, and equipped with a double row of tube lamps.

tion, and has a rosette which is sunk 3 cm into the ceiling. The opal glass bowl then projects only 7 cm from the ceiling. Because of this slight projection, damage to the bowls in carrying large steamer trunks into the cabins is avoided. With the same purpose in view the illumination of the corridors is carried out with tubular lamps placed as high as possible above the doors.

These opal glass "Philinea" tubular lamps of 25 and 40 W (30 and 50 cm long) are also installed in the cabins wherever it seemed necessary, in addition to the plafonniers of which there are one, two or three according to the size of the cabin. They are mounted in fittings with pistons having a double spring action, and are therefore fairly free of vibration. Besides in the corridors these tubes are also found above the toilet mirror and writing desk, and a corner of the cabin where one might wish to sit and read and beside the mirrors over the wash hand stands. Fastened to the wall over the beds there are small shaded reading lamps with small tubular lamps of 25 W.

In order to illustrate the various types of illumination we include a series of photographs taken on board the "Nieuw Amsterdam" of the Holland-America Line. The illumination was designed by the Philips consulting bureau for illumination and the technical staff of the steamship line, in collaboration with all the architects and decorators who

Fig. 7. The "Grand Hall" of the "Nieuw Amsterdam". This is the fashionable meeting place of the guests before and after dinner. The lighting is entirely indirect. The hall is decorated in light grey, as a suitable background for the colourful toilets of the ladies and the black and white of the gentlemen. The light is projected over uniform grey backgrounds from behind all the decorated surfaces of the walls and ceilings. Behind the pilasters and mural paintings there are continuous lines of tube-lamps; behind the lowered ceiling, executed in cast aluminium, there are rows of 40 W lamps with 4 lamps per metre. In front of the pilasters there are 4 large glass bowls and in the colums near the entrances to the hall there are 4 reflectors which provide for the further illumination of the ceiling.

Fig. 8. Third class smoking room of the "Nieuw Amsterdam". In this room a different solution has been found of the problem of introducing variation into the lighting of the low-ceilinged rooms of a ship. In the middle bent milk glass plates make up a large light panel behind which the lamps are hidden, while around the edges of this panel opal glass wall brackets and small fittings built into the ceiling provide additional light.

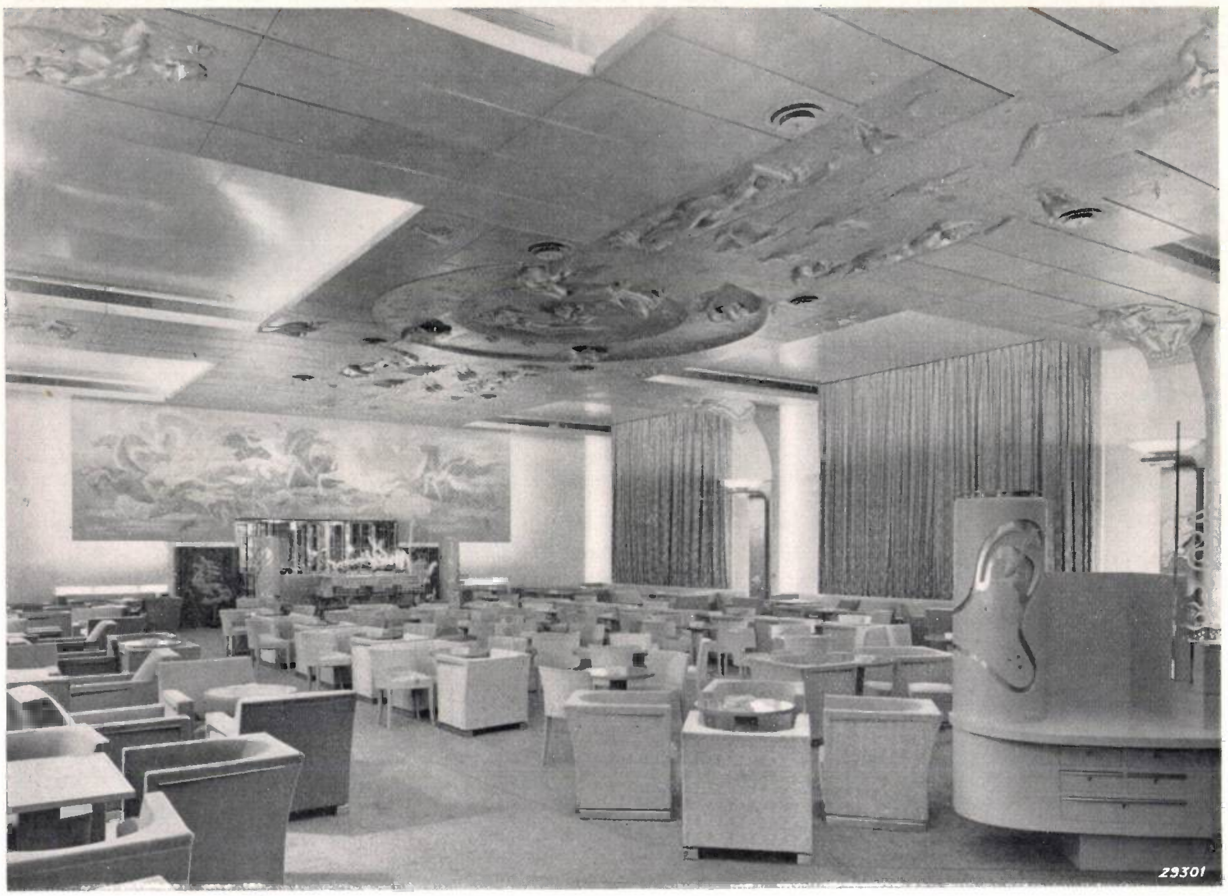


Fig. 7

Fig. 8

Fig. 9. The swimming pool of the "Nieuw Amsterdam" also has the low ceiling typical of ships. The plafonniers described in the text were used for the built-in illumination from above. By making the bowls of "Philiphan" glass (see Philips techn. Rev. 3, 47, 1938) all the colours are deepened and the skin takes on a warm colour against the beautiful green of the pool.

fitted out the interiors. The text underneath the photographs gives in every case a description of the details of the lighting installation.

We should like to say a few words about the efficiency of the illumination of ships. In general it is low. Even when light colours are used, the building into the ceiling of the fittings makes the contribution of light reflected from the ceiling small, and the walls also contribute only very little because the rooms are not only low-ceilinged but have a large floor space.

For example, in the dining room of the s.s. "Ruys" of the Koninklijke Paketvaartmaatschappij, where the illumination was installed under collaboration of our consulting bureau and the technical staff of the K.P.M., an average intensity of illumination of 90 lux was measured, while the floor surface is 190 sq.m. It was found necessary to install 265 lamps of 40 W for this purpose. The 265 lamps give a light flux of $265 \times 450 = 119\,250$ lumens, of which $90 \times 190 = 17\,100$ lumens are incident on the plane where the illumination was measured. The efficiency is therefore 14.3 per cent. The efficiency may also be measured by dividing the number of lumens which falls on the plane in question ($190 \times 90 = 17\,100$) by the total number of watts ($265 \times 40 = 10\,600$). The result is 1.6 lumens per watt, while the yield of the 40 watt lamps used amounts to 11.25 lumens per watt. Compared with the results obtained in other cases these values are indeed very low.

