THE BELL SYSTEM
TECHNICAL JOURNAL

VOLUME XXXII MARCH 1953 NUMBER 2

Ferrite Core Inductors

By H. A. STONE, JR.

(Manuscript received November 19, 1952)

This paper describes the use of ferrite materials as cores for inductors and develops methods for taking maximum advantage of their properties in the design of inductors for communication circuits.

INTRODUCTION

The extent to which the theoretical capabilities of wave filters and networks can be realized in practice usually depends on how high a Q, ratio of reactance to resistance, can be obtained in the inductors. In the voice and carrier telephone frequency ranges the dissipation in mica capacitors is so small compared to that in inductors that it can generally be neglected. Even paper capacitors, in the lower frequency ranges, compare favorably in Q with the best available coils. Consequently, there has been considerable incentive to develop improved magnetic materials that will permit the realization of higher Q inductors for filter and network use.

Work along this line has resulted in the development of the permalloys, and later the molybdenum permalloys which, powdered, insulated and pressed into shapes, have become the standard materials for wave filter coils in the voice and carrier frequency ranges.[1, 2]

Although the permalloys represent a vast improvement over the soft iron that preceded them they share the fundamental disadvantage of all metals, that they are good conductors. Any conductor in the vicinity of an alternating magnetic field has eddy currents induced in it, and, if the conductor is the core of a coil, the power loss associated with these currents appears as added resistance in the coil windings. To restrict the eddy current paths it is customary to powder the material and insulate the particles with some kind of inorganic filler. However, there are two practical limitations on this procedure. First, there is the mechanical difficulty of grinding particles to smaller than the few microns diameter that now represents the commercial limit. Second, as the size of the individual particles is made smaller more air gaps are introduced in the magnetic circuit by the insulating material, and the effective over-all permeability of the core is reduced. If carried too far the advantages in using the magnetic material in the first place are largely lost.

The need for high resistivity magnetic materials has been recognized for many years, and, in fact, the naturally occurring magnetic iron oxide, Fe_3O_4, or lodestone, is such a material. Its permeability, however, is only about 3 or 4 and its use in coil work has been very restricted. The problem of producing an oxide or mixture of oxides having higher permeabilities was investigated by Philips Gloeilampenfabrieken in Holland during the late war and they were successful in producing oxide mixtures, or ferrites, having permeabilities of 1500 and higher.[3]

After the war, Bell Telephone Laboratories initiated a program of ferrite development, and at the present time ferrite cores as well as inductors and transformers which take advantage of their unique properties are in commercial production.

The most commonly used ferrites consist of solid solutions of oxides of manganese, zinc and iron, or nickel, zinc and iron. They are prepared by mixing the materials, pressing them into the required shape, and heat treating them under carefully controlled conditions. The resulting ferrite parts are hard and brittle but they can be machined with diamond tools.

Although ferrites belong in the class of so-called semiconductors their conductivities are of the order of a millionth of those of metals and the eddy current losses are proportionately smaller. There are frequency limitations in ferrites but these occur at higher frequencies than are ordinarily of concern in magnetic core inductor work.

Besides offering the coil designer new possibilities because of low eddy current losses the ferrites have a secondary advantage in that they do not have to be subdivided to keep those losses low. The powdered metallic materials are fragile, are difficult to press except in simple shapes such as rings or cylindrical plugs, and are difficult to machine. As a result the coil designer, for virtually every use, is restricted to the toroidal type coil which, although it has some important advantages, is inherently expensive to wind and adjust. With ferrites, on the other hand, a variety of shapes can be produced. They have ample strength and can be ground and machined.

The combined magnetic and mechanical advantages of ferrites put them in an entirely new class as core materials for inductors. To exploit these advantages properly it is necessary to reconsider some of the fundamental aspects of coil design, since some of the assumptions that applied to powdered core coils are no longer valid, and others have to be modified.

DISSIPATION FACTOR

Legg has shown that the core loss characteristics of metallic magnetic materials can be described by three constants, defined as eddy current (e), hysteresis (a), and residual (c), coefficients, respectively.[1] The total increment of resistance (R_m), due to core losses is given by

$$R_m = e\mu f^2 L + a\mu B_m f L + c\mu f L \tag{1}$$

where μ = effective permeability of the core structure

f = frequency in c.p.s.

L = inductance in henrys

B_m = max. flux density in gausses

This expression is valid for practical applications so long as the flux density is low.

Measurements on ferrite cores indicate that the same formula can be used, with appropriate constants, at least for frequencies up to 200 or 300 kc, and, again, for low flux density applications. However, the emphasis on the terms becomes quite different. The residual loss, as its name implies, is usually of negligible importance in metallic magnetic materials, but in ferrites because of the very low value of the eddy current constant, the residual loss often emerges as a controlling factor.

Beside the core losses the other factors in an inductor that contribute to the total measured resistance are the dc resistance of the winding, ac losses in the winding and the parasitic capacitances within the structure. The latter two, especially the capacitance effects, have always had to be taken into account in higher frequency work, but their contribution to the total loss becomes a matter of first order importance when ferrites are used since they are no longer small in comparison with the eddy current loss in the core.

The objective in the design of inductors for wave filter and similar applications is to provide a specified inductance with as low an associated resistance as practicable. The quality factor, Q, the ratio of reactance to resistance, is the traditional measure of the extent to which this has been achieved. In the discussion that follows, however, it will be more convenient to use the reciprocal of Q as the measure of quality. This permits combining the loss components in a simple additive manner. Thus

$$\text{Dissipation Factor} = D = \frac{1}{Q} = \frac{R_{dc} + R_e + R_h + R_r + R_c + R_s}{\omega L} \quad (2)$$

where R_{dc} = dc resistance of the winding

R_e = eddy current loss

R_h = hysteresis loss

R_r = residual loss

R_c = increment of measured resistance due to distributed capacitance

R_s = ac loss in wire of the winding.

or

$$D = D_{dc} + D_e + D_h + D_r + D_c + D_s \quad (3)$$

where $D_n = \dfrac{R_n}{\omega L}$ (n being any of the above subscripts).

Having been given the inductance, and the operating frequency and current, for a desired inductor, and having chosen the core material that will be used, each of the D's above will be functions of the following variables: Permeability, Volume and Proportions of the core structure. We will now investigate how each of these factors can be manipulated to yield the highest Q (or lowest D) coil.

EFFECT OF PERMEABILITY ON DISSIPATION FACTOR

The permeability of a permalloy powder core is determined by the fineness of the powder and the amount of insulation with which it is mixed before firing. Obviously it is commercially impracticable to manufacture cores having a great variety of permeabilities, and it has happened that four values have been standardized for commercial use; 125, 60, 26 and 14. The coil designer working with permalloy powder, beyond choosing the most appropriate of these four values, seldom finds it practicable to control the permeability. This is due to the difficulty of pressing shapes suitable for use with air gaps, and the fragility of the parts. With ferrite, on the other hand, the parts can readily be adapted to the control of permeability by insertion of air gaps, and thus establishment of the optimum permeability becomes an important step toward achieving the desired coil performance.

The first four of the "D's" in (3), dc resistance and the core losses, are direct functions of the permeability and can be discussed to a certain extent independently of the last two.

The inductance of a magnetic core inductor is given by

$$L = \frac{kN^2A\mu}{\ell} \tag{4}$$

where k = constant

N = number of turns in winding

A = cross section of core

ℓ = mean length of magnetic path through core.

If the geometric proportions of the core have been prescribed, A and ℓ will bear a constant relationship to $V^{2/3}$ and $V^{1/3}$, respectively, and (4) may be written

$$L = k_1N^2V^{1/3}\mu \tag{5}$$

where k_1 = constant

V = core volume.

The dc resistance in the winding is

$$R_{dc} = \frac{\rho N^2\lambda}{k_wW} \tag{6}$$

where ρ = resistivity of the conductor

λ = mean length of turn

W = available area of cross section through which turns can be linked with the core

k_w = winding efficiency. This is the ratio of the actual cross section of conductor to the available area, W.

Again, for a core of given proportions (6) may be written

$$R_{dc} = \frac{k_2N^2}{V^{1/3}} \tag{7}$$

Eliminating N from (5) and (7)

$$R_{dc} = \frac{k_3 L}{V^{2/3}\mu} \tag{8}$$

and

$$D_{dc} = \frac{k_4}{V^{2/3}\mu f} \tag{9}$$

where k_2, k_3, k_4 are constants.

From (1) and (2) it is seen that

$$D_e = k_5 \mu f \tag{10}$$

$$D_h = k_6 B_m \mu = \frac{k_{10}\mu^{3/2} L^{1/2} i}{V^{1/2}} \tag{11}$$

$$D_r = k_7 \mu \tag{12}$$

in which the flux density,

$$B_m = \sqrt{2}\,\mu H = k_8 \mu \frac{Ni}{\ell} = k_9 \sqrt{\frac{\mu L i^2}{\ell A}} = k_9 \sqrt{\frac{\mu L i^2}{V}}. \tag{13}$$

H is the field strength due to the r.m.s. current, i, and k_5 to k_{10} are constants.

Combining (9) through (12)

$$D = \frac{k_4}{V^{2/3}\mu f} + k_5 \mu f + k_{10} \frac{\mu^{3/2} L^{1/2} i}{V^{1/2}} + k_7 \mu. \tag{14}$$

For a specified inductance at a given frequency and current the optimum permeability can be found from (14) by inserting the numerical values and solving graphically. It is very unlikely, however, that this cumbersome procedure would be necessary in practical work. Usually, depending on the design requirements, one or another of the core loss factors will be comparatively large and the others can be neglected. We will examine separately the three cases where residual, hysteresis and eddy current loss, respectively, predominate over the other core losses, and show how the permeability can be adjusted to minimize the dissipation factor.

1. *DC Resistance and Residual Loss Predominate*

This condition, which formerly was very seldom encountered, is becoming of increasing practical importance for two reasons. First, the eddy current losses in ferrite are extremely small. Second, the introduction of the transistor has created many new applications for inductors for use at very low power levels, and with proportionately lower hysteresis losses. For this condition we have, from (14)

$$D = \frac{k_4}{V^{2/3}\mu f} + K_7\mu \tag{15}$$

Solving for the value of μ, $\mu_{\text{opt.}}$, which will yield the lowest dissipation

$$\begin{aligned} \frac{dD}{d\mu} &= -\frac{k^4}{V^{2/3}\mu^2 f} + k_7 \\ \mu_{\text{opt.}} &= \frac{{k_4}^{1/2}}{k_7^{1/2}V^{1/3}f^{1/2}} \end{aligned} \tag{16}$$

and

$$D_{\text{opt.}} = 2\,\frac{{k_4}^{1/2}{k_7}^{1/2}}{V^{1/3}f^{1/2}}\,. \tag{17}$$

It may be noted from (16) and (17) that the optimum permeability is that which will result in the dc resistance and residual loss being equal.

2. *DC Resistance and Hysteresis Loss Predominate*

This condition, while it does not often apply to conventional size inductors for transmission networks, may occur when miniaturization of coils is contemplated. It will be noted from (14) that hysteresis is the only core loss component that is dependent on the volume of the core. We have, from (14)

$$D = \frac{k_4}{V^{2/3}\mu f} + \frac{k_{10}\mu^{3/2}L^{1/2}i}{V^{1/2}}\,. \tag{18}$$

The optimum permeability and dissipation factor derived from this equation are

$$\mu_{\text{opt.}} = \sqrt[5]{\frac{4k_4^2}{9k_{10}^2 f^2 L i^2 V^{1/3}}} \tag{19}$$

$$D_{\text{opt.}} = [(3/2)^{2/5} + (2/3)^{3/5}]\sqrt[5]{\frac{k_4^3 k_{10}^2 L i^2}{f^3 V^3}}\,. \tag{20}$$

Inspection shows that

(a) The optimum permeability is such that the ratio of dc resistance to hysteresis loss is 3/2.

(b) This value will depend not only on the frequency and volume of core, but on the inductance and current level as well.

3. *DC Resistance and Eddy Current Loss Predominate*

This condition is the most commonly occurring of all when metallic magnetic materials are used but is of very little importance for ferrite coil design. From (14)

$$D = \frac{k_4}{V^{2/3}\mu f} + k_5\mu f \qquad (21)$$

$$\mu_{\text{opt.}} = \frac{k_4^{1/2}}{k_5^{1/2}V^{1/3}f} \qquad (22)$$

$$D_{\text{opt.}} = \frac{2k_4^{1/2}k_5^{1/2}}{V^{1/3}}. \qquad (23)$$

Here we note that, as in the case for residual loss, optimum permeability is that which assures that the core loss is equal to the dc resistance. If both eddy current and residual losses were significant the sum of these should be equal to the DC resistance. It is also seen from (22) and (23) that although the optimum permeability is a function of frequency the resulting dissipation factor is independent of the frequency.

We have developed expressions for optimum permeability for three conditions of core loss. It is now of interest to know how critical the adjustment of permeability is. Fig. 1 shows the effect on the dissipation factor of deviations from optimum in permeability. It is seen that for

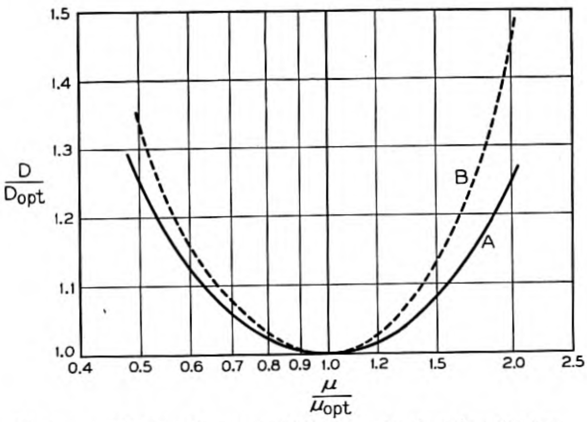


Fig. 1 — Effect of permeability on dissipation factor.

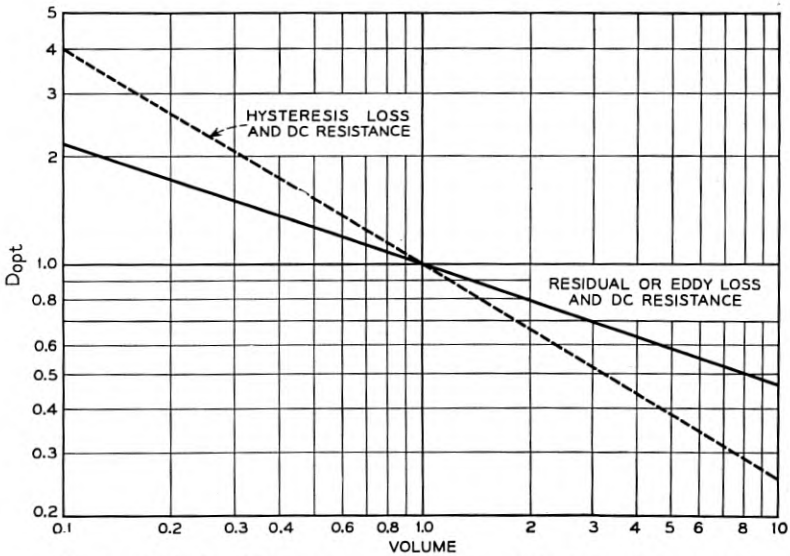


Fig. 2 — Relationship between core volume and dissipation factor.

deviations from optimum up to about 20 per cent the penalty in dissipation is not severe, but beyond this point the dissipation factor increases rapidly.

EFFECT OF VOLUME ON DISSIPATION FACTOR

It has been shown in the preceding section that the permeability can be adjusted for optimum coil quality but that the value to which it is adjusted depends on which of the loss factors predominates. It will also be seen from the foregoing that both the value of optimum permeability and the dissipation factor that can be realized by so adjusting the permeability depend on the volume of the core. The curves of Fig. 2 are derived from (17), (20) and (23), respectively, considering $D_{opt.}$ as a function of the independent variable, V. That is, they show how the dissipation factor will vary as a function of volume assuming that the permeability is adjusted as the volume is changed so that the lowest possible dissipation factor is always realized.

As would be expected, these curves show that, regardless of the relative magnitudes of the core losses, the dissipation factor can be decreased by increasing the size of the coil. There are serious limitations, however, on how far this process can be carried, and these are due to D_c and D_s, the resistive components resulting from capacitance and ac loss in the wire.

This ac loss is due to a combination of skin effect and eddy currents in the conductor. It is proportional to the sixth power of the diameter of each of the strands of wire that go to make up the conductor, and the square of the number of strands[4]. If the size of an inductor is increased, therefore, the ac loss may rapidly assume important proportions. Some help can be obtained by finer stranding of the wire and this has, in some cases, been carried to the extent of using 810 separately insulated strands in a single conductor. Even if fine stranding is sufficient to reduce the ac loss to a tolerable value, a penalty is paid in the form of higher dc resistance, due to the space occupied by the separate insulation on the strands. This amounts to a decrease in the winding efficiency, k_w, in (6).

The effect of distributed capacitance on the dissipation factor of a coil is a function not only of the volume, directly, but of the absolute value of the dissipation factor. If the dissipation factor is low compared to unity it is given, approximately, by

$$D_c = \frac{(D_0 + d)\dfrac{C}{C_g}}{1 - \dfrac{C}{C_g}} \tag{24}$$

where $D_0 = D - D_c$ is the dissipation factor that would obtain were it not for capacitance

d = dissipation factor of the distributed capacitance

C = distributed capacitance

C_g = resonating capacitance of the inductor.

If (24) is written in an alternative form

$$\frac{D}{D_0} = \frac{1 + \dfrac{d}{D_0}\dfrac{C}{C_g}}{1 - \dfrac{C}{C_g}} \tag{25}$$

it becomes evident that a coil with an inherently high Q is more seriously affected, proportionately, by distributed capacitance than is a low Q coil. Fig. 3 shows this information graphically, using the value 0.01 for d, which measurements on model inductors indicate to be of an appropriate magnitude.

Each of the curves in Fig. 3 is based on a constant value for C/C_g. In practice, however, as volume is increased it becomes more difficult to

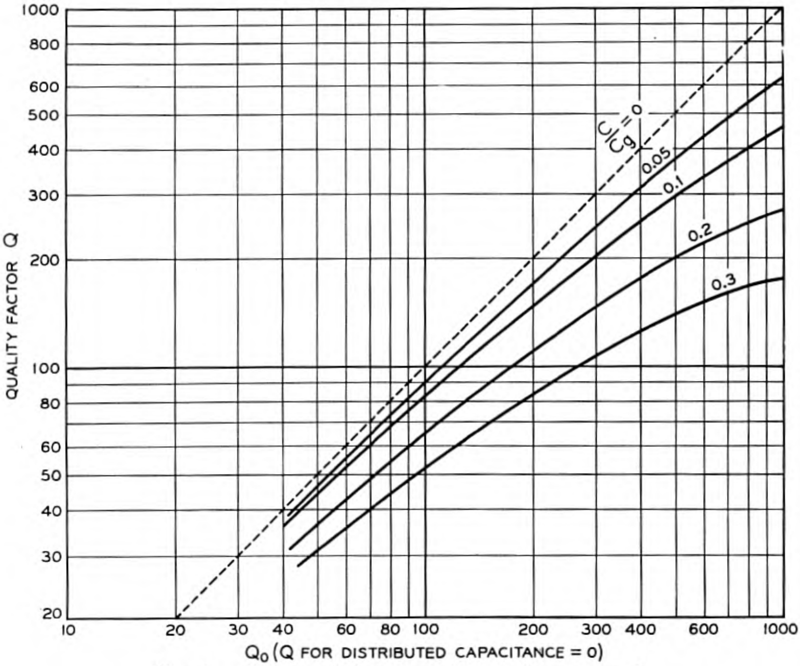


Fig. 3 — Effect of distributed capacitance on Q.

maintain small values of distributed capacitance. It will be necessary to increase the separation of windings from each other and from the core and this will result in a lower winding efficiency and a higher dc resistance.

Since the limitations imposed by ac losses in the wire and by distributed capacitance depend on the absolute magnitudes of the design parameters and the mechanical details of the inductor structure they do not lend themselves to representation in practicable generalized formulas as do the core losses. However, the following information on some model inductors will illustrate the magnitudes of these limitations.

A ferrite core coil was constructed similar to that shown in Fig. 4, but having a core volume of only 0.04 cubic inches. It was a 5 mh coil for use at 100 kc and had a distributed capacitance of 10 mmf. It was wound with a single strand conductor and the winding efficiency, k_w, was 0.4. Since it was intended for use at very low power levels the hysteresis loss was negligible. The permeability was optimized in accordance with case 1, above, for residual loss predominating. The measured Q was very close to 300. Thus, $D = 0.0033$. From the above data we note that $C/C_g = 0.02$, and from (25) we can calculate that $D_0 = 0.0030$,

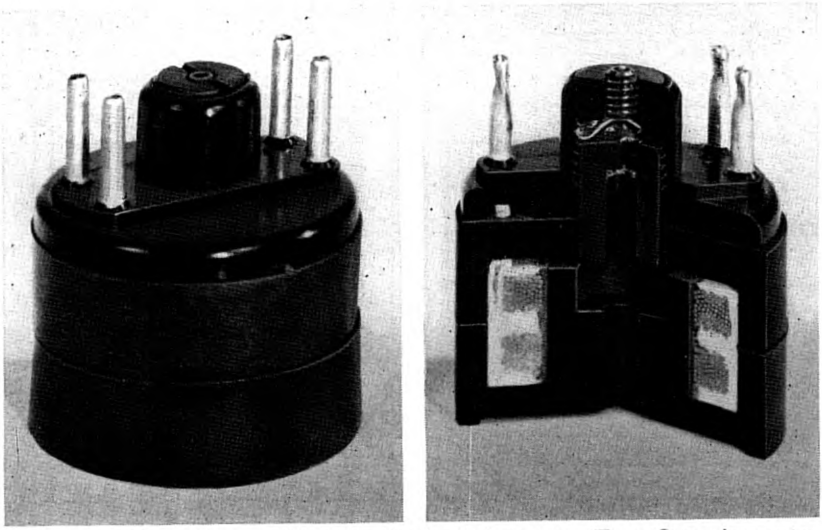


Fig. 4 — The 1509 type adjustable ferrite inductor for Type-O carrier system filters.

using the value 0.01 for d. Thus the "geometric Q", that which would obtain were it not for the distributed capacitance, is about 330.

The 1509 type inductor, shown in Fig. 4, was designed for use in the type-O carrier system.[5] It has a volume of 1.4 cubic inches, 35 times that of the small coil discussed above. From the standpoint of dc resistance and core loss alone, (7) indicates that the dissipation factor should be

$$D_0 = \frac{1}{35^{1/3}} (0.0030) = 0.00092,$$

or that the Q should be greater than 1000.

In order to keep the capacitance to the same order as that of the small coil, and to use stranded wire to avoid excessive ac loss, the winding efficiency, k_w, had to be reduced from 0.4 to 0.13. The effect of this on dissipation can be shown from (17), remembering that k_4 varies inversely as k_w:

$$\frac{D_0'}{D_0} = \frac{(3k_4)^{1/2}}{k_4^{1/2}} = 1.73$$

$$D_0' = (1.73)(0.0092) = 0.00159.$$

Thus, the Q (still neglecting the effect of the 10 mmf distributed capacitance that remains after reducing the winding efficiency) should be 630.

From (24)

$$D_c = \frac{(0.00159 + 0.01)0.02}{0.98} = 0.00024$$

and

$$D = D_0 + D_c = 0.00183$$

or the actual measured Q of the large inductor is 550, about half that indicated by core loss considerations alone.

EFFECT OF CORE PROPORTIONS ON DISSIPATION FACTOR

In the foregoing discussion of coil volume it has been assumed that the core porportions remained fixed as the volume was changed. It is now of interest to know what these proportions should be to insure the lowest dissipation.

The general type of structure under consideration consists of a closed cylindrical container and a center post of magnetic material, and an air gap which might be anywhere in the magnetic path. It is assumed that the thickness of the shells is such that the area of the magnetic path is uniform and equal to the area of cross section of the post. It is also assumed that the flux is uniformly distributed within the cross-section of the core. Fig. 5 represents such a core schematically.

Given a fixed over-all coil volume it is desired to know what proportions should apply to the outside diameter, the diameter of the center post and the axial height of the structure, to provide the lowest dissipation factor. This will be examined first for the case where residual or eddy

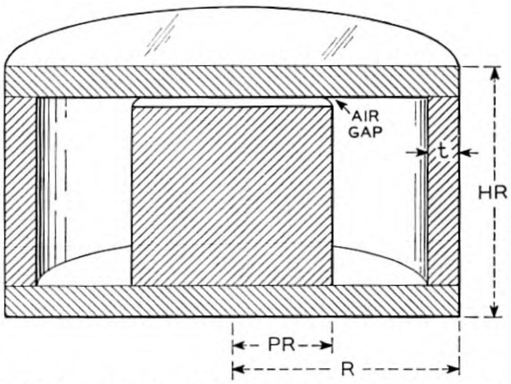


Fig. 5 — Optimum proportions for post and shell core assembly.

current losses are large compared to hysteresis loss, and then for the situation where hysteresis loss predominates. Referring to Fig. 5, the following relationships are noted.

$$t \text{ (thickness of shell) is given by } (R - t) = R\sqrt{1 - P^2} \tag{26}$$

$$A \text{ (magnetic cross-section)} = \pi P^2 R^2 \tag{27}$$

$$\ell \text{ (mean magnetic path)} = R(3\sqrt{1 - P^2} - 1 + 2H) \tag{28}$$

$$\begin{aligned} &W \text{ (available winding cross-section)} \\ &\qquad = R^2(\sqrt{1 - P^2} - P)(2\sqrt{1 - P^2} - 2 + H) \end{aligned} \tag{29}$$

$$\lambda \text{ (mean length of turn)} = \pi R(\sqrt{1 - P^2} + P) \tag{30}$$

$$V_c \text{ (coil volume)} = \pi R^3 H \tag{31}$$

1. *DC Resistance, and Residual or Eddy Current Losses, Predominate*

The dissipation factor, due to the dc resistance, is seen from (4) and (6) to be

$$D_{dc} = \frac{R_{dc}}{2\pi f L} = k_{11} \frac{\lambda \ell}{WA\mu} \tag{32}$$

where $k_{11} = (\rho/2\pi k k_w)$ is a constant with respect to the core proportions, assuming the winding efficiency is not materially affected by changes in shape.

Since, regardless of shape, we will want to adjust the air gap so that μ is optimum, and since we know from the previous discussion that this will occur when the dc resistance is equal to the sum of the core losses, we have, from (10) and (12)

$$D_{dc} = D_e + D_r = (k_5 f + k_7)\mu \tag{33}$$

Eliminating μ from (32) and (33):

$$D_{dc} = k_{12} \sqrt{\frac{\lambda \ell}{WA}} \tag{34}$$

where $k_{12} = \sqrt{k_{11}(k_5 f + k_7)}$ is again a constant with respect to the core dimensions. The dissipation factor is

$$D = 2D_{dc} = 2k_{12} \sqrt{\frac{\lambda \ell}{WA}} \tag{35}$$

Putting in the values, from (27) to (30), for λ, ℓ, W and A:

$$D = \frac{2k_{12}}{R}\sqrt{\frac{(\sqrt{1-P^2}+P)(3\sqrt{1-P^2}-1+2H)}{P^2(\sqrt{1-P^2}-P)(2\sqrt{1-P^2}-2+H)}} \quad (36)$$

Since the volume of the coil, given in (31), is assumed to be constant, we can eliminate R from the above equation:

$$D = k_{13}\sqrt{\frac{(\sqrt{1-P^2}+P)(3\sqrt{1-P^2}-1+2H)H^{2/3}}{P^2(\sqrt{1-P^2}-P)(2\sqrt{1-P^2}-2+H)}} \quad (37)$$

where $k_{13} = \dfrac{2\pi^{1/3}k_{12}}{V_c^{1/3}}$

We now have an expression for the dissipation factor in terms of the two variables, P and H, which determine the proportions of the coil structure. The effects of these proportions on the dissipation factor are shown in Fig. 6. It will be seen that best results are achieved when the radius of the post is approximately 0.45 of the outside radius and the axial height is about 1.2 times this radius. Fig. 5 is drawn to this scale, and the 1509 type coil shown in Fig. 4 approximates these proportions.

2. *DC Resistance and Hysteresis Loss Predominate*

We have noted that for this case optimum permeability is that which results in the following relationship between the dc and hysteresis dissipations:

$$3D_h = 2D_{dc} \quad (38)$$

From (11) and (13)

$$D_h = k_6B_m\mu = \frac{k_6k_9\mu^{3/2}L^{1/2}i}{\ell^{1/2}A^{1/2}} \quad (39)$$

Putting the values from (32) and (39) in (38), we can eliminate μ and express D_{dc} in terms of the dimensional variables of the coil:

$$D_{dc} = k_{14}\frac{\lambda^{3/5}\ell^{2/5}}{W^{3/5}A^{4/5}} \quad (40)$$

where $k_{14} = (\frac{3}{2}k_6k_9k_{11}{}^{3/2}L^{1/2}i)^{2/5}$

The total dissipation factor is

$$D = D_{dc} + D_h = \tfrac{5}{3}D_{dc} = \tfrac{5}{3}k_{14}\frac{\lambda^{3/5}\ell^{2/5}}{W^{3/5}A^{4/5}} \quad (41)$$

Using (27) to (30) we can express D in terms of the coil proportions

$$D = \tfrac{5}{3}k_{14} \sqrt[5]{\frac{(\sqrt{1 - P^2} + P)^3(3\sqrt{1 - P^2} - 1 + 2H)^2}{R^9P^8\pi(\sqrt{1 - P^2} - P)^3(2\sqrt{1 - P^2} - 2 + H)^3}} \tag{42}$$

Again, eliminating R by use of the constant volume relationship, (31)

$$D = k_{15} \sqrt[5]{\frac{H^3(\sqrt{1 - P^2} + P)^3(3\sqrt{1 - P^2} - 1 + 2H)^2}{P^8(\sqrt{1 - P^2})^3(2\sqrt{1 - P^2} - 2 + H)^3}} \tag{43}$$

where $k_{15} = \dfrac{5\pi^{2/5}k_{14}}{3V_c^{3/5}}$

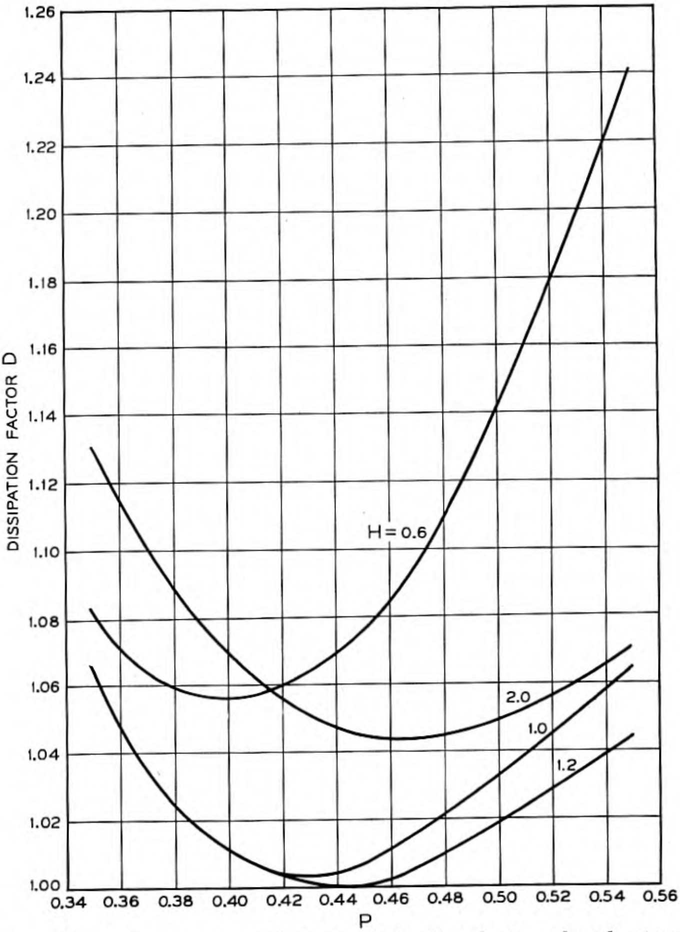


Fig. 6 — Effect of core proportions on dissipation factor when hysteresis loss can be neglected.

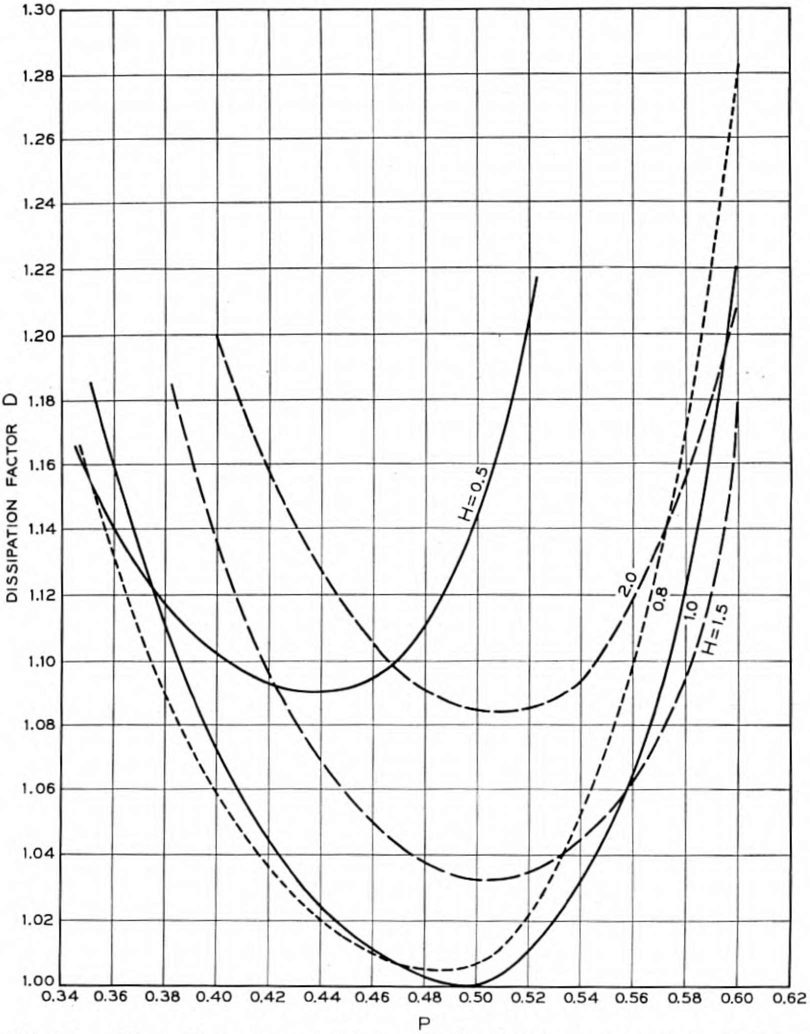


Fig. 7 — Effect of core proportions on dissipation factor when hysteresis loss is predominate.

This information is shown graphically in Fig. 7. It will be seen that when the hysteresis losses are important both the post diameter and the axial height should be about half of the overall diameter. These proportions are not far different from those derived as optimum from (37). This is fortunate since it means that cores of the same proportions will be suitable in different applications regardless of which core losses predominate.

METHODS OF INDUCTANCE ADJUSTMENT

The fact that ferrites can be molded in a variety of shapes, can be machined, and can be used with discrete air gaps permits considerable freedom in the mechanical methods that can be used to provide continuous adjustability of inductance. Whether the need is for factory adjustment to prescribed values or for adjustment after the coil is assembled in its equipment, the basic methods are the same.

We have noted, in (4) that the inductance of a magnetic core coil is given by

$$L = \frac{kN^2A\mu}{\ell} \tag{4}$$

in which k is a constant depending on the units used, and the other four quantities are design variables. Any of these, or any combinations of them, can be manipulated to produce variations in the inductance. They are: length of magnetic path, ℓ; cross section of magnetic path, A; number of turns, N; and effective permeability, μ. There is an additional variable hidden by the formula's assumption that perfect coupling exists among all the turns in the winding. Inductance can be adjusted by changing the coupling between parts of the winding. Each of these five variables, which are of widely differing practical value, will be commented upon at least briefly.

1. *Adjustment of Inductance by Change in Magnetic Path Length, ℓ*

Adjustment by a change in ℓ without an accompanying change in μ requires that the air gap also be changed since the effective permeability is a function of the ratio of the air gap dimensions to those of the total structure.

$$\mu = \frac{\mu_m}{1 + \mu_m \dfrac{g}{a}} \tag{44}$$

where μ_m = permeability of the core material

g = ratio of the length of air gap to total length, ℓ

a = ratio of effective cross section of gap to core cross section, A.

It would be mechanically possible to design an adjustable inductor of this sort but it is unlikely that there would be any practical advantage in maintaining constant permeability over the adjusting range. On the contrary, it would probably be more advantageous as well as mechanically simpler to have the air gap length constant so that the increase in inductance due to shortening the magnetic path would be augmented by an increase in permeability. In a device such as shown schematically in Fig. 8, as the two windings approach each other the over-all magnetic path decreases and at the same time the area of the air gap increases. Actually, the preponderant effect is due to the air gap change and the effect of variation in length of path becomes of secondary importance. This is likely to be the case generally, when both path length and air gap change simultaneously, since most of the reluctance is in the air gap.

2. *Adjustment of Inductance by Change in Magnetic Cross Section, A*

To provide for change of inductance by constriction of a part of the magnetic cross section is apt to be undesirable since it forces a concentration of flux in the constricted part of the magnetic circuit and may introduce unduly high hysteresis losses. Even if the levels are low enough so that this is not a consideration the amount of inductance variation that can be achieved even by a large constriction in part of the core is relatively small. To illustrate this we will consider a structure such as shown schematically in Fig. 9, in which one sector of the core can be varied in effective cross section. The reluctance of the structure is equal to the sum of the reluctances of the fixed part of the core, the sector of length $n\ell$, whose cross section aA can be varied, and the air gap:

$$\begin{aligned} R &= \frac{(1-n)\ell}{\mu_m A} + \frac{n\ell}{\mu_m a A} + \frac{g\ell}{A} \\ &= \frac{\ell}{\mu_m A}\left[1 + n\left(\frac{1}{a} - 1\right) + g\mu_m\right]. \end{aligned} \tag{45}$$

Let us assume the arbitrary but reasonable values of 2000 for μ_m, 0.01 for g, and 0.1 for n. Then approximately

$$R = \frac{\ell}{2000A}\left(20 + \frac{0.1}{a}\right).$$

It will be seen that as a is varied from 1.0, corresponding to the full cross section of the main core, to one-tenth of that, the change in inductance, which is inversely proportional to reluctance, will only amount to about 5 per cent.

3. *Adjustment of Inductance by Change in Number of Turns, N*

An adaptation of turns adjustment for use in a shell type structure is shown in Fig. 10.[6] The center post with the winding on it can revolve and turns can be removed by pulling on the outer lead, or added by rotating the knob at the end of the shaft. Continuous adjustment is possible since it is not necessary to add or remove integral numbers of turns. Although an inductor of this sort involves some mechanical complexity it has the advantage that the core parts do not have to be precisely machined. The turns adjustment can be used to compensate for sizable variations in the dimensions of the core parts and the resulting air gaps.