AUDITORIUM ACOUSTICS AND SOUND ABSORPTION

by R. VERMEULEN.

534.845

The absorption of sound by the walls plays an important part in the acoustics of an auditorium. In this article the manner is discussed in which the absorption of sound takes place by thick or porous walls and panels. Further the method is indicated of obtaining the data on the absorptive capacity of different materials which are necessary for calculating the reverberation time of an auditorium. An example is finally given of the calculation of the reverberation time of an auditorium.

In order to improve the acoustics of an auditorium it seems more obvious to amplify the sound artificially [1]) than deliberately to destroy part of the sometimes small amount of sound energy. Nevertheless it will be no surprise to readers of the earlier treatises on auditorium acoustics in connection with reverberation [2]) and intelligibility [3]) that increasing the absorptive capacity of the walls is in many cases an efficient method of improving either the intelligibility of the spoken word or the quality of music. The fact that this is done at the expense of the intensity of the sound is no objection, since this can be compensated for by amplifier installations. Because of the fact that the duration of a given sound becomes shorter, speech will become more intelligible and music clearer and more brilliant. This method must not be pushed too far, since a certain degree of reverberation, especially in the case of music, is indispensable, and composers take into account, although perhaps unconsciously, a certain amount of reverberation in composing their works.

In order not to work blindly as far as acoustics are concerned in designing a concert hall or theatre, and in order not to reach a satisfactory result only after several unsuccessful and usually costly experiments with different wall coverings, it is desirable to be able to calculate the reverberation. In the well known formula of Sabine [2]) for the reverberation time [4]):

$$T = 0.16 \frac{V}{A} \quad \cdots\cdots\cdots \quad (1)$$

The value of V, the volume of the room in m^3 can always be filled in easily. The calculation of the total absorption A in m^2, however, requires more information. This total absorption is composed of the sum of the absorptions of the different wall surfaces [5]), and is expressed in "square metres of open window", in which the latter is considered to be completely absorptive [6]). The absorption A_i of each part of the walls is calculated from the surface S_i in m^2, and the absorption coefficient α_i of the material of which it is made:

$$A = \Sigma A_i = \Sigma S_i \alpha_i \quad \cdots\cdots \quad (2)$$

Knowledge of the absorption coefficients α is the factor which makes it possible to calculate beforehand the reverberation time of a hall. Since the weakening of the sound waves takes place upon their reflection by the walls, we shall first deal generally with the mechanism of reflection of waves.

Reflection and absorption

In the air sound is propagated as a longitudinal vibration which is accompanied by a periodic change of air pressure. The momentary value of these changes we shall call the sound pressure p. For a plane wave it is proportional to the momentary value of the speed of vibration v. The quotient p/v of sound pressure and vibration speed is the so-called wave resistance W of the material [7]). The reflection and absorption of a sound wave in passing from air into another medium are determined solely by the difference in wave resistance of the two substances. The wave resistance can be represented as the product of the density ϱ and the speed of propagation c of sound:

$$W = \frac{p}{v} = \varrho c. \quad \cdots\cdots \quad (3)$$

When a plane sound wave falls perpendicularly upon the plane bounding two substances with different sound resistance, the wave cannot proceed as a whole; this follows from the transition conditions

1) J. de Boer, Philips techn. Rev. 3, 225, 1938.

2) A. Th. van Urk, Philips techn. Rev. 3. 65, 1938.

3) R. Vermeulen, Philips techn. Rev. 3, 143, 1938.

4) By reverberation time is meant the time in which the average sound level in the hall decreases 60 bB.

5) The absorption of the sound waves in the air itself, which may be appreciable at high frequencies, is here left out of consideration.

6) Absorption is also expressed in "Sabines" which units of absorption correspond with one square foot of open window = 0.093 m^2 of open window.

7) R. Vermeulen, Philips techn. Rev. 2, 47, 1937.

at that boundary plane. These conditions require that on both sides of the plane both the momentary value of the pressure p and that of the speed of vibration v shall be the same. If there is only one wave on both sides, then the wave resistance W, which is equal to the quotient p/v, would also have to be the same for these two waves, while the hypothesis was that the two substances have different wave resistances. The above-mentioned transition conditions for p and v can, however, be satisfied if, in addition to a direct wave proceeding through the second substance, there is also a reflected wave in the first substance which is superposed upon the incident wave.

If we indicate the quantities which relate to the incident, reflected and transmitted waves with the subscripts 0, 1 and 2 respectively, the following equations are valid:

$$p_0 = W_1 v_0;\ p_1 = -W_1 v_1;\ p_2 = W_2 v_2 \text{ and} \quad (4)$$

$$p_0 + p_1 = p_2;\ v_0 + v_1 = v_2 \quad \cdots\cdots\cdots \quad (5)$$

The pressure and the speed of the reflected and transmitted wave may easily be calculated from equations (4) and (5), as has been done in the text under *fig. 1*. For the reflection coefficient R, that is the ratio between the sound pressure of the reflected and of the incident wave one then finds:

$$R = \frac{p_1}{p_0} = \frac{-v_1}{v_0} = \frac{W_2 - W_1}{W_2 + W_1} \cdot \cdot \cdot \quad (6)$$

This reflection coefficient is therefore the quotient of two amplitudes; for the absorption coefficient a, however, it is not customary to take the quotient of two amplitudes, but to define it as the ratio between the decrease in intensity of the reflected wave and the intensity of the incident wave (I_0), so that:

$$a = \frac{I_0 - I_1}{I_0} = \frac{I_2}{I_0}, \text{ or: } a = 1 - R^2, \cdot \cdot \quad (7)$$

where I_1 is the intensity of the reflected wave and I_2 that of the refracted wave. Therefore the following expression for the absorption coefficient is valid:

$$a = \frac{4\ W_1 W_2}{(W_1 + W_2)^2} \cdot \cdot \cdot \cdot \cdot \cdot \quad (8)$$

In order to be able to use formulae (6) and (8) it is necessary to know the wave resistance of air and other materials. Since the density of air is 1.2 mg/cc at room temperature and the speed of propagation of sound is 340 m/sec, the wave resistance becomes

$$W_1 = 41 \text{ gr cm}^{-2} \text{ sec}^{-1} \cdot \cdot \cdot \cdot \cdot \quad (9)$$