4. *Adjustment of Inductance by Change in Permeability, μ*

Adjustment by change in permeability, that is, change in length or area of air gaps, can be accomplished in a variety of ways to meet differing design needs and can be mechanically simple and economical. For most purposes permeability adjustment will offer more advantages than any of the other methods.

Whatever the method of adjustment used its effectiveness will depend on proper correlation between the mechanical motion that produces the change and the inductance itself. For most filter and network applications the following considerations will apply:

(a) The slope of the line showing inductance plotted against the displacement that produces the adjustment should be reasonably constant over the adjusting range.

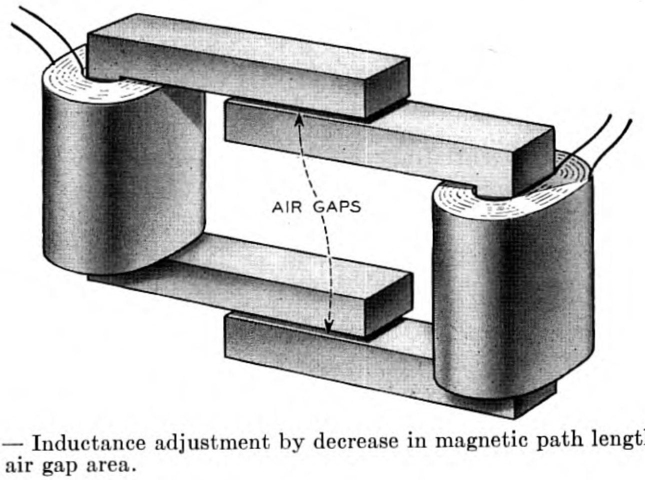


Fig. 8 — Inductance adjustment by decrease in magnetic path length and increase in air gap area.

(b) The slope should not be negative at any position in the adjusting range as this would introduce the possibility of false or double balance when the coil is being adjusted by null or peaking methods.

(c) The slope should not be so small that play in the mechanical parts will cause changes in slope of the same magnitude as the slope itself.

(d) Within the above limitation the smaller the slope the more precisely the adjustment can be made.

(e) Conversely, the greater the slope the greater the range of adjustment for a given amount of mechanical motion.

(f) It follows from (d) and (e) that the amount of mechanical motion available determines the product of range and precision. Where the mechanical motion is rotary, such as with adjustment by turning a screw, it is possible to adjust a coil to a precision of about 1/400 of a revolution without undue difficulty. If the range is covered by N revolutions of the adjusting screw, and the over-all range is $\pm R$ per cent of the mean value:

$$P = \frac{2R}{400N} = \frac{R}{200N}$$

where P = the precision of adjustment in per cent of the nominal value.

In the coil shown in Fig. 4, six turns of the adjusting screw are effective in producing an over-all change of ± 15 per cent in the inductance. The precision with which the adjustment can be made is, therefore, very close to ± 0.01 per cent of the value desired.

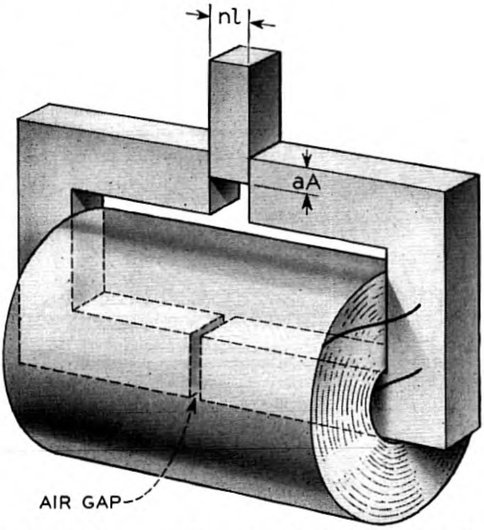


Fig. 9 — Inductance adjustment by constriction in magnetic cross section.

(g) It would appear from the above that by gearing or other means of increasing the mechanical motion that governs the inductance change it should be possible to improve the precision. This is true up to the point where consideration (c) is violated, to the point where play in the mechanical parts permits variations of the same order as the nominal precision.

An obvious form of permeability adjustment would consist of a means for varying the distance between two plane magnetic surfaces, as shown schematically in Fig. 11 (a). One disadvantage of this, or of any other means that involves a change in air gap length, is in the nonlinearity of adjustment. The effective permeability of a coil is given in (44). In most voice and carrier frequency applications of ferrite the magnitudes of g and a will be such that, approximately

$$\mu = \frac{a}{g} \tag{46}$$

and, if the cross section of the air gap is the same as that of the core

$$\mu = \frac{1}{g}.$$

A typical adjustment curve resulting from this inverse relationship is shown in Fig. 11 (b).

In addition to its nonlinearity the simple butted gap has a shortcoming in that the mechanical motions involved are of the order of only a few hundredths of an inch, which requires that parts be very accurately fitted. This can be somewhat alleviated by using cone or wedge shaped

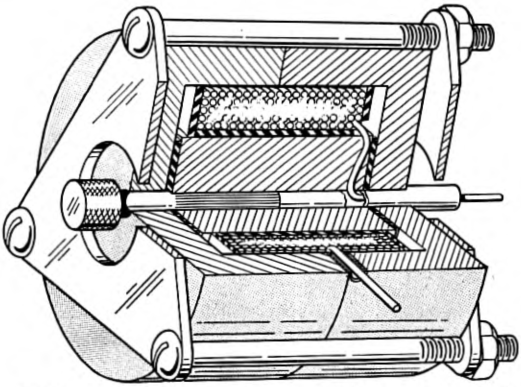


Fig. 10 — Inductance adjustment by addition or removal of turns.

gaps, as illustrated in Fig. 11 (c). Here the distance d, travelled by the screw is longer than the effective air gap g, by the amount

$$d = \frac{g}{\sin\ \theta}.$$

One means of overcoming the nonlinear characteristic of gaps such as these is to introduce compensation in the form of a secondary gap that opens as the main gap closes.[7] Such an arrangement is shown in Fig. 12 (a). When there are two gaps in series their reluctances are additive and their effect on permeability is given approximately by

$$\mu = \frac{1}{g_1 + g_2}. \tag{47}$$

Fig. 12 (b) shows the inductance characteristics that would result from either of the two gaps alone, and the effect of the two gaps in series.

From (46) it can be seen that the effective permeability varies ap-

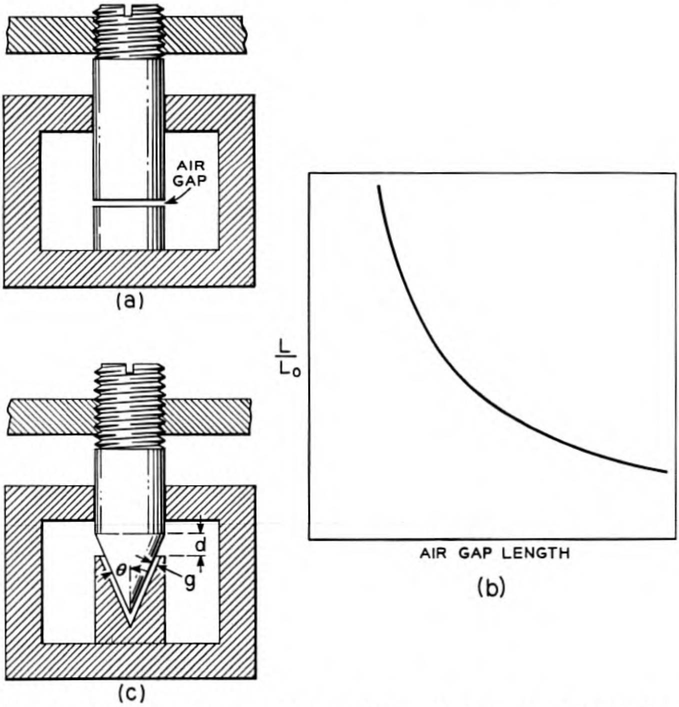


Fig. 11 — Inductance adjustment by variation of air gap length. (a) Gap formed by parallel plane surfaces. (b) Adjustment characteristic. (c) Cone or wedge shaped gaps.

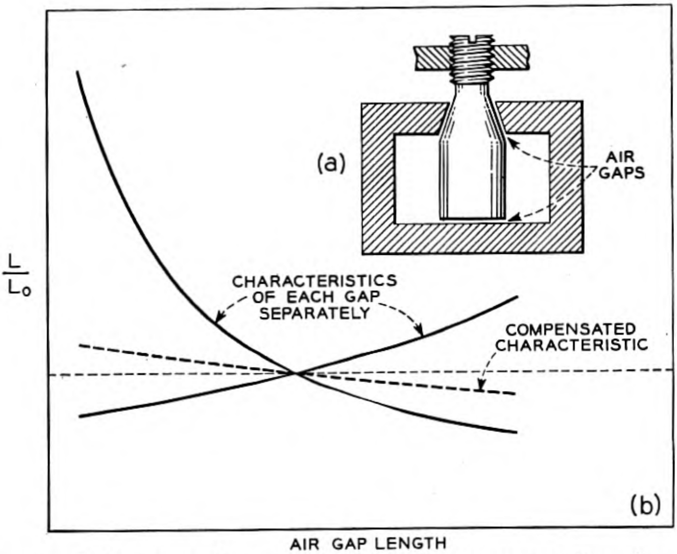


Fig. 12 — Inductance adjustment with partially compensating air gaps.

proximately directly as the ratio of air gap area to magnetic core area. Adjustment by variation of the area of the gap, therefore, is not subject to the nonlinearity that results from manipulating the air gap length. A simple method for providing adjustment by varying the effective area of the air gap is shown in Fig. 13. As one shell is rotated with respect to

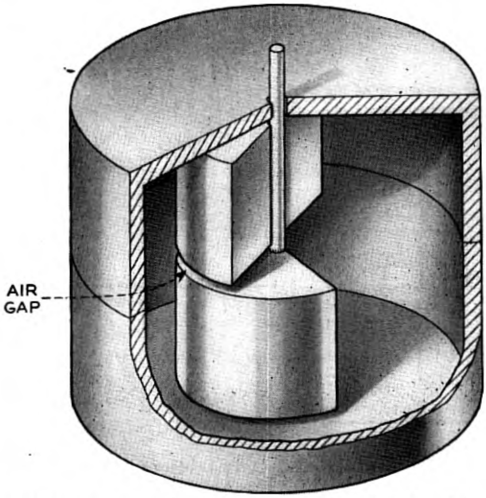


Fig. 13 — Inductance adjustment by variation in air gap area.

the other the area of registration between the semicircular faces of the center posts is reduced or increased. A simple arrangement such as this provides good linearity but has two limitations: (1) The total mechanical motion available for adjustment is only 1/2 revolution. (2) As the cores are rotated to reduce the air gap area and the inductance, the magnetic flux in the cores tends to concentrate more and more in the vicinity of this reduced area, and may under some conditions give rise to high hysteresis loss.

The method of adjustment used in the 1509 type inductor, Fig. 4, is essentially a variable area method but it is designed in such a way as to overcome these two limitations.[8] The air gap arrangement, visible in the cutaway view, consists of two gaps in parallel. The main annular gap is fixed and is large enough in area to insure that under no anticipated conditions of operation will the flux concentration be too high. The screw adjustment moves a cylindrical ferrite part into a depression in the center post, as shown. The effective cross section of this adjustable gap is approximately determined by the amount of surface of the cylinder within the depression. The total useful adjusting range corresponds to about six full turns of the adjusting screw.

5. *Adjustment of Inductance by Change in Coupling*

The overall inductance of two coils of equal inductance connected in series is

$$L = 2(1 + k)L_1 \tag{48}$$

where L = series inductance

L_1 = inductance of either coil

k = coupling coefficient

k may have any value between -1 and $+1$, these values corresponding to complete coupling and the windings connected in series opposing and series aiding, respectively. It is practicable to make inductors whose coupling can be continuously adjusted from very high positive values through zero to equally high negative values. This type of design is especially useful where a very wide range of inductance variation is desired. It will be seen from (48) that with couplings of plus and minus 90 per cent, respectively, in the extreme positions a range of 19 to 1 in inductance variation would result. A coil of this type is shown in Fig. 14.

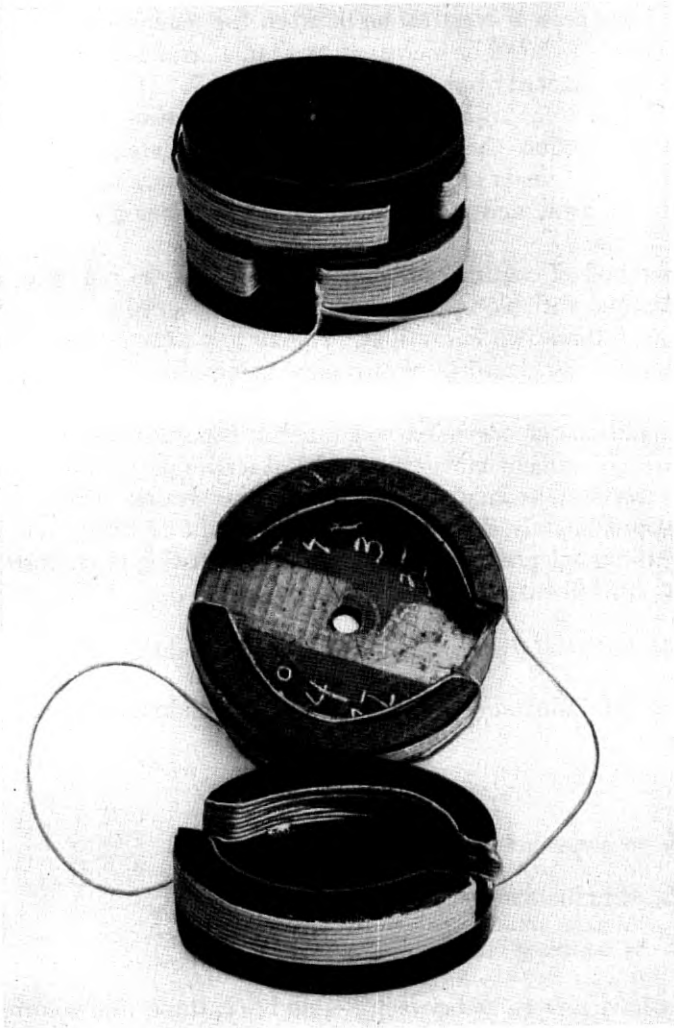


Fig. 14 — Magnetic variometer. Inductance adjustment by variation in mutual inductance.

CONCLUSION

The combined magnetic and mechanical characteritics of ferrite permit the design of inductors superior to those using metallic cores in at least three important respects:

1. Higher values of Q are obtainable than have ever before been practicable. In the range from about 50 to 200 kc, frequently used for telephone carrier, it is very difficult to realize a Q above about 300 in a

metallic core coil. Ferrite coils, on the other hand, have been made having Q's as high as 1000, and inductors with Q's between 500 and 600 are available commercially.

2. Formulas have been derived which show how ferrite can be used to realize the best inductor characteristics in the smallest volume. The 1509 type inductor, for instance, is only about 1/3 as large as the nearest equivalent permalloy core coil, yet its Q is over twice as high.

3. It is physically practicable in ferrite coil designs to include inductance adjustment facilities to meet a wide range of requirements.

It should not be concluded that the day of metallic cored inductors is over. For power applications, especially those involving direct current, present-day ferrites are inferior to silicon iron and permalloy. At voice frequencies there are many applications for which ferrite is, at best, no better than some of the older materials, although in others it has distinct advantages. In higher frequency ranges, however, and especially for low power level applications, the advantages of ferrites are outstanding enough to justify the expectation that they will largely replace the older iron and nickel-iron powders.

ACKNOWLEDGEMENTS

The development of ferrites and their application to inductors has been carried out by several teams in Bell Telephone Laboratories, including J. H. Scaff, F. J. Schnettler and their associates in the metallurgical department, V. E. Legg and C. D. Owens in the magnetic applications group, and S. G. Hale, R. S. Duncan and others in the inductor development area. I don't know who was first to conceive of using "dissipation factor" instead of "Q" to simplify his mathematics, but it was not the author. Equation (24) is derived by inserting $1/D_n$ for Q in an expression originally worked out by P. S. Darnell.

REFERENCES

1. Legg, V. E., Survey of Magnetic Materials and Applications in the Telephone System. Bell Sys. Tech. Jl., **18,** pp. 438–464, July, 1939.
2. Legg, V. E., and F. J. Given, Compressed Powdered Molybdenum Permalloy, Bell Sys. Tech. Jl., **19,** pp. 385–406, July, 1940.
3. Snoek, J. L., Non-Metallic Magnetic Materials for High Frequencies, Philips Tech. Review, **8,** pp. 353–360, December, 1946.
4. Wien, M., Über den Durchgang schneller Wechselströme durch Drahtrollen, Ann. D. Phys. [4], **14,** P. I, 1904.
5. Edwards, P. G., and L. R. Montfort, Type-O Carrier System, Bell Sys. Tech. Jl., **21,** pp. 638–723, July, 1952.
6. Legg, V. E., and C. D. Owens, Patent Application Serial No. 184602, Filing Date — 9/13/50.
7. Hale, S. G., and C. W. Nuttman, Patent Application Serial No. 263564, Filing Date — 12/27/51.
8. Duncan, R. S., and H. A. Stone, Patent Application Serial No. 262248, Filing Date — 12/18/51.

A Throwdown Machine for Telephone Traffic Studies

BY G. R. FROST, WILLIAM KEISTER AND
ALISTAIR E. RITCHIE

(Manuscript received December 3, 1952)

In order to study the traffic-carrying characteristics of the No. 5 crossbar switching system, a machine has been built to simulate the operation of the system. This machine, known as a throwdown machine, is controlled by a team of four operators. Its input is a statistically accurate representation of telephone traffic and its output is a detailed record of the course of each call through the system. This paper discusses the design principles of the throwdown machine, its operation, and the type of results obtained.

INTRODUCTION

Existing analytical methods are inadequate for investigating many statistical problems in which a large number of variables and their interactions must be considered. The problem of evaluating the performance and traffic capacity of a large automatic telephone switching system is one example. Others involve logistics, air and highway traffic control, and certain phases of military and naval strategy. All these require the assimilation of large quantities of data, processing the data according to certain procedures which are often empirical, and producing final information from which performance of the system or the excellence of the procedures can be judged.

These problems fall in the general category of "systems evaluation." The types of systems considered are those that are capable of a large number of variations depending on the nature of the input data, and must be judged on a statistical basis. One method of study might be to operate and observe an actual system. There are a number of objections to this. Operation may be so slow that the accumulation of sufficient data may require excessive time or, as in the case of a telephone switching system, so rapid that it is impractical to make the necessary observations. Operating the system under controlled conditions in these cases may be too expensive or indeed impossible. Then, too, the system may be proposed only and not yet exist.

One solution is to devise a method of simulating the performance of the actual or proposed system which through the use a suitable time scale will permit the necessary information to be obtained. The simulation may be done entirely on paper by recording each state of the system and modifying this state with each bit of input information according to the system plan, as though a log were being kept of the performance of the actual system.

Since the general problem involves large quantities of input data which are statistical in nature, all possible variations cannot be studied. A sufficient number of typical situations must be tried to obtain statistically reliable results. These methods have been extensively used in telephone traffic studies and are called "throwdown" studies. The name stems from the use of dice in the early study of telephone traffic problems. Each die is designated to represent a particular independent event and the faces of this die are designated according to the probability of the event taking place. By repeatedly "throwing down" a number of such dice and observing the results, the probability of a particular combination of events taking place can be estimated. Other similar methods based on selections from lists of random numbers have been used in telephone traffic studies for a number of years. Recently, mathematicians, using digital computers, have employed similar statistical methods in problems relating to the diffusion of gases, electron ballistics and the solution of certain types of differential equations. They have called this the "Monte Carlo" method.

Various mechanical aids can be used in running a throwdown study. This paper will discuss the techniques of throwdown studies and will describe a semi-automatic throwdown machine which was constructed for studies of the new No. 5 crossbar switching system used in local telephone central offices. A general view of this machine is shown in Fig. 1. It is a system of electrical switching circuits, signal lamps and mechanical devices which simulates a large telephone switching system and its associated subscribers. The machine is controlled by a team of four operators. Artificially generated telephone traffic is processed by this machine in a manner analogous to the action of the actual system. Detailed records are made of the progress of each call and the traffic situations encountered. After a sufficient number of calls have been processed, the recorded information can be analyzed by statistical methods to obtain desired information. The action of the system is simulated in sufficient detail to insure that results are representative of actual

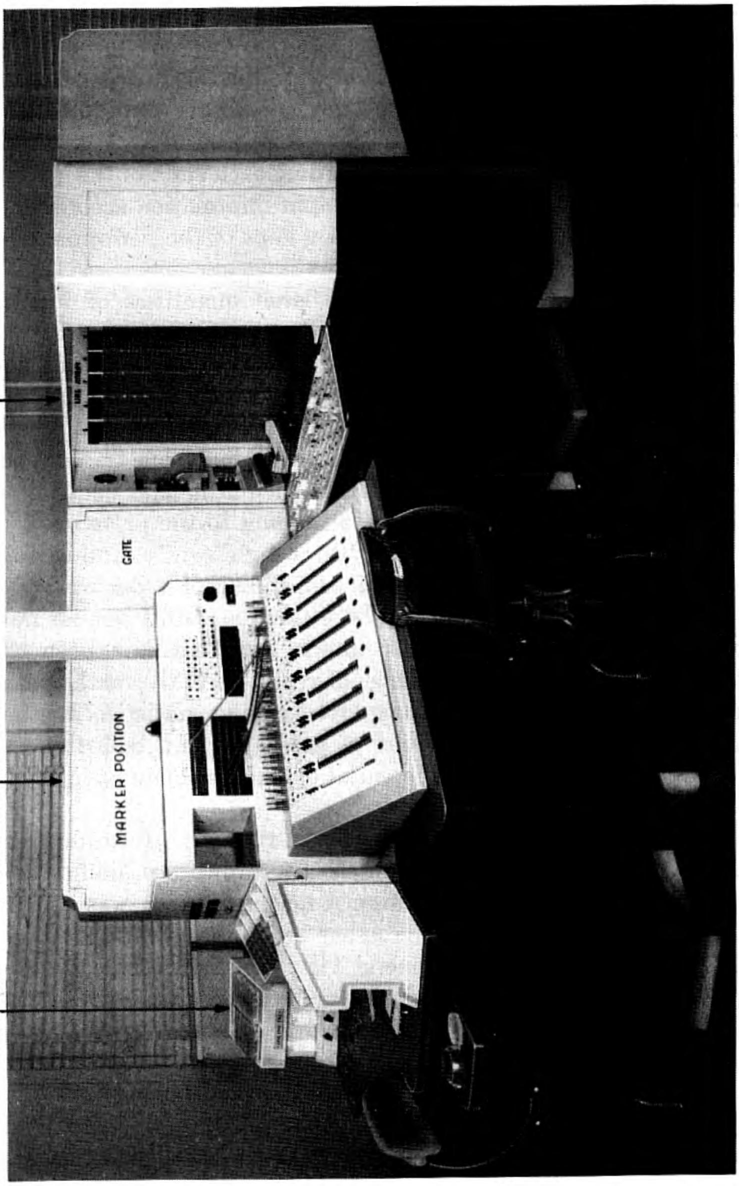


Fig. 1—General view of No. 5 system throwdown machine.

system performance. The level and character of submitted traffic can be varied, and a wide range of system sizes with varying quantities of control circuits can be tested.

Before the throwdown machine was built a "paper" throwdown trial of a small No. 5 crossbar system installation had been conducted by Mr. R. I. Wilkinson. This was run by a team of girls using card files, ledgers and written records, and using dice to make certain random decisions. The machine is basically a mechanization of these early methods to make possible the testing of larger installations in a reasonable time. The methods of generating data for the machine were developed by Mr. Wilkinson and many of the decisions relating to telephone traffic and statistical problems which were encountered in designing the machine were solved in consultation with him.

THE TELEPHONE TRAFFIC PROBLEM

A large automatic telephone switching system of the common control type is not a simple mechanism nor is evaluating its performance and traffic carrying capacity a simple problem. The economy of these systems depends upon the efficient use of relatively small groups of circuits on a time-sharing basis to serve a large number of subscribers. Each group of circuits is specialized to perform certain of the functions necessary in establishing a connection, and circuits from several groups must cooperate to handle every call. A sequence of actions with appropriate alternatives at several stages where busy conditions may be encountered is completely prescribed for every call. However, this sequence is subject to interference due to simultaneous requests to use the same control circuits. Competition is resolved by preference arrangements which cause some requests to be delayed while others are being served. Delays will increase the holding time of circuits with the possibility of causing traffic congestion at other points in the system.

With a number of subscribers originating calls at random, it becomes difficult to predict what the reactions of the system will be at various traffic levels. Although some parts of the problem can be solved by analytical methods employing probability theory, it is doubtful that mathematical means, beyond rough approximations, are available for evaluating the performance of an entire system.

Where systems have been built and placed in operation, the performance can be judged from observations of the working system. There are obvious weaknesses in this procedure. Only by collecting large quantities of information can the performance at various load levels be determined. Practical systems, as a matter of economy, are not equipped with indicating mechanisms to show performance at all stages. Events take place rapidly and the causes leading up to a particular traffic situation cannot be easily observed. Traffic loads cannot be repeated under controlled conditions. Size of installation and quantities of working equipment can be varied only by small amounts in a working office. To test offices of various sizes requires that these offices be built and installed. Variations in the system operation require actual changes in a working system. This can be done only to a limited extent.

In the case of newly developed systems, estimates of performance can be made on the basis of engineering judgment and experience with older systems of a similar type. This can be followed by a trial installation of a working system of a modest size which will test the system for flaws in design as well as provide information on traffic capacity which can be extrapolated to indicate approximate quantities of equipment for larger installations. At best this is a slow and expensive process and engineering data must be continually revised as experience is gained with larger installations. When the system is to be used extensively in installations of various sizes, methods of evaluating system performance in advance of actual construction are desired. This situation occurred in the development of the No. 5 crossbar system and was the occasion for building the present throwdown machine.

THROWDOWN TECHNIQUES

Since throwdown techniques have played an important part in the development of telephone traffic theory, a brief discussion of the basic principles will be given.

A single throwdown test will indicate the performance of a system under a specific set of conditions. In a typical telephone traffic study, a given traffic load would first be assumed and a simulated system installation to handle this load would be engineered on the basis of the best available information. The test run then will show the performance of the system under these particular conditions, and indicate both the adequacy of the initial engineering procedures and possible improvements. To obtain a proper balance between equipment quantities and traffic load may require several additional runs, varying the equipment quantities, traffic load or both.

The procedure in a throwdown study is to first obtain data representing the traffic to be handled by the system. The traffic data can be generated artificially by the use of random numbers. The method is based on a knowledge of the statistical behavior of the various factors entering into the composition of real traffic. Random numbers are drawn for each factor. These numbers are assigned values according to frequency distributions obtained analytically or from field observations. The regulating data are combined to produce a description of a sample of traffic which would be encountered under the assumed conditions and then processed by methods which simulate the performance of the actual system.

As a simple example of the throwdown procedures, suppose that it is desired to determine how often on the average an "all trunks busy" condition will occur in a particular group of trunks handling inter-office calls. A certain period of time is first selected and the number of calls expected within this period is determined. Two random numbers are then drawn for each call. One random number specifies the time, within the period, of origination of the call. The other random number, weighted according to an exponention distribution which will be discussed later, gives the holding time of the call.

With the data of call origination times and holding times prepared, the throwdown run can be started. The calls are listed in the order of their originating times. The first call is assigned to the first idle trunk. A record that this trunk is busy is made and the time at which it will become idle determined by adding the assigned holding time to the time of origination. This is also recorded. The call which follows in time of origination is then assigned to the next idle trunk and the process continued for succeeding calls. Before each call is established the release times of all busy trunks are scanned to determine whether any busy trunk should be made idle. In setting up each call idle trunks are chosen from the group in the same order of preference that would be used in the system being simulated.

Thus, the performance of an actual system is reproduced with considerable accuracy and detailed records of this performance can be made. From a study of these records the desired information can be determined. The probability of encountering an "all trunks busy" condition can be found, the average number of trunks busy can be determined, or a frequency distribution chart showing the percentage of the time the number of busy trunks is above any given number can be constructed. If proper records are kept such information as the average number of trunks searched over in locating an idle trunk can be determined. If the trunks were reached through a graded multiple or if they were in subgroups with a common overflow group, simple extensions of the above procedures would be followed.

This particular problem can be solved by analytical methods and is presented here only to illustrate the application of throwdown techniques. However, as problems become more complex it becomes increasingly difficult to apply analytical methods. Even in relatively simple systems the interplay of variables may be so involved that existing theoretical methods are entirely inadequate. Various simplifying assumptions must be made and there is often the doubt that some important factor has not been overlooked in formulating the mathematical theory.

The most fruitful use of throwdown methods has been to check results obtained by theory and to obtain data upon which mathematical theories can be based. Throwdown techniques, of course, can be used to obtain direct results, but the functioning of a system will be better understood if there is at least some attempt to develop a theory which explains how various forces act together to produce observed results. Such theories may suggest modifications of the system which will improve its performance.

It can be seen that the planning of a throwdown study requires a thorough knowledge of both the functioning of the system being simulated and the characteristics of the input data being processed by this system. The validity of results will depend upon the faithfulness with which the artificially prepared input data represent real data and the accuracy with which the throwdown routines represent the real system performance. The following two sections of this paper will describe the No. 5 crossbar system and the characteristics of the subscribers using this system. Later sections describe the methods used in the machine for simulating the dynamic performance of an operating system with its subscribers.

THE NUMBER 5 CROSSBAR SWITCHING SYSTEM*

No. 5 crossbar is a marker-controlled system designed primarily for local central office application in the residential sections of large cities and the fringe areas surrounding these cities.

In regions of this type a relatively large proportion of all calls are completed to subscribers within the same office. Since the surrounding offices to which connection must be made are likely to be of widely diversified types, the system is designed to interconnect with any existing type of central office. No. 5 is also capable of serving isolated centers from about 3,000 lines up, and multioffice areas including the largest

* F. A. Korn and James S. Ferguson, *Trans. A.I.E.E.*, **69**, Part 1, pp. 244–254, 1950.

metropolitan business exchanges. Although the system includes facilities for toll and tandem switching, these were not included in the throwdown studies and will not be discussed.

The Switching Network

The No. 5 system is built around an interconnecting network consisting of two types of switching frames utilizing crossbar switches and known as line link frames and trunk link frames. This is illustrated in block diagram form on Fig. 2. Each frame is double-ended and provides means for connecting any point on one side of the frame to any point on the other side. The connecting paths are known as links.

All subscriber lines in the office connect to one side of the line link frames, each of which can serve, roughly, 300 to 500 lines; and all trunks connect to one side of the trunk link frames, each of which has 160 trunk appearances. The other sides of line and trunk link frames are connected together in such a manner that each line link frame has access to all trunk link frames over several paths.

These interconnecting paths are known as junctors. The basic maximum number of line link and trunk link frames is 20 and 10 respectively, and this is the size embodied in the throwdown machine. However, in practice, multipling arrangements can be employed to double the number of frames to give greater subscriber and traffic capacity.

With the system described above, any subscriber line can be connected to any trunk over one of several paths, each consisting of two links and a junctor and known as a channel. On connections to outgoing or incoming trunks, a single channel is required; on connections through intraoffice trunks (connection between two local subscribers), two channels, one from each end of the trunk, are required. The method of combining links and junctors to form channels will be described later.

Dial Registration—Originating Registers

The circuits that receive and store the dialed signals from the subscriber are known as originating registers. These circuits, in quantity as determined by desired quality or grade of service, are distributed over the trunk link frames as equally as possible. A connection between subscriber and register is set up through the switching network just as between subscriber and trunk. The registers call in control equipment for setting up the subscriber's connection when dialing is completed.

Common Control Circuits—Markers

The switching of all connections in the office is performed by a group of common control circuits known as markers, any one of which may be utilized on a particular call. The principal functions of the marker are (1) to determine or receive the specific location of a calling circuit; (2) to translate input signals into the specific location of a called circuit or group of circuits; (3) to test for availability and seize a called circuit or one of a group of circuits; (4) to locate, test and seize a switching path between calling and called circuits; (5) to set up the connection; and (6) to take alternative action in case of trouble or busy conditions. A marker performs these functions in a very short period of time so that a few circuits can handle the requirements of an office. In the original design of No. 5 crossbar, a single type of marker handled all connections. This was the arrangement specifically handled by the throwdown machine. Later design has introduced three types of markers; dial tone, completing, and combined.

As an example of the function of the control circuits, when a subscriber originates a call, a connection is automatically established from the subscriber line circuit on a line link frame via a marker connector to an available marker. The marker identifies the location of the line and establishes the fact that it is a new call requiring a register. It tests all registers and trunk link frames and chooses an idle frame with idle registers. The marker then gains access to the correct line link and trunk link frames via the frame connectors, chooses an idle register, tests all usable channels, picks a particular channel and operates the crossbar switch magnets to close the connection between line and register. After storing the line location in the register for later use, the marker disconnects itself.

When the subscriber completes dialing, the register connects itself to an idle marker via the marker connector. It transfers to the marker the location of the originating line and the called number. If the call is local to the office, the marker determines the location of the called line from the number group circuit (a translating device) and tests and chooses an intraoffice trunk. The marker then gains access to the link frames through the frame connector, tests the called line for busy, picks a channel, and establishes the connection, thereafter removing itself and the register from the connection.

During the course of the foregoing events, a call may encounter various delays beyond the minimum circuit operating time in setting up the connection. Delay may be caused by meeting a temporary busy condition of markers, registers or the particular number group link frames required. Busy condition of lines or trunks may result in rerouting the call. In general, the grade of service is measured by the delay in connecting a subscriber to a register (dial tone delay) and the probability of not finding an idle trunk or channel.

Intraoffice Trunks

The intraoffice trunks, used on locally originated and completed subscriber connections, include the supervisory, charging and ringing functions. The trunk is held on a connection for the duration of the call in distinction to markers and registers which have short holding times.

Outgoing Trunks and Senders

Calls to other offices are completed via outgoing trunks which incorporate the supervisory and charging functions. In order to transmit information to the distant office, an outgoing sender is connected to the trunk for a short period of time by means of an outgoing sender link. In establishing the call, the marker first connects itself to the sender via the sender connector in order to register in the sender the called number, and then sets up the trunk-sender linkage. When the sender, which may be one of several varieties depending upon the type of signaling required by the distant office, has transmitted the number to the distant office, it drops off the connection.

Incoming Trunks and Registers

Calls from a distant office are completed through an incoming trunk, which includes supervisory and ringing functions. In order to receive signals from the distant office, the trunk connects itself to an incoming register by means of an incoming register link. When the register, which again may be one of several varieties, has received the called number, it obtains a marker through the marker connector for setting up the connection. When its functions have been accomplished, the register disconnects.

Tone Trunks

This group of trunks includes those to which calls are routed when line busy or all trunk or channel busy conditions are met. Subscriber error or excessive delay in dialing also result in routing to these trunks. The tone trunks return distinctive tone signals to the subscriber. If the marker is unable to set up a connection to a tone trunk, the originating register is able to return the tone signal.

Number Groups

Each number group is a translating circuit which provides terminating information for a consecutive block of 1000 directory numbers. When a marker transmits the called line number to a number group, it receives back the line location on a line link frame, the type of ringing required, whether or not the line is in a PBX group, and certain minor information. The number group also includes facilities for selecting an idle line in a PBX group. As with link frames, only one marker can connect to a number group at a time, but markers can connect to different number groups simultaneously without interference.

Connectors

The marker connector, providing access from line link frames and register to markers, include circuits which assign preference of all line link frames and registers for specific markers. This helps distribute marker usage.

In addition, when several line link frames and registers are competing simultaneously for busy markers, a gating arrangement allocates the order of service to reduce excessive delays. Although shown as one block on Fig. 2, the marker connector circuits are provided one per line link frame and one per sub-group of ten or less registers.

The frame connector, which provides access from markers to link frames, includes preference and lockout features since only one of several competing markers can connect to a given frame at one time. Although shown as one circuit on Fig. 2, the frame connectors are actually distributed over the frames.

Outgoing senders of a particular type are divided into two groups and the sender connector permits connection from two markers, each to one sender in each group simultaneously. Preference and lockout features obtain.

The number group connectors, provided one per number group, are similar to frame connectors and give access from markers to number group.

Handling a Typical Call

An understanding of the operational intricacies of a telephone switching system cannot be gained by a discussion of components and can only be developed by a study of the course of events in setting up calls through the system. As has been intimated earlier, the establishment of calls is largely a matter of marker operation. In order to illustrate the marker functions, Figs. 3 and 4 show two charts indicating the order of events in establishing a connection, first, between a calling or originating subscriber line and a register, and second, after dialing, between two subscriber lines within the same office. These two types of connection are known as a dial tone call and an intraoffice call, respectively.

Figs. 3 and 4 are drawn as sequence charts with time flowing downward. There is no attempt to maintain an accurate time scale; the x marks on the vertical line merely represent the relative order in which important control functions take place. In actuality, of course, the time between x marks is known with fair precision. Brief descriptions of the control functions are listed to the left of the vertical lines. The call illustrated is presumed to encounter no difficulties in completion. However, points at which blocking might occur are marked with an asterisk to the right of the lines. If any of the difficulties noted were to develop, the marker would have to take alternative action which will be illustrated later. Also shown to the right of the lines are potential points of delay, where a call may have to wait until a connector, a marker, or a desired frame becomes idle. It must be remembered that during moderate or heavy traffic, several or all of the markers are working simultaneously and tending to interfere with each other.

In a well-balanced and soundly engineered central office, the aggregation of parts are nicely adjusted to give on the average no more than certain preassigned values of delay and blocking at some average busy hour traffic level chosen as a base. A typical example of permissible delay is no more than 1 per cent of calls having a dial tone delay greater than three seconds. When traffic is heavier than the engineering base, the percentages of delay or blocking will increase.