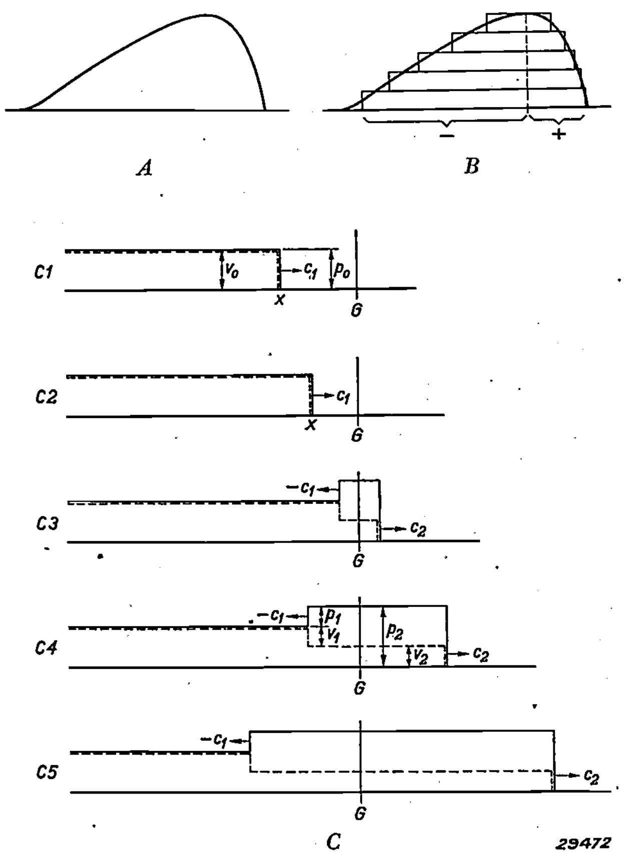


Fig. 1. Reflection of a wave impulse.

It is sufficient to discuss the reflection of a "unit impulse" (as it was called by Heaviside), because it is always possible to represent any wave (A) as made up of a large number of positive (+) and negative (—) elementary impulses (B). In such a unit impulse ($C1$) to the left of x there is a constant pressure p_0 (continuous line ———) and a velocity v_0 toward the right (broken line — — —). To the right of point x, where the wave front is situated at the moment, there is a state of rest, and the pressure is the normal atmospheric pressure. The wave front is displaced with the velocity of sound c to the right ($C2$). In the wave the ratio between the pressure p_0 and the velocity v_0, as in every plane wave, is equal to the wave resistance: $W_1 = \varrho_1 c_1 = \frac{p_0}{v_0}$, where ϱ_1 is the density. Upon reaching the boundary G ($C3$) between the two media, a wave begins in the second medium ($C4$) in which the ratio between the increase of pressure p_2 and the velocity v_2 must now be equal to the wave resistance in the second medium: $W_2 = \varrho_2 c_2 = \frac{p_2}{v_1}$. Since, however, the pressure and velocity must be continuous at the boundary plane, a wave in the reverse direction (p_1, v_1) must occur, in order that the following equations may be satisfied:

$$p_0 + p_1 = p_2, \text{ and } v_0 + v_1 = v_2,$$

with:

$$p_0 = W_1 v_0 \text{ and } p_2 = W_2 v_2$$

while in connection with the negative direction of propagation of the reflected wave: $p_1 = -W_1 v_1$. It follows from this that:

$$\frac{p_1}{p_0} = \frac{-v_1}{v_0} = \frac{W_2 - W_1}{W_2 + W_1},$$

$$\frac{p_2}{p_0} = \frac{2\ W_2}{W_2 + W_1} \text{ and}$$

$$\frac{v_2}{v_0} = \frac{2\ W_1}{W_2 + W_1}.$$

In the transition of sound from air into a wall of a given material, the phenomena which take place depend very much upon the way in which the sound is propagated in this material. As the simplest case let us suppose that the second substance is homogeneous and isotropic, and moreover that it extends infinitely on the other side of the boundary surface. This last assumption means neglecting the influence of reflection at other boundary surfaces. Since solid substances on the average certainly have a density 1000 times that of air, while in general the speed of sound through them is also greater than in air, the wave resistance W_2 will also be several thousand times that of air. It follows from formula (8) that such a wall can absorb practically no sound, and that therefore there will be practically total reflection of the sound, as may be seen clearly from formula (6). Therefore homogeneous solid substances are entirely unsuitable as absorptive material for sound.

In practice we are generally not concerned with this simple case. The wall is far from infinitely thick, and moreover sound is absorbed in the air-filled pores of many technical materials. Before going deeper into this subject we shall first extend our consideration of reflection and absorption to obliquely incident waves.

Obliquely incident waves

Although the treatment remains fundamentally the same, there is an essential difference between the reflection of normally incident waves and of obliquely incident waves, because on the one hand it is clear that in the transition conditions mutual equality of only the normal components of the vibration speed is required while, on the other hand, in the propagation of sound in the second material the wave resistance still continues to determine the ratio between the pressure and the *total* vibration speed. The formulae (4) therefore remain unaltered, but instead of formulae (5) we now obtain the following:

$$p_0 + p_1 = p_2\,;\; v_0 \cos \Theta_0 + v_1 \cos \Theta_1 = v_2 \cos \Theta_2\,, \quad (10)$$

where $v \cos \Theta$ is the normal component of the speed, and the angles Θ are given in *fig. 2.* Formulae (4) for the wave resistance may be adapted in the following way for oblique incidence:

$$p_0 = W_1 v_0 = \frac{W_1}{\cos \Theta_0} \cdot v_0 \cos \Theta_0 = \frac{W_1}{\cos \Theta_0}\, v_{n0}\,, \quad \text{etc.} \quad (11)$$

so that in the formulae (4) and (5) W becomes $W/\cos \Theta$ v, v_n:

$$p_0 = \frac{W_1}{\cos \Theta_0}\, v_{n0}\,;\quad p_1 = \frac{-W_1}{\cos \Theta_1}\, v_{n1}\,;$$

$$p_2 = \frac{W_2}{\cos \Theta_2}\, v_{n2} \quad \text{and} \quad (4^*)$$

$$p_0 + p_1 = p_2\,;\;\; v_{n0} + v_{n1} = v_{n2} \quad (5^*)$$

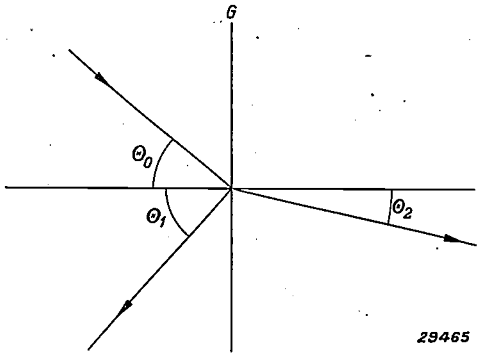


Fig. 2. Reflection and refraction at a boundary surface. Θ_0 angle of incidence, Θ_1 angle of reflection and Θ_2 angle of refraction. The relation between the angles Θ is given by Snellius' law:

$$\frac{\sin \Theta_1}{\sin \Theta_2} = \frac{c_1}{c_2} \text{ and } \Theta_0 = \Theta_1\,.$$

It may therefore be seen that the absorption depends upon the angle of incidence. Several other remarkable conclusions may be drawn from these formulae. Thus for materials for which $W_2 > W_1$ the absorption will be complete at a certain angle Θ_1, and there will therefore be no reflection, since with increasing Θ_1, cos Θ_1 decreases more rapidly than cos Θ_2, so that finally:

$$\frac{W_1}{\cos \Theta_1} = \frac{W_2}{\cos \Theta_2} \quad (12)$$

For larger values of Θ_1 the reflection coefficient becomes negative, which means that due to the reflected wave the change in pressure at the wall becomes smaller than in the case of the undisturbed incident wave.

If, moreover, the speed of propagation c_2 in the second medium is greater than that (c_1) in the first, total reflection will occur with an angle of incidence greater than $\sin^{-1} c_1/c_2$, just as with light. The difference from the case of light lies in the fact that this total reflection with *light* only occurs on the transition of the wave from glass, for instance into air, while with *sound* it occurs upon transition from air into a solid substance. This is, however, not of great importance, since the reflection coefficient at solid substances, as already mentioned, usually lies close to 1.

If a sound wave moves parallel to an absorbing surface, it constitutes a limiting case for $\Theta_1 \rightarrow \pi/2$, i.e. cos $\Theta \to 0$. According to (6) the reflection coefficient is then

$$R \to -1 \, ; \quad \cdots \cdots \quad (13)$$

this means that the reflected and the incident wave will just compensate each other, since at all points:

$$p_1 = -p_0 \text{ and } v_1 = -v_0 \quad \cdots \quad (14)$$

Experiments have actually shown that the primary wave disappears upon shearing incidence on a wall. In *fig. 3* the results of such experiments by von Békésy [8]) are reproduced.

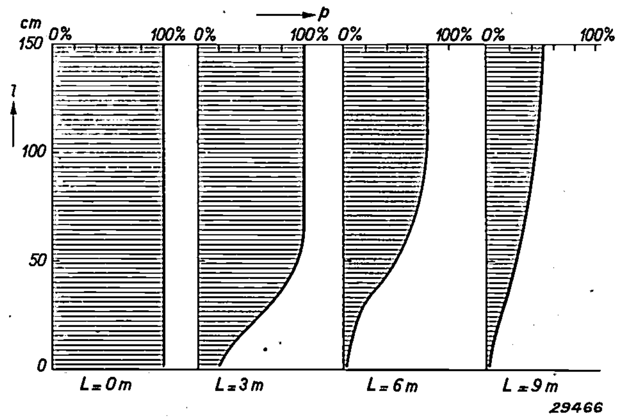


Fig. 3. A plane sound wave which proceeds parallel to a wall of absorbing material, is reduced in intensity close to the wall. The longer the distance along the wall covered by the wave, the greater the distance from the wall at which this weakening may be observed. The sound pressure p in per cent of the original value is plotted as a function of the distance l to the wall for several values of the distance L covered by the sound wave along the wall.

Porous materials

Most walls do not in the main absorb sound by being themselves set into motion under the influence of the sound vibration, but because of the fact that there are many pores in the material in which the sound wave is propagated. Because of its viscosity the motion of the air along the walls of the tiny canals is retarded, and this leads to the damping of the sound waves. In order to study the effect of this damping we begin with the equations of motion which form the basis of wave motion.

For a plane wave proceeding in the direction x through a homogeneous material without internal friction, the following differential equation is valid:

$$-\frac{\delta p}{\delta x} = + \varrho \frac{\delta v_x}{\delta t} \, , \quad \cdots \quad (15)$$

which expresses the fact that the force which acts upon unit volume is equal to the mass times the acceleration. If friction exists, a difference in pressure between front and rear sides of the element of volume will also be necessary to overcome this resistance. Let this increase be proportional to the velocity v_x. Then the differential equation becomes

$$-\frac{\delta p}{\delta x} = + \varrho \frac{\delta v_x}{\delta t} + r \, v_x \, , \quad \cdots \quad (16)$$

in which the factor r may be called the specific resistance. With sinusoidal vibrations with a frequency $\omega/2\pi$ the above may be written in the following form:

$$-\frac{\delta p}{\delta x} = (j\omega\varrho + r) \, v_x = j\omega\varrho \left(1 - j \frac{r}{\omega\varrho}\right) v_x =$$

$$= \varrho \, (1 - j/\nu) \, \frac{\delta v_x}{\delta t} \, , \quad \cdots \quad (17)$$

in which thus a "reduced frequency"

$$\nu = \omega \, \varrho / r \quad \cdots \cdots \quad (18)$$

has been introduced. By comparing formulae (15) and (17) with each other it may be seen that it is only necessary to substitute $\varrho \, (1 - j/\nu)$ for ϱ in the previous formulae. Since the velocity of sound is inversely proportional to the square root of the density, it is changed to $c = c_0/\sqrt{1 - j/\nu}$. In this complex speed of propagation the fact is expressed that the waves during propagation also diminish in amplitude. Therefore for the behaviour of the sound wave in the canals of the second medium there is a separate wave resistance:

$$W_2' = \varrho_0 \, (1 - j/\nu) \cdot \frac{c_0}{\sqrt{1 - j/\nu}} = \varrho_0 \, c_0 \sqrt{1 - j/\nu} . \quad (19)$$

Since, however, only a part σ of a cross section parallel to the boundary plane is occupied by the air canals, and the part $1-\sigma$ allows of no air displacement, the wave resistance W_2 of the canals toward the outside appears to be $1/\sigma$ times as great

$$W_2 = \frac{W_2'}{\sigma} = \frac{\varrho_0 c_0}{\sigma} \sqrt{1 - j/\nu} . \quad \cdots \quad (20)$$

By means of this value W_2 of the wave resistance the absorption can be calculated of a layer of porous material placed in front of a heavy wall. The result will depend upon the thickness d of the layer, the specific resistance r and the frequency $\omega/2\pi$. It will then be found that it is possible [9]) to express the absorption coefficient α in two dimensionless combinations of these quantities, namely the "reduced frequency" ν already introduced and the "reduced resistance":

[8]) G. von Békésy, Z. techn. Phys. **14**, 6, 1938.

[9]) L. Cremer, Elektr. Nachr. Techn. **10**, 242, 1933; **12**, 338 and 366, 1935.

$$\delta = \frac{dr}{\varrho_0 c_0} \quad \cdots\cdots\cdots \quad (21)$$

The value dr is the direct current resistance of a layer with the thickness d, which could therefore be measured, if necessary, by determining the current of air which flows between front and rear sides under the influence of a small difference in pressure.