A summation of all the possible alternative sequences which a marker may have to take when trunk busy, line busy and channel busy conditions are encountered becomes extremely complex. Although no attempt will be made to discuss this in detail, a chart showing the operational variations of a marker on an intraoffice call is presented on Fig. 5. Even this figure does not include all possible variations since, for example, the contingency of all tone trunks being busy is not shown on the diagram. This chart, similar in form to Fig. 4, will later be found useful in discussing the throwdown machine. In order to simplify the presentation, some of the control events are combined in time with the frame seizure which precedes the event. The normal course of a call

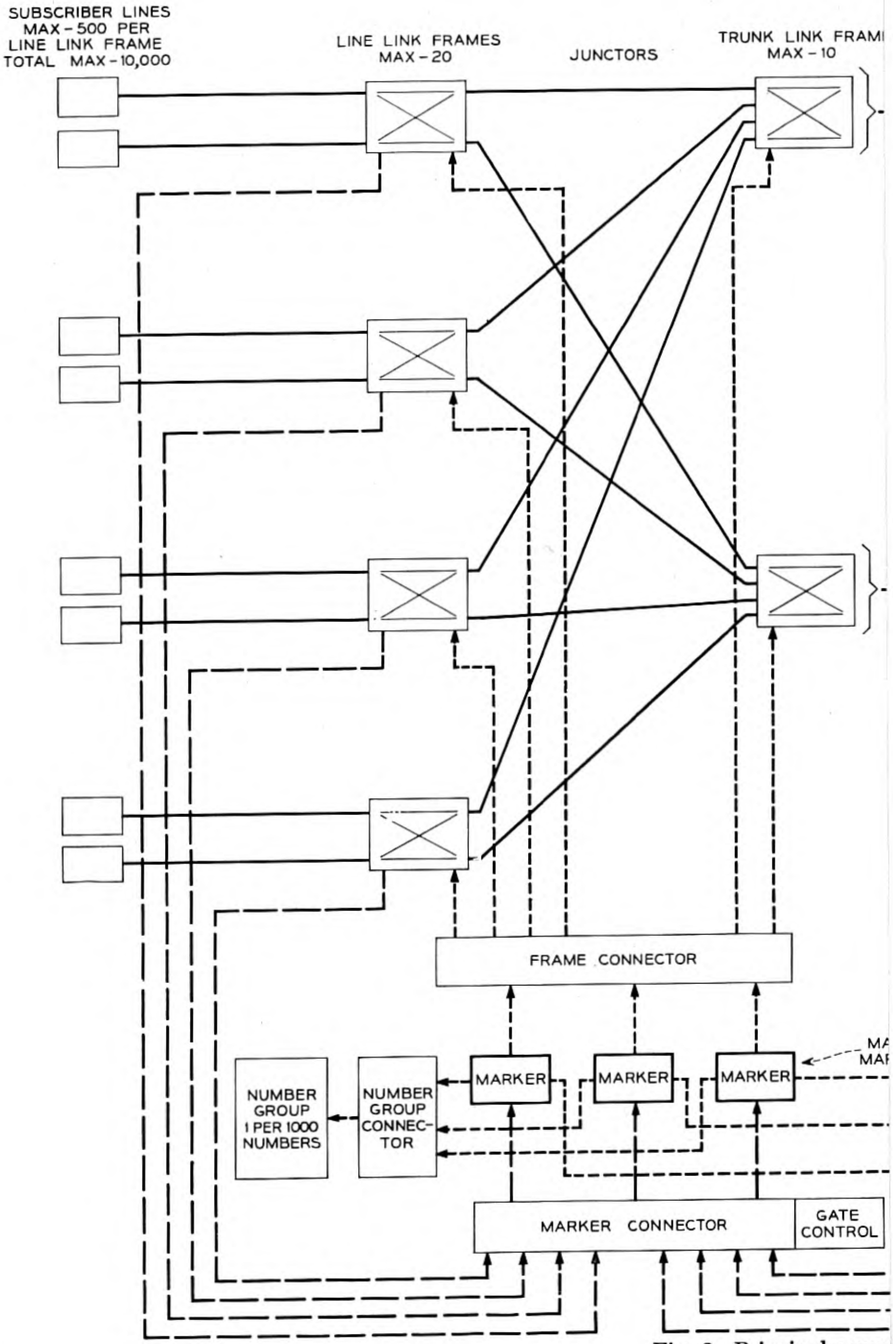
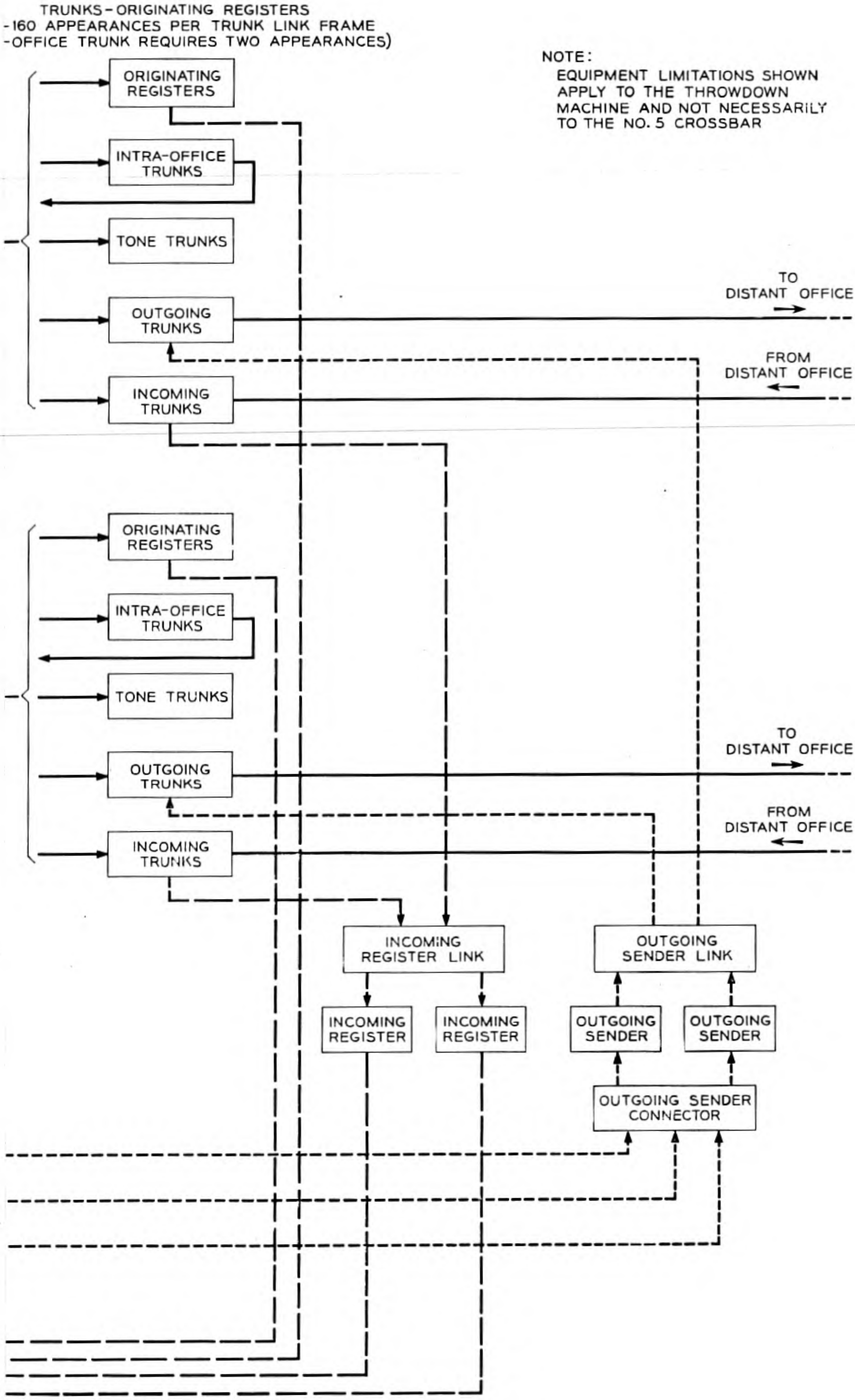


Fig. 2—Principal comp

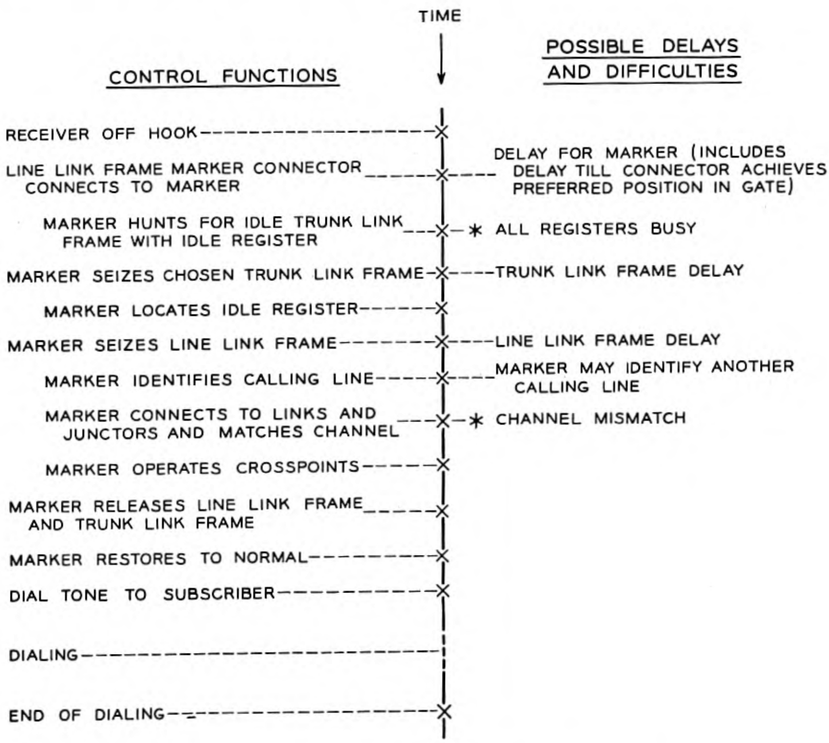


5 crossbar office.

is in a vertical direction and it is only when a busy condition or a special function such as PBX hunting is encountered that the call shifts to the right or left. On a particular call, the only control functions performed are those marked with an x.

In the situation illustrated, the size of the office is assumed to be such that two groups of intraoffice trunks are provided. The marker makes a more or less random choice of the first group to be tested, and, if this group proves to be all busy, will automatically test the second group.

This brief description of the Number 5 Crossbar System is not intended to be exhaustive. It discusses only the important features of the system and those which will help make the description of the throwdown machine more intelligible. Certain more involved items will be discussed in greater detail in later sections concerned with the functioning of the throwdown machine.



* MARKER MUST TAKE ALTERNATIVE ACTION IF IT ENCOUNTERS THIS CONDITION

Fig. 3—Establishing a dial tone call.

CHARACTERISTICS OF SUBSCRIBERS AFFECTING DATA

Since the purpose of the throwdown machine is to evaluate performance and traffic capacity of a simulated switching system under conditions met in service, subscriber data fed into the machine must represent, as nearly as possible, the characteristic actions of real subscribers. Fortunately, telephone subscriber characteristics can be studied in working switching systems which are similar to the one under throwdown evaluation. Little error is introduced by such subscriber data

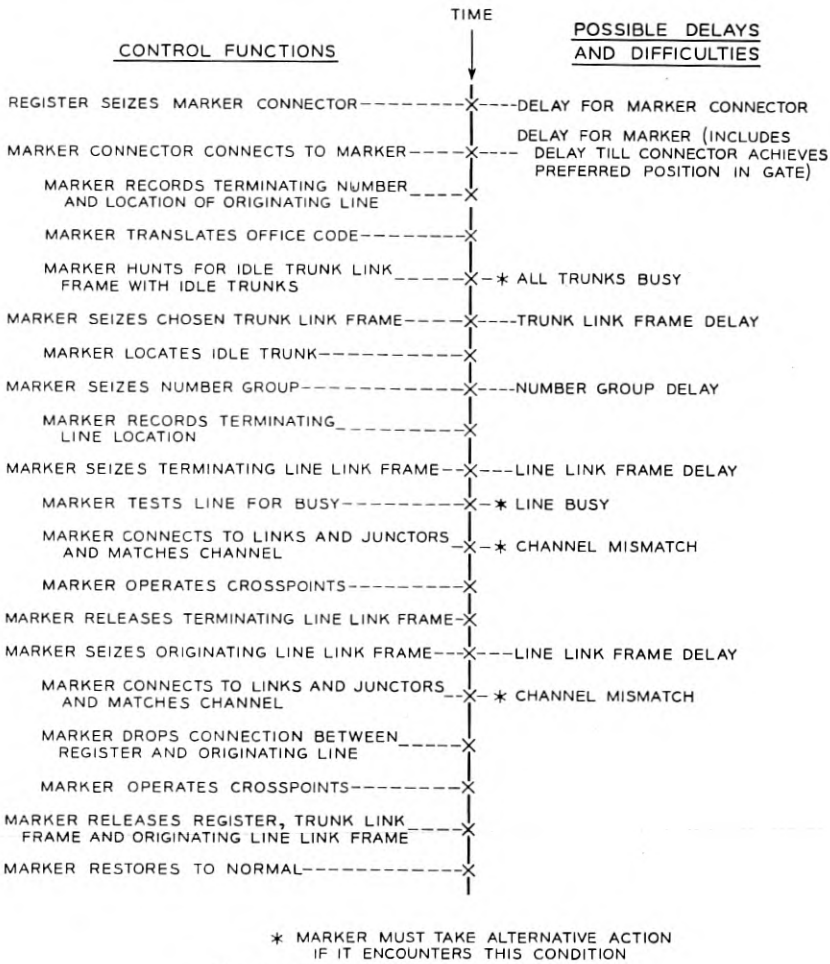


Fig. 4—Establishing an intraoffice call.

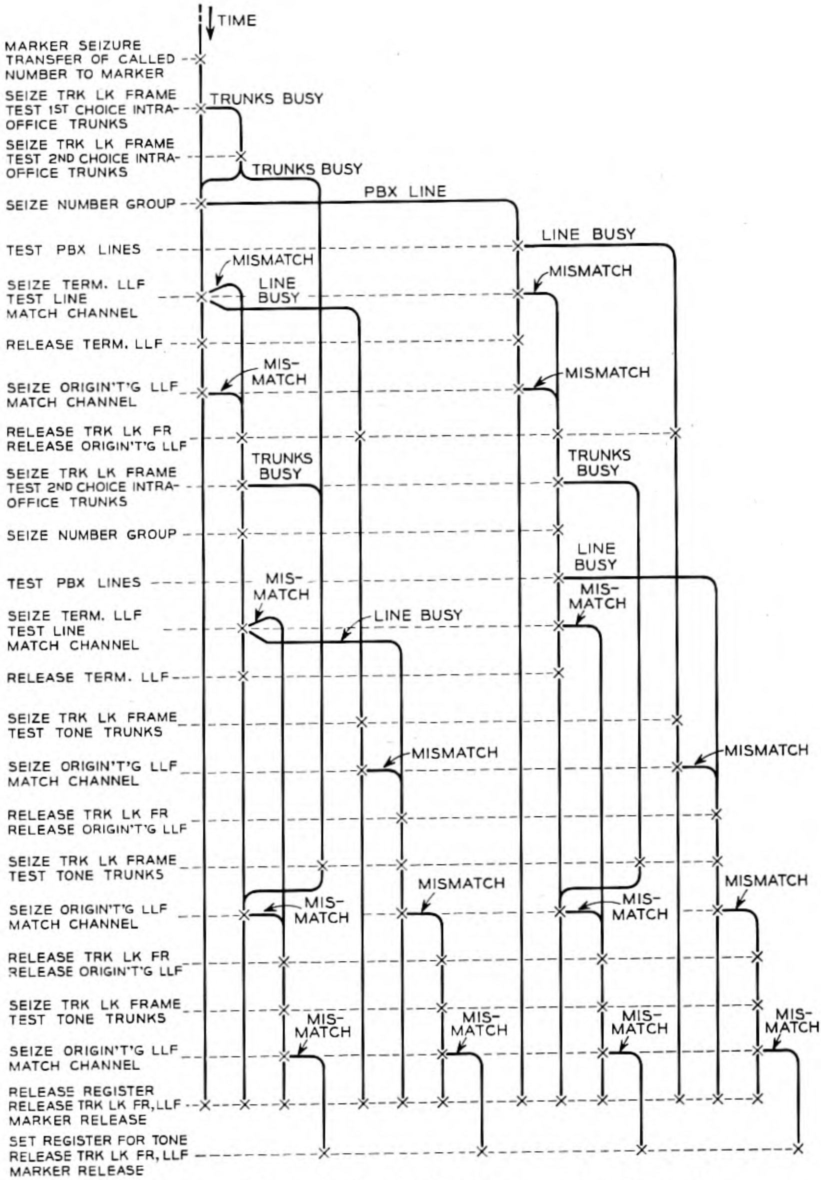


Fig. 5—Possible variations in handling an intraoffice call.

inasmuch as subscriber behavior is very slightly influenced by the type of system serving their telephones.

Subscribers, although they are individuals, exhibit many "group characteristics" dictated not by the requirements of telephone communication but by their mode of life. This fact allows statistical treatment of many observed action distributions without introduction of significant error. However, these group actions also present problems of congestion in telephone plant which require detailed throwdown study for solution.

As an example of group characteristic, subscribers do not originate a steady barrage of calls over the twenty-four hours of the day. During mid-morning and mid-afternoon hours traffic is built to a peak value, whereas during certain of the remaining hours it is reduced to a minimum. In some residential areas peak traffic may also occur during the early evening. Throwdown evaluations of simulated switching systems, however are primarily concerned with the busy hour, the hour in which the greatest number of calls are originated, regardless of its actual time of day occurrence.

Useful datum obtained from busy hour field observations is the calling rate per subscriber (calls per hour) which can be used to set up traffic load conditions on the simulated switching system. The calling rate characteristics can be measured as average calls per hour placed by subscribers in a number of group classifications. An example used in a particular throwdown study of simulated system response to a given traffic load is given in Table I. The values given in this table represent average day to day calling rates. Weather conditions, pre-holiday periods or special events have been found to raise substantially the average calling rate in affected classifications. Values adjusted for these conditions are useful in projecting percentage of overload that can be offered to systems engineered for average daily loads.

Subscribers, however, in originating calls, act independently within their classified group in maintaining the average calling rate. Originating times of calls, therefore, occur at random within the hour. Throwdown input data representing subscriber originating time behavior are produced by assigning to each call, of the total within the studied hour, a six digit number from a list of random numbers. If the hour is divided into one million parts the assigned random number determines the millionth part of the hour in which each call will originate.

Observations made in the field have shown that subscribers, upon receiving dial tone, do not always follow through to dial a full code. Among possible causes are failure to hang up after completion of a call, answering the wrong telephone where two or more are adjacent, dialing

TABLE I

Group Classification	Average Calling Rate Per Subscriber During Busy Hour
Heavy demand individual line subscribers (such as doctor and professional services	5.0
Medium demand individual line subscribers (such as small business and some residential subscribers)	0.85
Light demand individual line subscribers (mainly residential subscribers)	0.02
P.B.X. line subscribers (such as large businesses, hotels, railroads, etc.)	3.0

before dial tone, and forgetting the number. Such actions produce waste usage of equipment within the switching system, and their study is pertinent to producing throwdown data. To simplify this study, all subscriber actions involving the alerting of central office equipment are designated "subscriber starts" and divided into four categories as given in Table II.

Since "no dials" and "partial" dials are largely due to subscriber errors in properly originating calls, many of these calls will be originated upon discovery of the error. False starts, on the other hand, are attributed to accidental origination with no intent to place a call. A flow chart illustrating these actions is shown in Fig. 6. The importance of this subscriber behavior is indicated by the percentage of waste usage calls (FS, ND, and PD) to ultimate good calls. Pen recorder tapes taken at particular central offices showed waste usage calls at 30% and good calls at 80.5% of ultimate good calls. For these specific cases false starts represented 7.5%, partial dials 7.5%, and no dials 15% of ultimate good calls.

TABLE II

Category	Description
Good calls	Calls on which the subscriber waits for dial tone and then dials the required full code
False starts	Calls, on which a sender or register is seized, but which are abandoned in less than two seconds without dialing
No dials	Calls lasting longer than false starts but on which no dialing occurs. These calls may exceed a certain length "time-out" period and be given a distinctive tone
Partial dials	Calls on which less than the required full code is dialed. These calls may be held beyond a certain length "time-out" period and be given a tone

It was also found that approximately 90% of the partial dials and the abandoned no dials were reoriginated. These percentages are quoted only to illustrate subscriber behavior under certain conditions of central office load at a particular office. Type of service, load conditions on the central office, and location can effect these percentages. For a more detailed analysis, see "Dialing Habits of Telephone Customers."*

Fig. 6 illustrates only the group behavior of subscribers. Individually, the subscribers will hold equipment on abandoned no dials, abandoned partial dials, and false starts for varying amounts of time. These varying individual holding times can be quantized into several average values which are equally likely to occur, or may be averaged to one value as shown on Fig. 6 depending upon the throwdown study requirements. The holding times on calls receiving tone are usually assumed to cease a few seconds after tone is received.