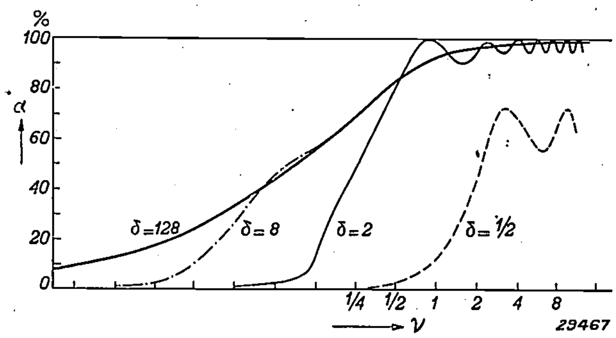


Fig. 4. The absorption coefficient α of the porous material placed in front of a hard wall is given above as a function of the "reduced frequency" $\nu = \omega\varrho/r$ with the reduced quantity $\delta = \frac{dr}{\varrho c}$ as parameter. When the reduced direct current resistance of the wall is large ($\delta = 128$), (*i.e.* when it is very thick or has a high specific resistance, and thus small pores) the absorption at low frequencies (*i.e.* at wave lengths longer than the thickness of the wall, thus $\nu < {}^1/_{20}$) is small, and increases regularly with increasing frequency, until at higher frequencies ($\nu > 4$) the absorption is practically complete. At smaller values ($128 > \delta > 1$) of the reduced resistance, the behaviour is practically the same for frequencies which have a wave length smaller than 6 times the thickness of the wall, but the absorption decreases rapidly toward lower frequencies. For very small values of the reduced resistance ($\delta < 1$) complete absorption is never attained. The reflection at the hard wall becomes increasingly less disturbed; the absorption does, however, change for different frequencies due to the occurrence of stationary vibrations in the material.

In *fig. 4* the absorption coefficient α is plotted as a function of ν for different values of δ as a parameter. It may be seen from this figure that the absorption decreases sharply for low frequencies, which is quite understandable when it is kept in mind that nodes of the air vibrations are formed at the walls, so that the velocity of the air can only reach a significant value at a distance from the wall of the order of $\lambda/2\pi$. If the wall is much thinner than $\lambda/2\pi$, it may not be expected that much energy will be dissipated by friction in the air motion in the pores.

When it is understood that one of the conditions to be observed in order to obtain a high absorption coefficient is the avoidance of reflection at the front surface, a material will be sought which is composed of several layers in such a way that the wave resistance of the first layer differs only slightly from that of air, and that of the following layers becomes successively greater in order to provide adequate adsorption of the wave. With a medium thickness of the wall covering, however, this method does not provide such a great improvement that it would make up for the complications.

Moving panels

Not only the presence of pores in a wall, as discussed above, can serve to make a wall more suitable for the absorption of sound. It is known that halls with a wall covering of wood often possess good acoustic properties. The thickness of such wooden panels is always so small compared with the wave length of audible sound that the mutual differences of phase of the movements in the wall are extremely slight and there can scarcely be any question of wave propagation *in* the panel. The behaviour of such a thin wall is now characterized, not by a wave resistance, but by its impedance to motion Z_2, which gives the ratio between the force F acting on the wall and the velocity v which it obtains as a consequence:

$$Z_2 = \frac{F}{v} = \frac{p}{v} S, \quad \cdots\cdots \quad (22)$$

where S represents the area of the thin wall considered. From equation (22) it may be seen that the impedance to motion per unit area, Z_2/S now takes the place of the wave resistance $W = p/v$ in the case of the infinitely thick wall; so that for the reflection coefficient with normal incidence the following equation takes the place of equation (6):

$$R = \frac{p_1}{p_0} = \frac{-v_1}{v_0} = \frac{Z_2/S - W_1}{Z_2/S + W_1} \quad \cdots \quad (23)$$

If the panel is fastened rigidly against a heavy wall, the impedance of the wall will thereby be made large. When it is fastened on laths with an air space between panel and wall it is much more mobile. The cushion of air will act as an elastic spring and will increase the resonance frequency of the wall to such an extent that in calculating its value a satisfactory approximation is obtained by considering the panel itself as absolutely flexible and considering only the stiffness of the air cushion and the mass of the panel.

Beginning with the law for adiabatic change of state of a gas, $pV^\varkappa$ = constant, one obtains upon differentiation:

$$\frac{\Delta p}{p} + \varkappa \frac{\Delta V}{V} = 0 \quad \cdots\cdots \quad (24)$$

The mechanical stiffness of a cushion of air with a surface S and a thickness d is defined as the ratio of the force ΔK on the surface to the decrease in thickness caused by that force $-\Delta d$.

$$-\frac{\Delta K}{\Delta d} = -\frac{S\Delta p}{\Delta V/S} = -\frac{\Delta p}{\Delta V}\cdot S^2 \cdot \quad (25)$$

The stiffness per unit of surface is therefore

$$-\frac{\Delta p}{\Delta V}\cdot S = +\frac{\varkappa p}{V} S = +\frac{\varkappa p}{d} = \frac{\varrho c^2}{d}, \quad (26)$$

since the velocity of sound is given by $c = \sqrt{\varkappa p/\varrho}$. For a mass m per unit of surface of the panel the following value is found for the resonance frequency f:

$$2\pi f = \omega = \sqrt{\frac{\varrho c^2}{dm}} \quad \cdots\cdots \quad (27)$$

In the neighbourhood of this frequency the impedance Z_2 of the wall can be much smaller than would be the case without the air cushion. It thereby becomes possible for the values of Z_2/S and W_1 (the wave resistance of air) to approach each other more closely in formula (23), and in this way the condition for a smaller coefficient of reflection is more nearly satisfied. Energy is then taken up by the panel from the incident waves, and it will depend upon the damping what part of this energy is destroyed and what part will be radiated again later as sound energy. It appears that under certain circumstances this latter radiation may lead to disturbing increase of the reverberation time with a certain frequency, but the absorption will usually be high due to a sufficiently rapid damping, and the reverberation will be shortened exactly at the resonance frequency. Experiments by Erwin Meyer[10]) have shown that the above considerations are somewhat too simple.

It is indeed true that the thickness of the cushion of air is usually so small compared with the wave length, that it may actually be considered as a simple spring within a wide frequency range, but that the other dimensions of the cushion of air are often such that stationary vibrations are very well possible, which not only make the theory incalculable, but which also decrease the absorption very much. If these vibrations are prevented (by dividing the air space into sections with partitions), or damped (by filling) a better agreement with the measurements is found (*fig. 5*). For damping it is not necessary to fill the air space completely with a porous material (fig. 5*b*); it is already enough for example if a 5 cm thick wadding is introduced around the edges (fig. 5*c*).

By making a certain number of perforations in the panel and filling the air space with porous material, a gradual transition can be obtained from the properties of a moving panel to those of a porous material. The holes for this purpose need not be large: if 4 per cent of the surface consists of holes, the effect is appreciable, with 15 per cent the action is almost the same as that of a porous material alone.

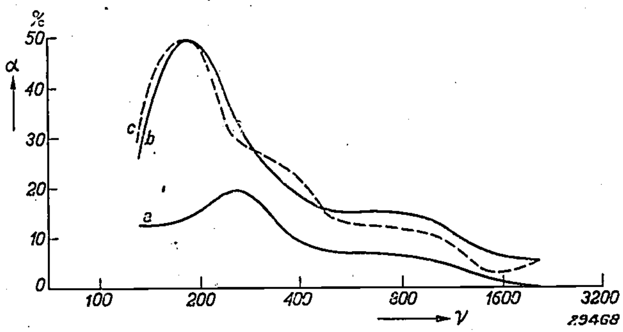


Fig. 5. Absorption coefficient α per m² expressed in per cent as a function of the frequency in c/s for 5 mm thick triplex panels of 1.5 m by 1 m, placed 5 cm in front of the wall.
a) undamped cushion of air.
b) air space filled with a thickness of 4 cm of wadding.
c) only the edges of the air space sealed with 5 cm of wadding.

The great importance of these resonating panels is, that a high absorption coefficient can be obtained in the low-frequency range where the porous substances were found not to give satisfactory results. At the same time there is the possibility of fixing the maximum absorption at a desired frequency by the choice of the thickness of the cushion of air. It will be found stated[11]) that in halls with a calculated short reverberation the acoustics are improved by wooden resonating panels. This experience should probably be interpreted to mean that the curve of the reverberation as a function of the frequency is made flatter in this way.

Measurement of absorption

While it is true that the above considerations give some idea of the way in which sound absorbing materials act, it is, however, very desirable to measure the absorption directly on a sample of the material. It would for example be possible to determine the reflection coefficient at different angles of incidence, and to calculate from the results the average value of the absorption coefficients over the different directions according to the formula:

$$a_r = 2\int_0^{\pi/2} a(\Theta)\ \sin\Theta\ \cos\Theta\ d\,\Theta, \quad (28)$$

in which a value is ascribed to the different absorption coefficients which is proportional to the solid angle for which they are valid. It is, however, still

[10]) E. Meyer, Elektr. Nachr. Techn. **13**, 98, 1936.

[11]) H. Baginall and A. Wood, Planning for good acoustics, p. 137, Methuen & Co. Ltd.

an open question whether this formula gives the value of the absorption coefficient which must be used in Sabine's formula for calculating the reverberation time. It seems therefore preferable to determine the absorption coefficient by measuring with relatively small samples of material the difference in the reverberation time of a test chamber in which the material is used and that in the same chamber when the material is not used.

If the volume of the test chamber is V, the area of the test material S_2 and the reverberation time without and with the material T_1 and T_2 respectively, formulae (1) and (2) give for the absorption coefficient:

$$\alpha = 0.16 \frac{V}{S_2}\left(\frac{1}{T_2} - \frac{1}{T_1}\right) \quad . \quad . \quad . \quad (29)$$

In order to use as little as possible of the test material and yet to have as accurate as possible measurements of the absorption coefficient, the characteristic absorption of the chamber itself upon which that of the test material is superposed should be kept as small as possible. The test chambers for these absorption measurements must therefore be so-called "hard" chambers (Hallräume, reverberation chamber). In such chambers, however, the deviations from Sabine's law make themselves felt, and the measured absorption coefficient is found not to be independent of

1) the position of the sample in the room,
2) the size of the sample,
3) the absorption of the rest of the wall space,
4) the size of the reverberation chamber.

It may be stated in general that too small samples give too high an absorption coefficient, and that the other influences are especially noticeable with materials having large absorption coefficients. The fact that this makes it difficult to compare the results obtained in different laboratories, may be seen from the comparison of measurements made in different countries [12]). Values of the absorption coefficient were measured with the same material at 265 c/s, for example, and the results varied from 0.33 to 0.58, while at 512 c/s they varied from 0.69 to 0.99. In *fig. 6* the influence of the size of the sample on the absorption coefficient is shown. Although it may be seen from the figure that with sufficiently large dimensions of the sample a practically constant absorption coefficient is obtained, even this final value is not always the same for two different laboratories.

An extraordinary result may be mentioned which was obtained in an unusually large reverberation test chamber with 1 m² of slag-wool at

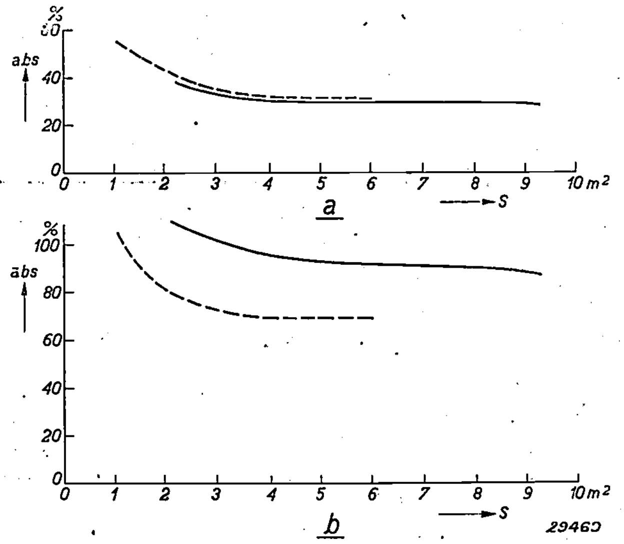


Fig. 6. Absorption coefficient as a function of the surface area of the absorbent material at 256 cycles/sec (a) and at 512 c/s (b), measured in two different American laboratories. Curves of same laboratory continuous and broken lines respectively.

2900 cycles per sec. A "record" value of 300 per cent was obtained for the absorption coefficient calculated according to formula (7). These deviations must not be taken too tragically since they occur chiefly under extreme conditions, and under more ordinary conditions and with reasonable application satisfactory results may be obtained. In order to give an idea of values found, several values of the absorption by different material as found in the literature are given in *table I.*

Table I

Absorption with 512 c/s.

Open window	100%
Opening of stage	25 ÷ 40%
Curtains	15 ÷ 50%
Rug	20 ÷ 40%
Slag-wool	60—80%
Felt	50—80%
Acoustic plaster	38—70%
Brick wall	3%
Wood	6—10%
Area occupied by public	96%
Per person	0,5 m² open window
Per unoccupied chair (upholstered)	0,3 m² „ „
Per unoccupied chair (not upholstered)	0,03 m² „ „

The most important contribution to the total absorption of a room is not only difficult to measure, but is also variable, since it is provided by the

[12]) R. M. Morris, G. M. Nixon and J. S. Parkinson, J. acoust. Soc. Amer. 9, 237, 1938 and E. Meyer, Akust. ZS. 2, 180, 1938.

Table II

Calculation of the reverberation time of Queen's Hall, London.

Volume of the hall: 422 000 cubic feet.
Number of seats : 2 026
Volume per seat : 208 cubic feet.

Material	Surface in sq.ft.	Absorption coefficient	Absorption in sq.ft. open window
Hard plaster	19 400	0.025	485
Glass	1 000	0.027	34
Vents	180	0.5	90
Wooden panels. . .	5 080	0.1	508
Dado linings. . . .	1 500	0.1	150
Painted cloth on wood	1 200	0.12	144
Wooden floor and stage	11 250	0.06 (0.10)	608
Carpet	4 790	0.2 (0.10)	863
Linoleum	3 280	0.03	89
Organ-case	1 000	0.08	80
Upholstered chairs	2 026 in number	1.3 sq.ft. per chair	2 633

Type of occupation	Number of persons	Absorption in sq.ft. per person	Absorption in sq.ft. open window
Orchestra . .	90	4,7	423
Choir	250	4,7	1 175
Audience (full)	2 026	4,7—1,3	6 888
Audience (1/3)	675	4,7—1,3	2 295

Type of occupation	Tota. absorbing surface in sq.ft.	Reverberation time in sec.
Empty hall	5 660	3.6
Choir + orchestra	7 166	2.9
1/3 audience + choir + orchestra	9 533	2.2
Audience + choir + orchestra	14 160	1.4

audience. The previously mentioned method [13]) makes it possible to determine the absorption coefficient during a concert without interruption (even at different frequencies), by using filters. Here also it is found difficult to speak of a single definite absorption coefficient or of an absorption in sq.ft. of open window per person: a completely filled hall has a considerably lower absorption per person than a half-filled one (*fig.* 7). On the other hand it is fortunate that this is so, because in this way the reverberation time will vary less with changing degree of occupation. In order to avoid large differences it is a good thing to use heavily upholstered chairs which make up for the absorption of the missing audience.