Subscribers, as individuals placing ultimate good calls, spend varying amounts of time, after receipt of dial tone, before start of dialing and

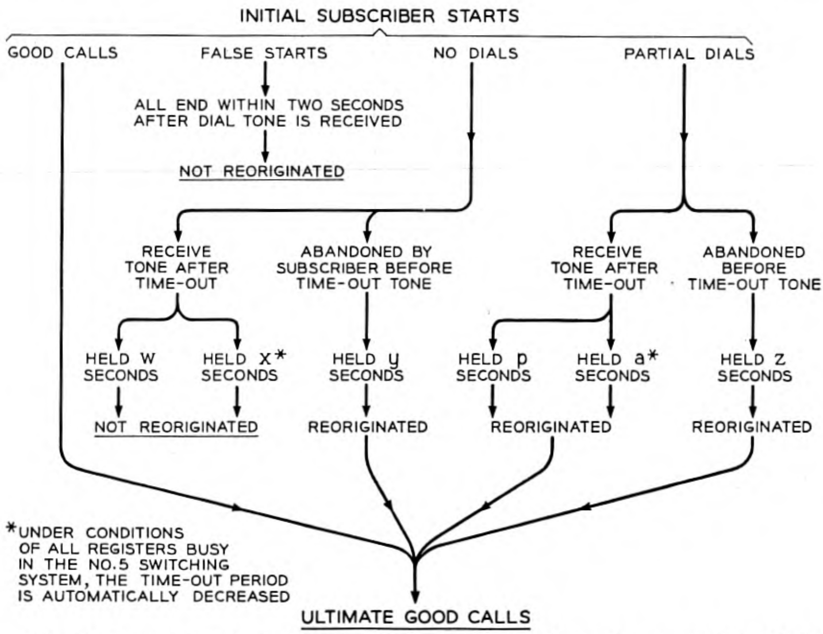


Fig. 6—Simplified characteristic action of subscribers in converting initial subscriber starts to ultimate good calls.

* Charles Clos and Roger I. Wilkinson, "Dialing Habits of Telephone Customers," Bell System Tech. J., **31**, pp. 32–67, Jan. 1952.

in dialing a full code. This behavior affects the holding time of registers receiving the dialed digits and must be considered in throwdown studies. Data collected on combined waiting and dialing time characteristics show a frequency and time distribution that can readily be quantized into a number of values, each equally likely to occur. When the number of values for a particular throwdown study are determined, each quantized dialing time is represented by a number. Each ultimate good call is then assigned a number from a random list of the representative numbers to establish the dialing time of the call.

Ultimate good calls will develop one of three terminating conditions attributable to subscriber behavior: 1) DA, called subscriber does not answer, 2) busy tone because of called subscriber line busy, or 3) answer by called subscriber. It is assumed from analysis of "don't answer" studies that, for certain throwdown evaluations of the switching system, approximately 10% of the ultimate good calls meet the DA condition. The number of busy tone terminations, of course, will develop during the throwdown study as a result of the average originating and terminating calling rate per subscriber served by the system.

Most subscribers, upon encountering a line busy condition make subsequent attempts to reach the called line. The number and frequency of attempts made depend upon the individual characteristics. A detailed analysis of this characteristic, suitable for use in throwdown studies, has appeared in a paper by Charles Clos.*

When calls are answered by called subscribers, the connections will be held for varying amounts of time. It has been determined from field observations that the frequency distribution of these holding times is closely approximated by an exponential distribution. For throwdown purposes a simplifying assumption can be made that holding time is not a continuous variable but is quantized so that a particular holding time will have one of several values. To determine these values an exponential distribution having the proper average is plotted as shown in Fig. 7. The area under the curve is then divided into the required number of equal parts (ten, for this example). A central value of holding time is determined to represent each subarea. Ten holding times are thus produced which are weighted according to the exponential distribution and which are equally likely to occur. These holding times can be designated 0 to 9 and assigned to the calls by choosing one-digit numbers at random for each call.

* Charles Clos, "An Aspect of the Dialing Behavior of Subscribers and Its Effect on the Trunk Plant," Bell System Tech. J., **27**, pp. 424–445, July, 1948.

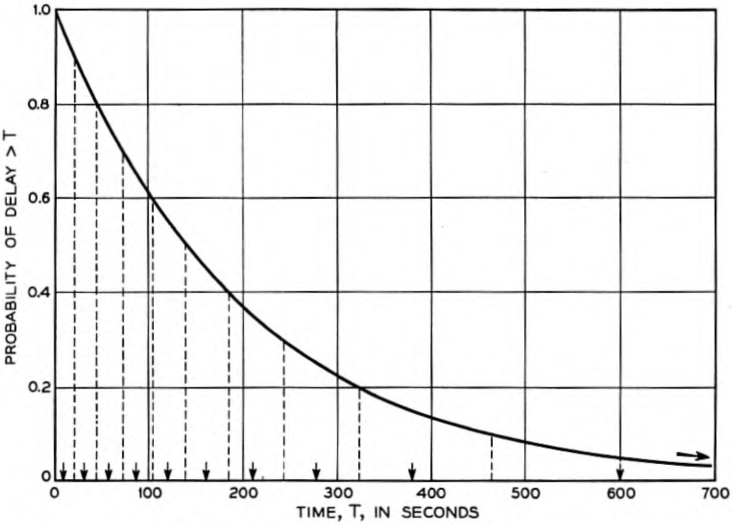


Fig. 7—Distribution of holding times.

GENERAL PLAN OF THE MACHINE

The broad plan of the throwdown involves a division of work between the team of operators and the circuits of the machine. The circuits keep track of system events, resolving complex sequences of actions concurrently taking place. Their chief purpose is to present signals to the four operators so that they will be able to perform the right actions at the proper time and thus dovetail together the events for a large number of calls in progress.

The operators keep manual records of the busy-idle states of items which occur in large quantities, such as lines, trunks and links. They also perform the searching operations necessary to locate these items when they are required to be made busy or idle.

In general the circuits signal the operators to perform actions; the operators in turn signal the circuits that the action is completed or some appropriate alternative taken. The circuits then determine the next action and present corresponding signals to the operators. Thus the circuits largely control the sequencing of events. However, in some cases where the sequencing would require extensive circuitry, the operators assist in determining sequence. For example calls returning to the control circuits after a subscriber completes dialing are interleaved with new calls according to written records maintained by the operators. Releasing connections are also placed in proper time sequence by the operators according to written records.

The actions of the operators are checked electrically in many cases. An improper setting of certain switches, which is inconsistent with the state of the system at the time will in some cases give an alarm and in others, block the progress of the machine until the error is corrected. Certain actions performed by the operators involve the insertion of plugs into jacks. Where alternative actions are possible in response to a given signal, separate groups of jacks correspond to the several alternatives. Insertion of a plug into a particular group will signal the circuits as to which alternative is taken. The circuits make one of several keys effective and the operator must then depress this key (corresponding to the action she has taken) before the circuits will advance.

Where several operators must cooperate to perform a given action, signals are passed between the operators by means of keys and all operators must respond before the circuits can advance. Wherever possible, overlap operation is employed. Signals are presented to several operators simultaneously and each operator starts the indicated action, signaling the circuits when it is completed. When the signal from the last operator is received, the circuit advances. In some cases an operator is allowed to start a particular action before the stage has been reached at which this action is required. For example the information necessary to choose links and determine a suitable path through the interconnecting network of the No. 5 crossbar system is available before it is necessary to establish the connection. Since this search is time consuming, the operator is allowed to start as soon as the information is available. At the proper time she is given a signal to complete the record of this connection, or if the call has been blocked before reaching this stage, she is given a "back out" signal instructing her to restore her records to their previous condition. Since this is a rare condition occurring less than one time in 100 tries, little useless work is done and the overall action is speeded up.

A block diagram showing the relations between the operators and the various major components of the machine is shown in Fig. 8 The CLOCK controls the flow of time in the machine and gives an indication of simulated present time in the traffic run being tested. The TIME DETECTORS, of which three are used, provide a means for the operators to indicate a future time at which some action must be taken and be signaled when this time arrives. The block labeled CONTROL CIRCUITS, which present action signals to the operators, represents the main body of circuits which maintain a current record of the states of the various complex units of the system, such as markers. The GATE provides a means for operators to feed traffic into the machine and determines the order in which each working item of traffic is taken up by the control circuits. The RANDOM CIRCUIT is an electronic "roulette wheel" which permits the operators to make random choices in disposing of certain traffic items. It provides random selections varying from one out of three to one out of ten.

The individual record of each call is made on a card of the form shown in Fig. 9 which is called a "call slip." These are of various types identified by distinctive designations and colors for the several varieties of calls which may occur. The basic types are: dial tone, intra-office, incoming, and outgoing.

As the call progresses through its various phases the slip is passed from one operator to another, each operator retaining the slip while it is in the phase under her control and recording on it in designated spaces the nature and time of the events taking place. When the call is completed, the slip carries a complete record of the call including the designation number and time of seizure of the various circuit units used in establishing the connection and the nature and duration of any delays encountered.

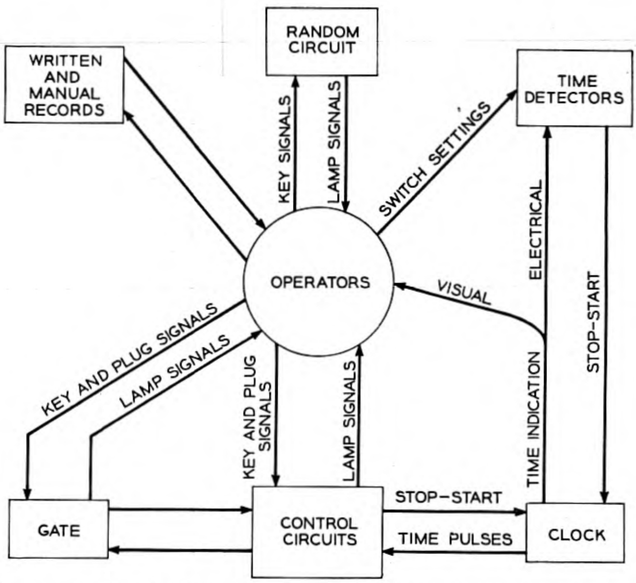


Fig. 8—Communication between the machine components and the operators.

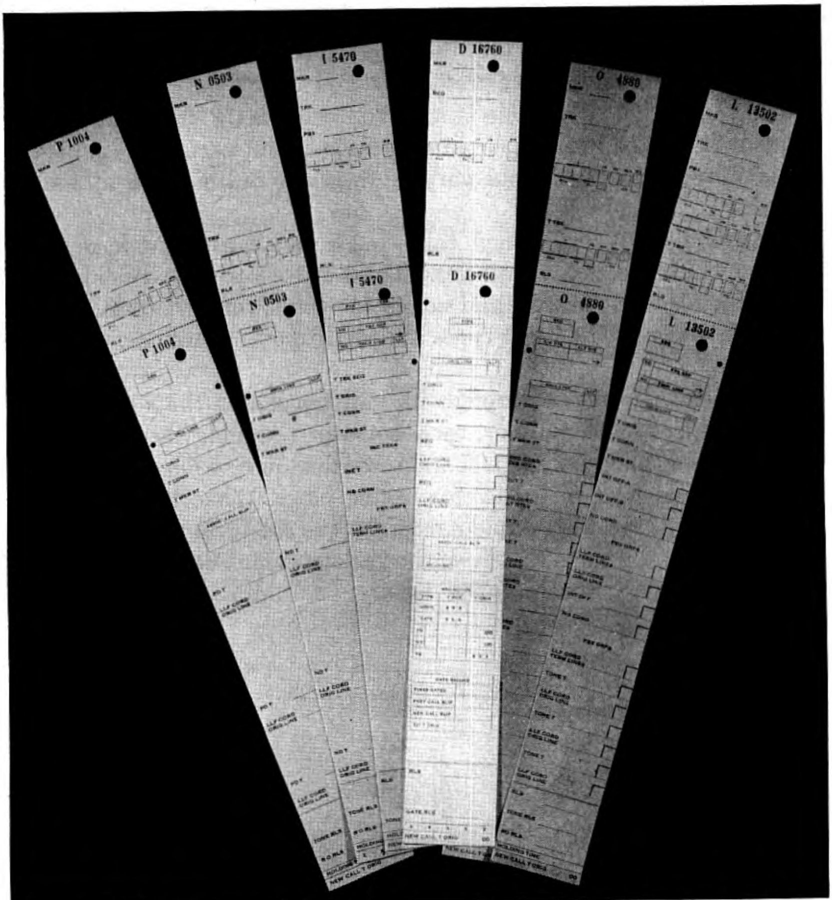


Fig. 9—Call slips on which records are made in the throwdown machine.

DESCRIPTION OF POSITIONS

The throwdown machine consists of five positions as shown in the plan view of Fig. 10. A photograph showing all but one position appears on Fig. 1. This division into positions results from the requirements of simulating the components of the No. 5 crossbar system and equalizing the work load on the attending operators. The arrangement of positions, as shown, provides a continuous clockwise flow of records and other items that must be passed from operator to operator. The five positions are known as: originating position, gate position, marker position, match position, and assignment position. Four operators attend the five posi-

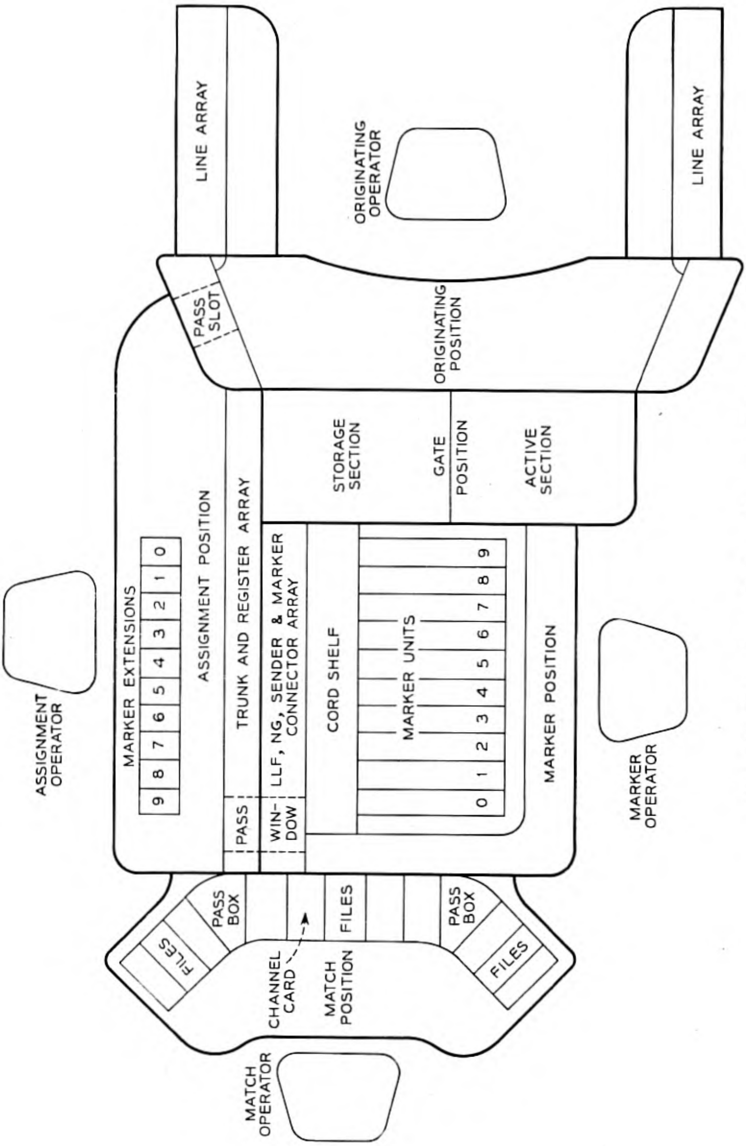


Fig. 10—Plan view of throwdown positions.

tions, the gate position being jointly served by the operators stationed at the originating and marker positions. Each operator is given the same name as the position at which she is stationed.

The general plan of traffic flow is such that all calls are originated or restarted at the originating position. For this reason, the major equipment items of this position include: the subscriber line array, which represents by pegs all of the subscriber lines associated with the switching system under throwdown evaluation; the trunks incoming to this system; and certain of the system components associated with incoming calls. The call slips, sorted in sequence as to originating times, are stored at the originating position for use in originating calls.

Means are provided for presetting the times at which the next originated or restarted calls will enter the throwdown machine. When actual time coincides with a time thus preset, the system stops and signals the originating operator to serve the waiting item. The originating position does not provide for actually entering an item into the machine, but indicates the time of entry, stops the machine, and supplies the items to be entered. For example, the operator, when signalled to originate a dial tone call in accordance with information furnished on the call slip, selects, from the line array, a peg representing the subscriber line, associates it with the call slip, and passes the two items to the gate for entry.

Since the means for determining the busy or idle condition of a subscriber line are provided by the presence or absence of that line's peg in the array, the originating position also enters into the operation on incoming and intraoffice calls.

The gate position, which simulates the marker connectors of the No. 5 crossbar system, serves as the entry point for all calls. The originating operator inserts call slips and pegs into the gate position from one side to start the call. The marker operator removes them at the other side for processing in the marker position. Relay circuits associated with the gate position control the flow of traffic through the gate in accordance with actual No. 5 crossbar operation.

The marker position provides means for associating the call slips with the individual simulated markers of the switching system. Since these markers control processing of the calls, the principle records of the calls' progress are obtainable through this association. The records are kept as time entries on the call slip and marked adjacent to action lamps determining these entries.

At the top of each marker unit in the marker position are cords which provide access to the switching system components under control of the marker. These components are line link frames, number groups, sender subgroups, and marker connectors. As the call progresses, the marker operator will connect and disconnect the marker with these components as directed by the action lamps provided.

The principle purpose of the assignment position is to provide equipment for test and choice of a trunk on each call. A trunk jack array which includes all trunks and registers of the switching system therefore appears on the face of this position. Since, in the actual No. 5 crossbar system, testing and choosing of trunks are performed by the markers, an extension of each marker also appears in the assignment position. The assignment operator, when required, makes a visual test for idle trunks to the proper destination. She then chooses and associates one of these idle trunks with the active marker by means of a marker cord which is plugged into the trunk jack.

In addition to this principal function of trunk choice, the assignment position is equipped to determine the disposition of calls when the marker has finished setting them up. Means for ascertaining conversation time, dialing time and other types of holding times is provided.

The positions so far described have simulated only the two ends of a connection, the subscriber line and the trunk or register. To complete the connection a channel must be set up between these ends. In the actual switching system the marker matches a line link, a junctor, and a trunk link to form the channel. The match position is provided to simulate this action. The match operator, through visual inspection of a set of channel cards, tests and makes busy the channels for each connection. Information as to the originating subscriber line reaches her through the pass box in the form of the upper portion of the call slip, Fig. 9. Information as to the trunk or register choice is passed verbally by the assignment operator. Since this operation is a function of the marker, marker extensions appear at the match position. These extensions are provided with action lamps to guide the match operator and with keys to inform the simulated marker circuits of the results of the match.

A pass window is cut between the marker position and the assignment position for passage of the call slip and originating peg at the time of marker release. Similarly, a pass box is provided for passing the upper portion of the call slip (match ticket) from the match operator to the assignment operator. The assignment operator is charged with the disposition of these items.