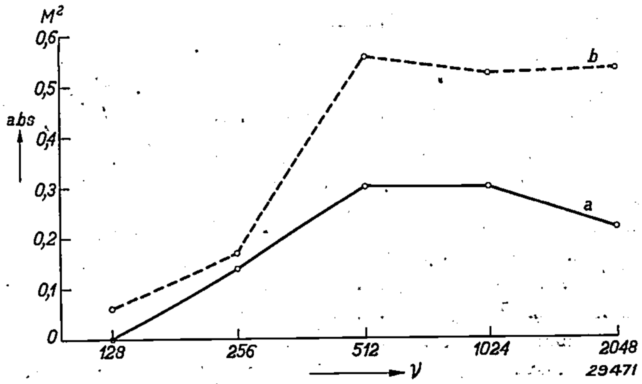


Fig. 7. The variation of the absorption per person in a hall as a function of the frequency ν.
a) hall completely occupied
b) hall half occupied.

Calculation of the reverberation time

In *table II* [14]) an example is given of the calculation of the reverberation time of an auditorium,

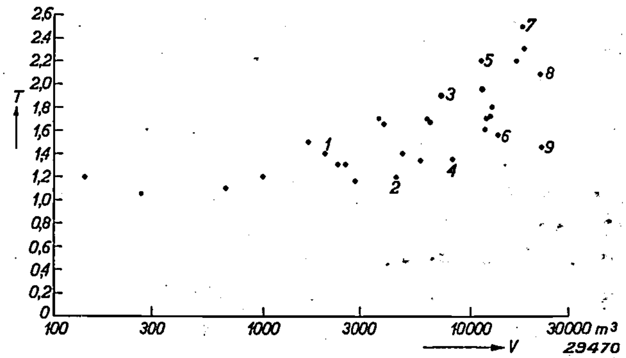


Fig. 8. Reverberation time T as a function of the volume V of several auditoria known to have good acoustics, *inter alia* the following:
1) Altes Gewandhaus, Leipzig.
2) Examination Hall, Cambridge. (Eng.).
3) St. Margaret's, Westminster.
4) Musikvereinsaal, Vienna.
5) Wagner Theater, Bayreuth.
6) Grand Theatre, Moscow.
7) Thomas Kirche, Leipzig.
8) Eastman Theatre, Rochester (U.S.A.).
9) Salle Pleyel (new), Paris.
This figure shows clearly that the longest reverberation time occurs in churches and the shortest in theaters, while concert halls usually lie in the middle.

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[13]) E. Meyer and V. Jordan, Elektr. Nachr. Techn. 12, 213, 1935; *cf.* also: Philips techn. Rev. 2, 266, 1937.

[14]) H. Baginall and A. Wood, Planning for good acoustics, p. 109 - 112.

Queen's Hall in London. The surface area of the different materials is calculated, and the absorption coefficient is estimated. In the case of the chairs the absorption per chair is used. In several cases 10 per cent is deducted because the surfaces in question are not entirely freely exposed to the sound. In the second half the absorption is calculated separately for the various groups of people present in the hall. The absorption of the occupied chairs is of course deducted.

It is now possible to calculate the total absorption and thus the reverberation time, according to Sabine for different combinations as done in the third part of the table. It may be seen that the audience has a very great influence. In calculating the absorption of a hall it is certainly desirable to find out whether or not the reverberation time becomes too long when the hall is only partially occupied. There must be sufficient permanent absorption in the hall in the form of upholstered seats or on the walls. Then in order not to have too short a reverberation time when the hall is fully occupied, the volume of the room must not be too small.

Fig. 8 gives an idea of the reverberation times usual in practical cases. The reverberation time T is represented as a function of the volume V of the hall for various halls known to possess good acoustics.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS' GLOEILAMPENFABRIEKEN

1319: W. de Groot and F. M. Penning: Die Ergiebigkeit der Neutronenproduktion durch D-Kanalstrahlen (Physica **5**, 512 - 520, June 1938).

From the yield of neutrons produced by deuterium ions upon striking plates of deuterium phosphate, lithium, beryllium and carbon as measured by Amaldi, Hafstad and Tuve, the effective diameters of these atomic nuclei were determined as functions of the speed of the projectile, use being made of the values found by Mano of the range in air of charged particles and of the atomic stopping power of these elements. With the help of the approximation formulae following from these values, it is possible to extrapolate the yields as found by Amaldi, Hafstad and Tuve for lower energies also. The results of Ladenburg and others, Scherrer, Kallmann and other authors can be made to agree with these formulae by a suitable choice of parameters. It is found necessary in some cases to determine once more the recalculation factor for the transition from a compound to pure deuterium, making use of the proportionality of the atomic stopping power to the square root of the nuclear charge number.

1320: J. H. de Boer: Interpretation of molecular phenomena by means of potential curves. IV. Fluorescence and phosphorescence; Activation energy (Ned. T. Natuurk. **5**, 57 - 64, Mar. 1938).

The significance of the Franck and Condon principle is discussed in relation to the position of absorption and emission bands of fluorescent substances. The chemical activation energy can be deduced from the potential curves. Activated adsorption and explosive substances are finally dealt with briefly.

1321: J. H. de Boer: Interpretation of molecular phenomena by means of potential curves. V. Colloids (Ned. T. Natuurk. **5**, 65 - 69, Apr. 1938).

Colloidal phenomena are discussed from the standpoint of potential curves according to the considerations of Hamaker (*cf* **1286**).

1322: C. J. Bakker, W. de Groot and F. M. Penning: Neutron-producing nuclear reactions and their yields (Ned. T. Natuurk. **5**, 102 - 115, May 1938).

A critical survey is given of the present insight into the production of neutrons by various nuclear reactions. The results obtained by others are compared with those of several measurements by the authors, and a satisfactory agreement is found. Special mention must be made of the fact that Penning was able to produce neutrons in a closed tube filled with deuterium at a low pressure with a discharge in a magnetic field in which the electrons must move in spiral paths between the electrodes. Deuterium molecular ions

Queen's Hall in London. The surface area of the different materials is calculated, and the absorption coefficient is estimated. In the case of the chairs the absorption per chair is used. In several cases 10 per cent is deducted because the surfaces in question are not entirely freely exposed to the sound. In the second half the absorption is calculated separately for the various groups of people present in the hall. The absorption of the occupied chairs is of course deducted.

It is now possible to calculate the total absorption and thus the reverberation time, according to Sabine for different combinations as done in the third part of the table. It may be seen that the audience has a very great influence. In calculating the absorption of a hall it is certainly desirable to find out whether or not the reverberation time becomes too long when the hall is only partially occupied. There must be sufficient permanent absorption in the hall in the form of upholstered seats or on the walls. Then in order not to have too short a reverberation time when the hall is fully occupied, the volume of the room must not be too small.

Fig. 8 gives an idea of the reverberation times usual in practical cases. The reverberation time T is represented as a function of the volume V of the hall for various halls known to possess good acoustics.

ABSTRACTS OF RECENT SCIENTIFIC PUBLICATIONS OF THE N.V. PHILIPS' GLOEILAMPENFABRIEKEN

1319: W. de Groot and F. M. Penning: Die Ergiebigkeit der Neutronenproduktion durch D-Kanalstrahlen (Physica **5,** 512 - 520, June 1938).

From the yield of neutrons produced by deuterium ions upon striking plates of deuterium phosphate, lithium, beryllium and carbon as measured by Amaldi, Hafstad and Tuve, the effective diameters of these atomic nuclei were determined as functions of the speed of the projectile, use being made of the values found by Mano of the range in air of charged particles and of the atomic stopping power of these elements. With the help of the approximation formulae following from these values, it is possible to extrapolate the yields as found by Amaldi, Hafstad and Tuve for lower energies also. The results of Ladenburg and others, Scherrer, Kallmann and other authors can be made to agree with these formulae by a suitable choice of parameters. It is found necessary in some cases to determine once more the recalculation factor for the transition from a compound to pure deuterium, making use of the proportionality of the atomic stopping power to the square root of the nuclear charge number.

1320: J. H. de Boer: Interpretation of molecular phenomena by means of potential curves. IV. Fluorescence and phosphorescence; Activation energy (Ned. T. Natuurk. **5,** 57 - 64, Mar. 1938).

The significance of the Franck and Condon principle is discussed in relation to the position of absorption and emission bands of fluorescent substances. The chemical activation energy can be deduced from the potential curves. Activated adsorption and explosive substances are finally dealt with briefly.

1321: J. H. de Boer: Interpretation of molecular phenomena by means of potential curves. V. Colloids (Ned. T. Natuurk. **5,** 65 - 69, Apr. 1938).

Colloidal phenomena are discussed from the standpoint of potential curves according to the considerations of Hamaker (*cf* **1286**).

1322: C. J. Bakker, W. de Groot and F. M. Penning: Neutron-producing nuclear reactions and their yields (Ned. T. Natuurk. **5,** 102 - 115, May 1938).

A critical survey is given of the present insight into the production of neutrons by various nuclear reactions. The results obtained by others are compared with those of several measurements by the authors, and a satisfactory agreement is found. Special mention must be made of the fact that Penning was able to produce neutrons in a closed tube filled with deuterium at a low pressure with a discharge in a magnetic field in which the electrons must move in spiral paths between the electrodes. Deuterium molecular ions were formed, which, after being accelerated, struck a plate covered with a compound of zirconium and deuterium, thereby producing neutrons.

1323: J. de Boer: Sensitivity at different frequencies of a spherical model of a pressure gradient microphone (Physica **5**, 542 - 552, July 1938).

The sensitivity was calculated of a spherical model of a pressure gradient microphone as a function of the frequency of the sound wave of the radius of the sphere. For wave lengths greater than 2π times the radius of the sphere the sensitivity is found to be independent of the frequency. The calculated results agree with the results of measurements.

1324: M. J. Druyvesteyn: The interaction between an electron beam and a plasma (Physica **5**, 561 - 567, July 1938).

The way in which a beam of electrons is retarded and scattered in direction and energy by a plasma is calculated. Since the interaction takes place according to the Coulomb law, an upper limit must be determined for the distance at which an electron of the beam is still affected by the electrons and ions of the plasma.

1325: W. Elenbaas: Der Einfluß der positiven Ionen auf die Beweglichkeit der Elektronen im Quecksilberbogen (Physica **5**, 568 - 580, July 1938).

The electrical field strength in a high pressure mercury discharge is measured at pressures lower than those ordinarily used, and in tubes with diameters of 11, 19 and 37 mm. In this pressure range differences are found between the measurements and the calculations made on the basis of measurements in the normal pressure range. It is found that the deviations may be accounted for by considering the influence of the positive ions on the mobility of the electrons as calculated by Gvosdover.

1326: C. J. Bakker: Current distribution fluctuations in multi-electrode radio valves (Physica **5**, 581 - 592, July 1938).

The fluctuations are discussed which occur in the current in a radio valve due to the distribution of the current over the different electrodes. Formulae are derived for the fluctuations in the anode current of a pentode. Experiments carried out on pentodes agree satisfactorily with the theoretical expressions.

1327: J. A. M. van Liempt: Der Reflexionskoeffizient einiger organischen Stoffe (Rec. Trav. chim. Pays Bas **57**, 694 - 696, June 1938).

From the Raman frequency of an organic molecule and the variation of its sublimation pressure the reflection coefficient of the organic substance in question can be calculated by means of the formula (*cf.* **1051**) for the sublimation pressure. For the reflexion coefficients of benzene, p-dibrombenzene and naphtalene, values from 0.99 to 1 are found in this way.

1328: J. D. Fast Außerordentlich grosse Löslichkeit von Stickstoff und Sauerstoff in einigen Metallen; studiert an Zirkon und Titan (Metallwirtschaft, **17**, 641 - 644, June 1938).

Titanium and zirconium have an abnormally large capacity for dissolving nitrogen and oxygen. One result of this fact is that these gases have a very unfavourable influence on the mechanical properties of these metals. In the case of thorium, nitrides and oxides are formed which do not dissolve in the metal, and which are therefore much less harmful. It follows from X-ray examination that at the most, 40 atom per cent of oxygen and 20 of nitrogen dissolve in zirconium. The transition point ($865 \pm 10°$ C) from hexagonal (α) to cubic (β) zirconium changes, due to the solution of oxygen or nitrogen, into a transition range, the width of which depends upon the amount of gas dissolved; with 10 atom per cent of oxygen it is more than 600°. Since the specific weight of a rod of zirconium rises from 6.54 to 6.60 upon taking up 10 atom per cent of oxygen, it may be concluded that the oxygen atoms are taken up in the open spaces of the metal lattice. In contrast to the case with hydrogen, the oxygen and nitrogen taken up by zirconium and titanium cannot be driven out by heating in a high vacuum.

1329: W. G. Burgers: Elektronenoptische Beobachtungen der Zwillingsbildung in Nickeleisen (Metallwirtschaft **17**, 648 - 650, June 1938).

The concept proposed by Carpenter and Tamura, that recrystallization twins are due to defective crystal growth, and need not necessarily develop, for example by starting from preformed twin crystals along slip places, is confirmed in experiments carried out with the aid of the electron microscope on crystals of nickel-iron which exhibit twin structure. This of course does not prove at all that the concept must be generally valid.