THE TIME SYSTEM

The timing system of the throwdown machine is based on a start-stop system of time pulses. Pulses generated by the clock, Fig. 11, drive indicators which display time and also drive circuit elements which count time units to cause events to occur in the proper sequence. When these time counters reach a stage where some action is to be performed by the operators, "stop-time" signals are produced. These signals lock in and cause the clock to halt. Simultaneously, action signals are displayed to the operators. When each operator completes the indicated action she depresses an appropriate key at her position which extinguishes the action signals and removes the stop-time signal. When the last operator has responded, all stop-time signals are removed, the time lock is opened and the clock again advances. Thus the clock, in effect, takes time-out while the operators perform the various manual searching and recording actions necessary to simulate and tabulate the performance of the crossbar system.

In the throwdown machine, each clock pulse represents one millionth of an hour or 3.6 milliseconds. This quantitization of time is based on two considerations. The first is that it is convenient to represent a particular time during an hour by a six-digit decimal number. The second is that the time represented by one time unit (3.6 milliseconds) is well under the average acting time of the relays and switches used in the system being simulated. Thus the events taking place in the actual system can be reproduced in sufficient detail. Some events of longer time duration

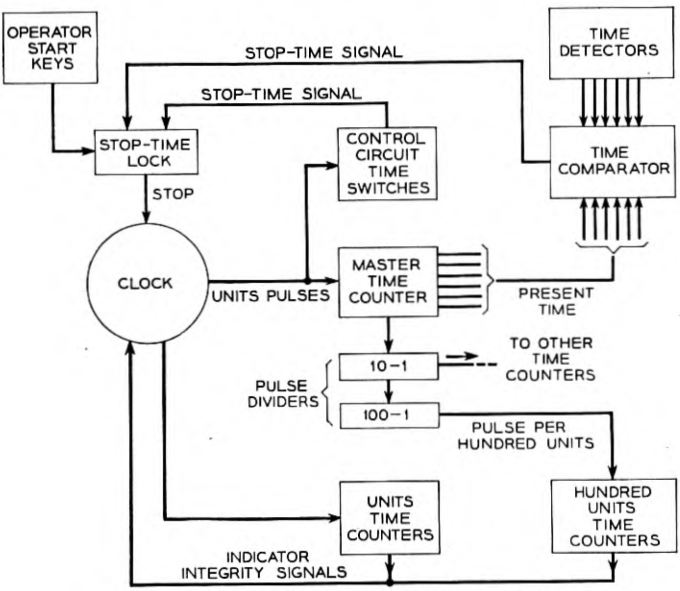


Fig. 11—Block diagram of the time system.

are timed in less detail. For this purpose the clock is also arranged to deliver pulses at one tenth and at one hundredth of the basic pulse rate.

The clock, circuit wise, is a form of free running relay pulse generator. It consists of a series of relays in which the first, in a released condition, causes the remainder of the series to operate in sequence. When the last relay of the series operates, it causes the first relay to operate. The remainder then release in sequence. When the last relay releases, it causes the first relay to release. This cycle, if not interfered with, is repeated continuously to produce pulses representing units of time. Time, thus, can be stopped by the simple expedient of allowing the stop time signals to hold or "lock" the first relay in the operated state.

Time is visually indicated in units and hundreds of units at the operators' positions by a group of telephone type message registers termed the *time counters*. Certain of the registers indicate present machine time for action recording purposes. Others are set at a specified number of units ahead of present time to indicate future times at which held items will be released or re-entered into the system:

To drive these counters and to safeguard their integrity, the counters are substituted for the last relay in the clock pulser relay series. The operating windings of all *units counters* are connected in parallel. Contacts, which make on each counter when the individual counter is advanced, are all connected in a series circuit to form the last relay contact. Failure of any units indicator to advance will, therefore, interrupt the pulse generator cycle and stop time until the trouble is cleared.

The integrity of the hundred units time counters is guarded in the same manner. On each hundred pulse when these counters are advanced, their windings and contacts also form a part of the pulser circuit.

The basic pulse repetition rate of the clock is approximately four pulses per second, being determined by the acting time of the counters and the various circuit elements which the clock pulses must drive. Since each pulse represents 0.0036 seconds, the ratio of basic machine time to real time is in the order of 70 to 1.

The clock pulses, Fig. 11, are counted by two types of time switches. One type, the control circuit time switch, is associated directly with the component control circuits of the throwdown machine. These time switches are not continuously driven but are automatically connected to the clock as required to time events in the progress of a call. The control circuit time switches, as discussed in more detail later, are returned to zero after each usage in preparation for timing the next event.

The second type of time switch, designated the *master time counter*, is continuously driven by the clock. This counter records simulated present time throughout the entire running of a throwdown test.

The master counter does not, in itself, generate stop time or action signals. Its primary purpose is to furnish correct present time to the time comparator circuit where such signals are generated. It also controls a pulse divider which furnishes pulses at one tenth and at one hundredth the basic pulse rate.

Since failure of the master time counter to advance on each pulse would introduce present time errors which are cumulative over the throwdown run, the integrity of this circuit is rigidly guarded. Checking circuits verify that each basic clock pulse advances the master time counter. The checking circuits, not shown on Fig. 11, upon detecting a failure to advance, produce a stop-time signal and an alarm signal which can be released only after the counter is brought to correct time. The master time counter consists of pulse driven rotary switches arranged so that each switch represents one decimal digit of time. To count one hour of simulated time, as provided in the master time counter, six switches are necessary. These switches record 000000 to 999999 units of 0.0036 seconds.

Means are provided for presetting such items as subscriber originating times, incoming call originating times, and hold release times by the *time detectors*. When a time so set coincides with simulated time, the clock is stopped by a stop-time signal and an action signal indicates that a call is to be originated or a held item released.

The time detectors consist of sets of ten-position manually controlled switches with each switch representing a decimal time digit. Since simulated time is divided into millionths of an hour, these switches are preset to the millionth interval, say 003162, in which an event is to occur. Information from each switch is transmitted to the *time comparator*, Fig. 12. Also transmitted to the time comparator, from the master time counter, is the simulated machine time, say at present, 003159. When the master time counter advances to a time 003162 which coincides with the detector time, the time comparator generates a stop-time signal to stop the clock. An action signal, associated with this particular time counter in coincidence is also lighted. The operator, after taking the appropriate action, resets the time counter to a future time—the time of the next waiting item in the category—and depresses the start time key. Checking circuits prevent advancement of the clock should the time detector accidentally be set to a time value which has already passed in the throwdown run, in the example, to a value less than 003162.

Three time detectors are provided in the throwdown machine. One each is used for setting originating times of subscriber starts and of incoming calls. The third is used for releasing held items. Since the holding times of these items are measured in hundreds of time units the last two digits of the time interval are dropped, and only four switches are required.

It has been mentioned that the ratio of basic machine (clock) time to real time is in the order of 70 to 1. However, in operation, the flow of time is halted frequently to permit actions by the operators. The average interval between stops in the traffic runs which have been processed is less than 10 time units. Thus the machine time is only a small fraction of the time consumed in processing a traffic sample. The ratio of total processing time to real time has turned out in practice to be between 1000 and 2000 to 1 depending on the nature of the traffic sample being tested.

GENERAL PLAN OF THE CONTROL CIRCUITS

The major part of the throwdown machine circuitry is associated with the marker sequence and timing controls and the gate preferencing arrangement. The circuit plan followed in these two cases will be briefly described in order to illustrate how the throwdown functions were implemented.

The gate circuits simulate the action of the marker connector circuits of the No. 5 crossbar system which control the access of line link frames and registers to the markers. These circuits assign traffic to idle markers according to the preference rules used in the actual system. The circuits resemble corresponding circuits of the system. They employ two relays per connector and one relay per marker and are arranged with cross-connection terminals so that the preference order and number of connectors can be varied as required.

Each call handled by a marker consists of a series of events occurring in time sequence with time intervals between events corresponding to the "work time" consumed by the marker in performing required functions. The sequence of events is not fixed at the start of a particular call but may be altered from stage to stage depending on the particular busy and idle conditions encountered. A block diagram of the control circuits used for simulating this action is shown in Fig. 12. They consist of a number of individual circuits provided on the basis of one per marker together with common circuits whose use is shared by all markers.

The fundamental plan is based on the use of two rotary stepping switches. One of these, the sequence switch, determines the sequence in which events take place, while the other, the timer switch, measures work time between events. In the throwdown machine each possible event which may occur in setting up a call is represented by a position on the sequence switch. Circuits through the switch cause action signals for the event represented by its position to be displayed to the operators. All of the events for one type of call are arranged in order on the switch terminals so that by omitting certain positions (events) all of the variations of this type of call can be represented. At the conclusion of an event the switch is directed to the position corresponding to the next appropriate event according to the setting of "memory" relays located in each marker unit. These relays operate at various stages of the call in response to key signals from the operators indicating the conditions they have encountered in attempting to respond to action signals. Thus, significant events are recorded in order to control the future progress of the call.

To provide for all types of calls it was found necessary to furnish three sequence switches for each marker. These switches have 22 positions and six arcs. Five arcs are used for displaying signals and for control while the sixth arc is used in conjunction with the timer switch to control the time at which signals are displayed. A timer switch is individual to each marker. It is a 22-position, six-arc switch arranged with auxiliary relays to count a maximum of 105 clock pulses. Terminals of the timer switch

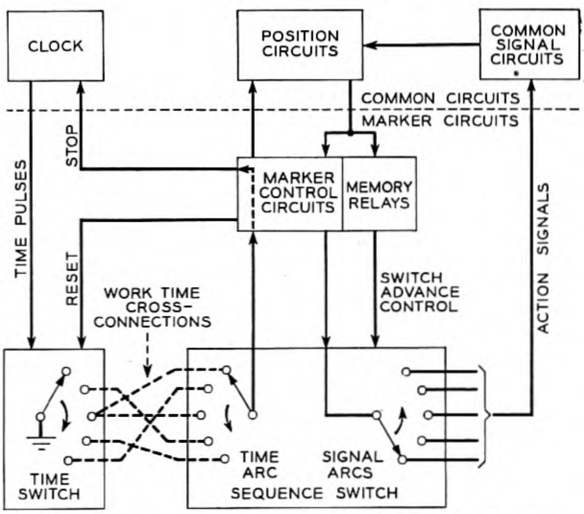


Fig. 12—Block diagram of marker progress circuits.

are cross-connected to the time arc of the sequence switch to fix the work time preceding the event represented by each sequence switch position.

In the general operating scheme, the sequence switch stands at a position representing the next event to take place. The timer switch is started from normal and counts clock pulses until it reaches the terminal cross-connected to the position on which the sequence switch stands. This initiates signals which stop the clock and cause action signals to be displayed. When the operators respond, the sequence switch advances to the terminal for the next event in the call, the time switch returns to normal and the clock restarts. The time switch then counts time units leading to the next event. Since several markers may be in use at the same time in different stages of their calls, two markers can reach an action point during the same clock pulse. A lockout circuit insures that only one marker at a time displays its action signal. At any time that a marker stops the clock, the timer switches of all other markers halt but continue their count when the clock is restarted. Thus, relative time relations are maintained while a true count of time consumed by an operating system is obtained.

The circuit action can be illustrated by a discussion of the events in a dial tone call. This call represents an attempt by a marker to establish a connection from an originating subscriber line to a dial register. The possible sequences of events are diagramed in Fig. 13. As indicated, this class of call employs eight sequence switch positions in addition to the normal position.

The call starts when the gate circuit assigns an idle marker to a call which has been originated at the proper time and placed in the gate. The assigned marker is prepared to process this type of call by the operation of an associated "class" key which selects and advances the sequence switch which carries the events of a dial tone call. Advance of the switch is from normal to Position 1 to control, at the proper time, signals for the first event, namely, seizure of a trunk link frame and selection of a register. The timer switch is set at zero and in a condition to step one terminal at a time in response to clock pulses. Terminal 1 of the sequence switch time arc is cross-connected to a terminal of the timer switch representing the marker work time between the time the marker is first seized and the time it attempts to seize a trunk link frame.

The operator in control of the marker now operates a START key and the clock starts pulsing, each pulse causing the marker timer switch to advance one step. When the specified work time has elapsed the timer switch reaches the terminal connected to Position 1 of the sequence switch. This passes a signal from the timer switch through the sequence switch causing the clock to stop and the signals associated with Position 1 of the sequence switch to be displayed. Signals at the assignment position identify the particular marker and instruct the assignment operator to obtain a trunk link frame and a register. Three conditions may occur: (1) a trunk link frame frame and register may be idle and available, (2) a register may be idle but the frame on which it appears is busy, being held by another marker, or (3) all registers may be busy.

In condition (1) the assignment operator obtains the frame and register according to the operating procedures and operates an OK key at her position. This extinguishes her signals and passes a signal to the marker operator instructing her to record the present time in the space provided for trunk seizure on the call slip. The proper space is indicated by one of a row of lamps beside the call slip. The marker operator then operates the START key for this marker causing the marker timer to return to normal and the sequence switch to advance to Terminal 2. The clock starts if no other marker is waiting to display signals.

In condition (2) the assignment operator observes that there will be a delay in obtaining a trunk link frame. She inserts a connector cord for this marker in a "delay" jack and operates a DELAY key. This extinguishes her signals and passes a signal to the marker operator to record

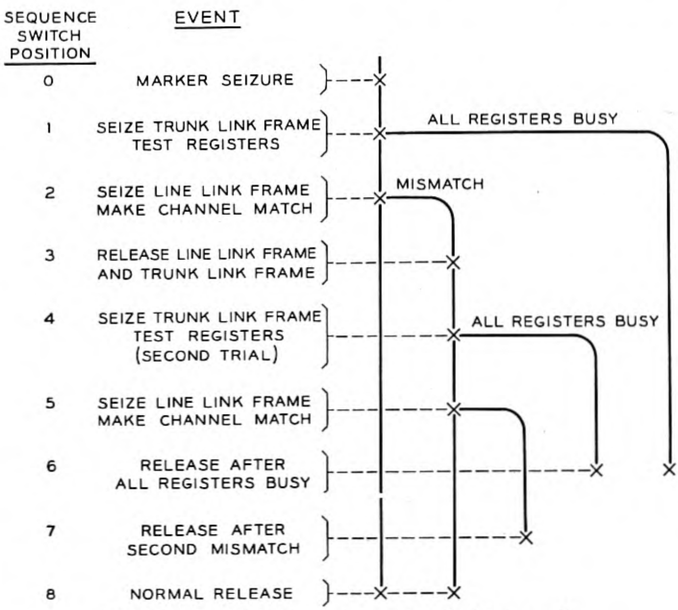


Fig. 13—Possible sequences of a dial tone call.

on the call slip that a delay has been encountered. The marker operator then operates the START key for this marker. This returns the timer to normal but since a delay condition has been established in the marker circuit by the actions of the assignment operator it does not advance the sequence switch but allows it to remain in Position 1. The marker in a delay condition does not permit its timer switch to step but removes the halting signal from the clock to allow time to advance and other markers to be served. All signals for this marker are extinguished during this time. As time advances some other marker will release a trunk link frame. This passes a signal through the delay jack to the delayed marker causing it to display again the signals requesting a trunk link frame and a register associated with sequence switch Position 1. The operators and circuits then proceed exactly as when these signals were first displayed.

In condition (3), the assignment operator observes that all registers are busy and operates her ALL BUSY key. This operates a memory relay in the marker circuit recording that the all busy condition has been encountered and that the future progress of the call should follow the sequence indicated in Fig. 13 by the side branch at Position 1. With all registers busy it is impossible for the marker to establish a connection. As the side branch shows, the alternative is to release and restore to normal. (Later attempts to serve this subscriber will be made until an idle register is obtained.) Thus, with the all busy condition recorded on the memory relays, the circuit will cause the sequence switch to advance, running over positions representing intermediate actions and coming to rest in Position 6 where it is prepared to display signals for the release of the marker. The operation of the ALL BUSY key also started the clock and restored the timer switch to normal so that it could measure suitable work time before displaying release signals.

The call progresses through successive events in a manner similar to that described above. After obtaining a trunk link frame and a register as in condition (1), the sequence switch stands in Position 2 while the timer switch counts work time preceding a request for a line link frame. During this time other markers may request service causing the clock to halt, interrupting the advance of time in all circuits until the operators have completed the required actions. After this marker has counted the specified time, a signal is passed through Position 2 of the sequence switch causing a request for a line link frame to be displayed at the marker position. A delay is handled as before, the marker waiting until the busy frame is released by some other marker. When the frame is obtained, a signal at the match position requests that operator to match

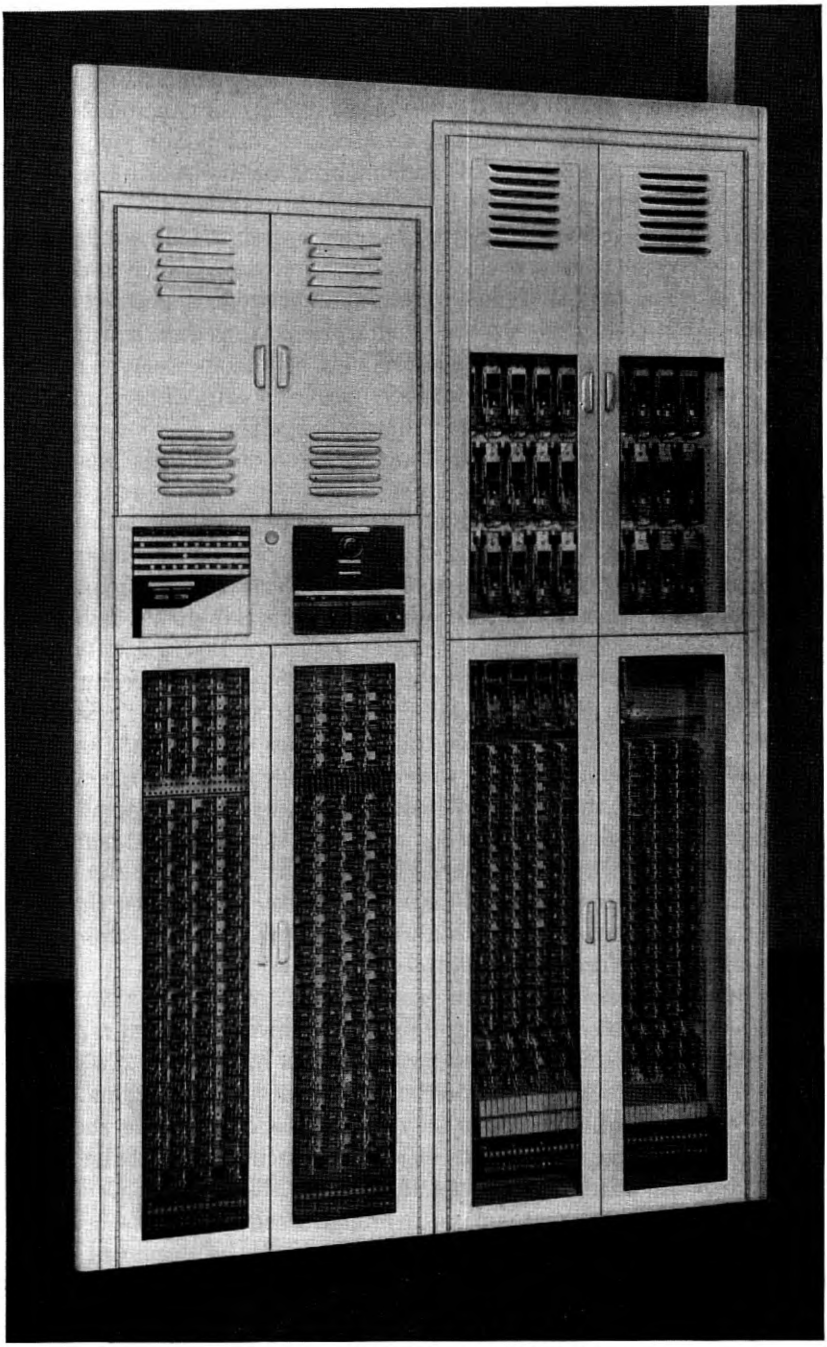


Fig. 14—Relay and switch cabinets.

a channel before allowing time to advance. In case of a mismatch the match operator depresses a MISMATCH key. This operates a memryo relay which will cause the marker to follow the alternate sequence indicated by the side branch at POSITION 2 in Fig. 13. As a safety precaution, the MISMATCH key is made effective only at the time a request for a match is made so that accidental operation at other times will not disturb the circuit action. With a mismatch recorded, the sequence switch advances to Position 3 and at the proper time displays signals to release the line link frame, trunk link frame and register. Thus the call advances with the possible alternates of all registers busy at Position 4 or a second mismatch at Position 5. The call proceeds to one of its possible conclusions where frames are released and the marker becomes idle.

The dial tone call which has been described is the least complex call handled by the machine. The intraoffice call which was diagramed in Fig. 4 is the most complex. It requires 22 sequence switch positions to represent events and may have 92 possible sequences depending on conditions encountered. To take care of variations in sequence in all calls a total of nine memory relays is provided in each marker.

The circuit equipment consists, largely, of telephone type electromagnetic relays and rotary stepping switches. Approximately 800 relays are used. The total number of switches is 57. Of these 47 are of the 22-position, 6-circuit type while the remainder are of the 44-position, 3-circuit type. This equipment is mounted in the two cabinets shown in Fig. 14 and in additional units located within the operating positions. Time indications are given by four-digit message registers, approximately 40 being used. The random circuit consists of a gas tube counting ring with several control relays. Output indications are given by miniature neon lamps. Signals are given to the operators by telephone switchboard lamps, 822 being used. Manual equipment used by the operators in signaling to the circuits consists of keys, switches and telephone plugs and jacks. The machine contains 187 keys and switches, 60 cords equipped with plugs and 509 jacks.

PREPARATION FOR A THROWDOWN RUN

There are two phases in the preparation for a throwdown run. One of these is the tentative engineering of an office of the size to be tested. The other is the preparation of data to represent traffic handled by this office. The first of these follows rather closely the general procedure that would be used in planning the installation of a new switching office.

The chief difference is that preliminary decisions concerning the size and general characteristics of the office to be tested will be made. For example it may be decided to test an office in the twenty line link frame size range serving mixed business and residential subscribers with a high percentage of calls completed within the office. Based on general knowledge of subscriber behavior, the number of subscribers necessary to present a suitable traffic load to this number of link frames will be determined. Values for average holding time and calling rate associated with these subscribers must also be developed. These may, in part, be determined from estimates or specific knowledge of the traffic capacity of the line link frames, taking care that the figures are typical of such a group of subscribers as determined by field observations. Since we are usually concerned with determining the maximum safe capacity of the system, full load or overload conditions will be assumed and the usual margins for future growth considered in engineering an actual office will be omitted. Decisions will also be made as to the number of other offices to which this office has trunks and the percentages of the total traffic originated and terminated in each of these offices.

Having made these preliminary assumptions concerning the nature and environment of the office to be tested, the office is then engineered according to the best available information. The numbers of registers and markers are determined and arranged in connector groups according to standard procedures. The sizes of the trunk groups to various connecting offices are determined and the placement of trunks on the trunk link frames chosen according to the usual practice. All similar factors concerning quantities and arrangement of equipment are determined. The throwdown machine is then set up to simulate this office. This will involve crossconnections in the gate circuits and arrangement and designation of the facilities provided for keeping the busy-idle records of such items as lines, trunks, registers and links.

The second phase in the preparation is to produce data representing calls presented to the system during the time interval to be studied. This is accomplished by choosing a random number for each call. This number must contain a sufficient number of digits to specify all the pertinent data necessary to describe the call. These digits are assigned to represent certain factors. For example, the first six digits represent the time of origination. The next two digits specify the type of call. This is done on a percentage basis. For example, if 40 per cent of the traffic is to be locally completed, the numbers 00 through 39 in these places would indicate an intraoffice call, if 25 per cent is to be outgoing to other offices the numbers 40 through 64 would indicate this type of call, and so on for the remaining types of calls. If divisions of less than one per cent are to be made, three or more digits could be used for this purpose.

The meanings of certain of the remaining digits will depend upon the type of call. If the call is originated within the office, the next five digits will give the identity of the calling line in terms of its frame location. If the call is incoming from another office certain of these five digits will be used to specify the office of origination and the trunk number. This again is on a percentage basis depending on the fraction of total incoming traffic expected from each office. Five other digits give the identity of the called line if the call is completed within the office; if outgoing they specify the terminating office. Other digits give the percentage of calls which will result in partial dialing by the calling subscriber. Since, as will be discussed later, there is a possibility that these will be re-originated later as good calls, all of the information for a good call is also determined for these calls. Additional random numbers determine various other aspects of the call such as the identity of the number group which will be used. As a suggestion to those who would undertake a throwdown study, it is advisable to include a number of surplus "utility" digits in the original random number. It invariably happens that some factor is overlooked in the initial planning or arises during the course of a test and these digits can be used in making decisions in these cases.

It should be noted that the random circuit is used for making certain random decisions in the course of a call at the time that a need for these decisions arises. For example, the holding time of an established connection and the probability of the called subscriber not answering is determined by this circuit. This circuit could have been eliminated by including digits in the original number for every possible situation of this type which might be encountered. This would cause much useless work in preparing the data since all situations are not encountered on every call.

A quantity of random numbers must be drawn to provide the desired load on the office. This is not a simple process of determining the number of calls expected during a given period and drawing this number of random numbers. One factor is that in generating data by the above procedures it will be found that certain numbers will represent calls originated by lines which are busy at the indicated time on a connection established previously. The number of such cases can be estimated from the expected number of busy lines in the office and a corresponding number of additional numbers drawn. When this situation is encountered in the course of the run the call is discarded. The other important factor is a "feedback" effect due to the calls meeting a situation which prevents successful completion and the probability that these calls are originated later.

A simple example is where the called line is found busy. As previously mentioned, there is then a possibility that subsequent attempts will be made at a later time. All attempts make use of circuits and control equipment and should be considered in determining the load capacity of the office. Other situations which produce this effect are illustrated in Fig. 15. Lines entering the figure at the left represent classes of calls which enter the system. A certain number of calls will be partial dials, no dials and false starts regardless of the performance of the system at the time the call is originated. The partial dials represent cases where the subscriber makes an error in dialing or does not wait for dial tone and dials a digit before he is connected to a register. These may be abandoned before the register obtains a marker or may "time out" in the register and be connected to an overflow tone trunk. In either case the subscriber may re-originate the call later. The probability of re-origination has been estimated from field data and the dotted lines in the figure represent these reoriginated calls.

In the throwdown machine the random circuit is used to determine which calls will re-originate and the elapsed time before the second at-

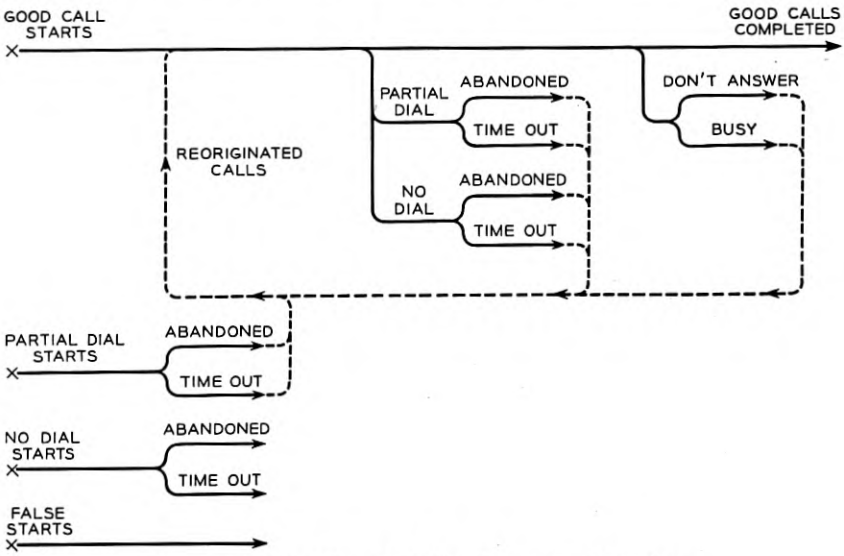


Fig. 15—Composition of the load on a crossbar office.

tempt. No dial starts represent such situations as a change of mind by the subscriber after lifting the receiver or accidentally removing the receiver and are not considered to be correlated with later trials. The same is true of false starts which are momentary start signals often hard to explain. A certain number of partial dial and no dial calls are the result of dial tone delays which may occur during heavy load conditions. These branch from the "good call" line in the figure and are due to the subscriber not waiting for dial tone and dialing part of the digits or even the entire number before being connected to a register. The probability of this occurrence will depend on the extent of dial tone delay. After dial tone delay on each call is known, successive uses of the random circuit determine whether the call is partial or no dial type and if the call is to be re-orginated at a later time. Rough estimates of the quantity of this type of traffic could have been made and included in the original traffic data. However, since they depend on the performance of the system there is a tendency toward a "snowballing" effect and it was thought best to handle it as described in order to detect this effect.

It can be seen that the exact load on a system is difficult to estimate on the basis of initial starts. The procedure then is to make the best possible estimate of initial starts necessary to produce a given load, taking into consideration all important known factors and, at the conclusion of a run, make a count to determine the exact number of calls of various types handled.

When the data for the various calls have been determined from the random numbers, the pertinent information must be transcribed on the call slips. For most calls originated in the office this will requre two call slips. One of these is for the dial tone stage of the call and is used in establishing a connection to a register. The other represents the later stage of the call where the register connects to a marker after dialing is completed and an attempt is made to establish an intraoffice or outgoing connection. The initial time of origination and the calling line number will be recorded on the dial tone slip. The slip for the second stage of the call will carry the calling line number and the called line or outgoing trunk number, but will not carry a time of origination. This is a function of dialing time and is determined at the conclusion of the dial tone stage by the random circuit. It is recorded on the second call slip at that time. The two slips are associated by recording the serial number of the associated slip on the dial tone slip. Incoming call slips carry the origination time, the calling trunk, and the called line numbers. Call slips for partial dial and no dial initial starts carry information similar to that on the dial tone slips.

In preparation for a run all slips bearing initial time entries are stacked in order of their times of origination. Slips for the second stage of a call are kept in a separate stack. At the conclusion of a dial tone action when the time of origination of the second stage is determined, the associated call slip is located, the time is entered on this slip and it is inserted in proper order in the stack of originating calls. In the case of re-originated calls, the information from the old call slips is recopied on a new slip with the new time of origination and these slips are placed with waiting calls in the proper time order.

Detailed Description of Equipment and Functions of the Operating Positions

THE ORIGINATING POSITION

The originating position, in relationship to the rest of the throwdown machine, appears to the extreme right of Figs. 1 and 10. Detailed views of the two wings of the position are shown on Figs. 16 and 17. The principal features of this position are arrays of jacks and wooden pegs representing subscriber lines and incoming trunks, together with associated time counters and detectors. The chief function of the originating operator is to enter all calls into the system at appropriate times. This requires that the stacks of call slips, visible in Figs. 16 and 17, be held at this position. These call slips are arranged in order of their originating times, with latest time on top, and carry the originating line identification number so that line peg and call slip can be associated when the call is to start.

The line array, split into two wings, contains jacks and pegs for all subscriber lines in the office. The jacks which are simply holes with no electrical function, are arranged in a coordinate grid to assist in quick location. Twenty frames are provided, each holding 500 lines for a total of 10,000 lines.

The array is divided into 20 horizontal sections, running across the two wings, each representing a line link frame. Each frame is divided vertically into 10 subgroups of 50 lines, each representing a horizontal group. Since the directory number assigned to a subscriber line in a crossbar office is purely arbitrary and has no physical significance, it is not used in the present case for line identification. Rather, an equipment number, which represents the location of the line on a line link frame, is used. This is a five digit number, stamped on the peg, and made up

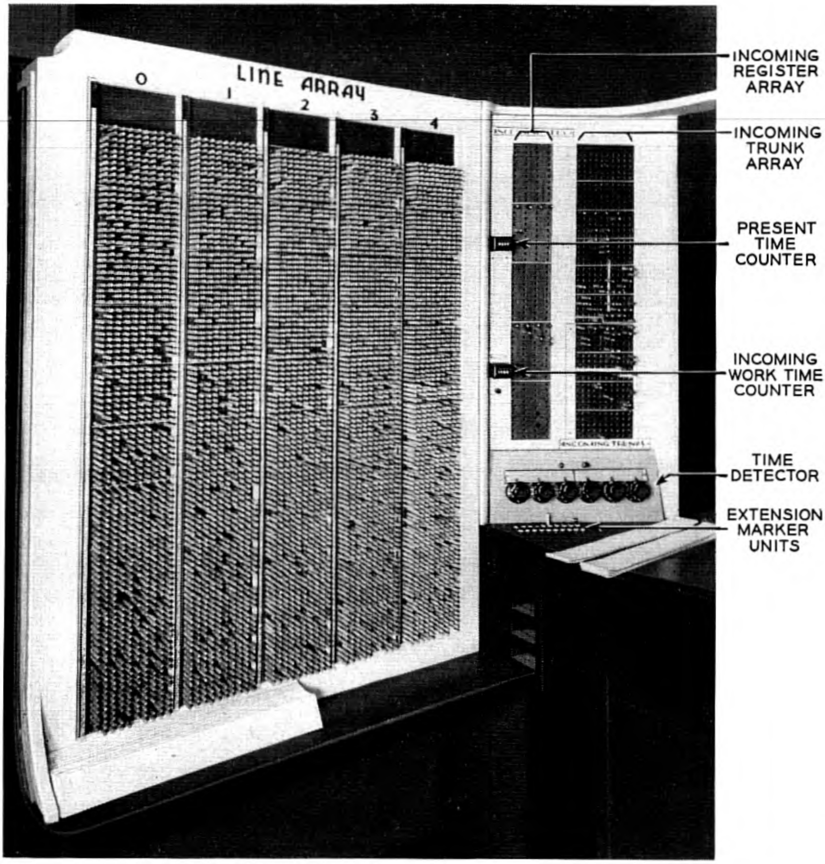


Fig. 16—Originating position—left side.

as follows:

HG—LLF—L

where HG—horizontal group No. 0–9
LLF—line link frame No. 00–19
L—line No. 00–49

The right upper five lines in each horizontal group of a frame (lines 45–49) are reserved for PBX use. A PBX can be assigned to line jacks occupying the same relative location in a vertical row (20 line link frames-same horizontal group). The pegs in a PBX group are marked with a distinctive color to facilitate hunting.

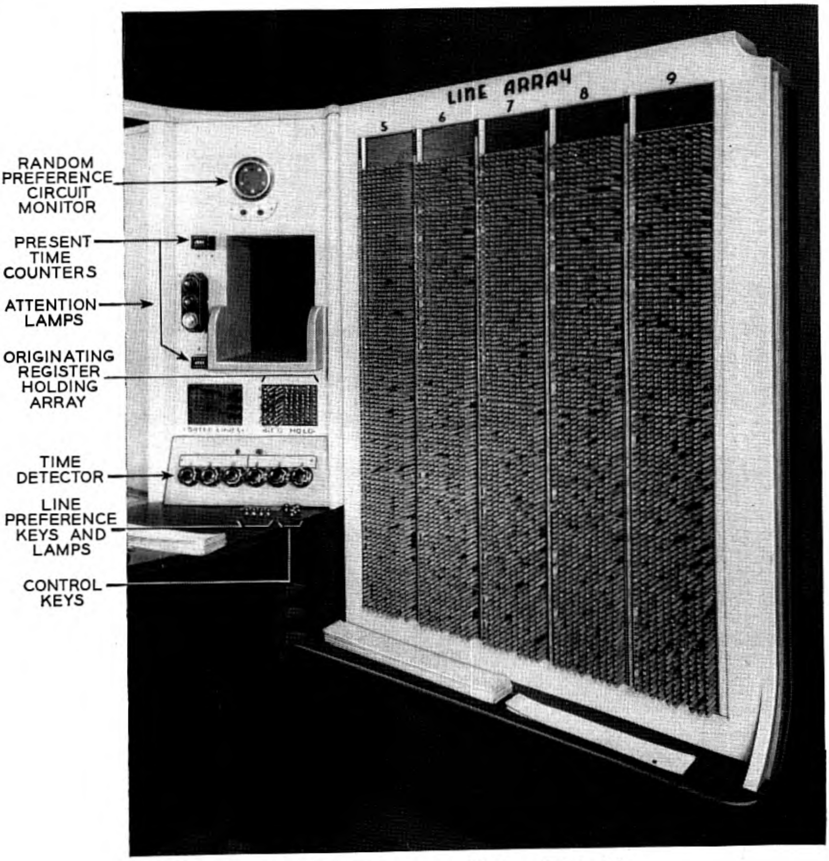


Fig. 17—Originating position—right side.

Line pegs are removed from the array while they are occupied on a call. The jack thus remains empty during the busy interval. The originating operator utilizes a line peg and a call slip in originating each dial tone call.

The incoming trunk and register arrays are located on the left column at this position as shown on Fig. 16. The incoming trunk array consists of pegs and jacks arranged in accordance with the trunk groups. The primary horizontal grouping is assigned a route number which is used only for identification purposes. The horizontal rows within a route represent the number of the trunk link frame switch on which the trunk is located. Vertical rows represent trunk link frames. Provision is made for a maximum of 400 incoming trunks. Trunk identification is a four-digit number made up as follows:

R—TLF—SW

where R—route No. 10, 11, 12, 13
TLF—trunk link frame No. 0–9
SW—switch No. 0–9

Jacks representing these same incoming trunks also appear at the assignment position.

The incoming registers consist of short sleeves which can fit over the trunk peg. When a call is originated by a particular trunk as indicated by the time on a waiting incoming call slip, the trunk is associated with the correct type of incoming register by slipping the sleeve corresponding to the latter over the trunk peg. The peg and sleeve then accompany the call slip into the system. Appropriate times for each action are entered on the call slip.

The incoming register sleeves are held on arrays adjacent to the trunks. They are identified by a three-digit number:

CONN—REG

where CONN—marker connector No. 0 or 1
REG—register No. 00–19

Waiting holes are provided at the register array for holding trunks if all registers are busy.

MARKER CONNECTOR OR GATE

This is a bridging position between the originating position and marker position. Through it must pass all call slips and pegs requiring marker service. The originating operator places call slips and pegs in the gate from one side and the marker operator removes then at the other.

The gate, shown on Fig. 18, provides jacks for pegs and slots for call slips arranged in blocks corresponding to the individual marker connectors for line link frames, originating registers and incoming registers. It is divided into two sections, a storage section in which calls wait for an idle connector, and an active section in which calls wait for an idle marker and removal by the marker operator. An associated relay circuit controls the flow of traffic through the gate, simulating actual No. 5 operation, and indicates by lamp the appropriate action.

When a call slip is ready to enter the gate, the originating operator obtains the corresponding line or register peg and places the peg and slip in a jack and slot, respectively, in the correct connector block of the storage section of the gate. The connector block number is ascertained from the peg identification or from the call slip. If the connector is free a CONN READY (R) lamp lights to indicate that the peg and slip should be advanced to the connector block in the active section of the gate. If the connector is not ready, the lamp will light at some subsequent time and the connector delay time will be noted on the call slip. If several calls pile up in a particular storage connector block, means are provided for maintaining correct preference for entry into the active section in the case of registers. With line originations, a random line preference circuit under key control at the right side of the originating position key shelf picks, on a random basis, the next entry into the active section of the gate (except in the case of preferred lines whose pegs will have a distinctive marking) and which have precedence over ordinary lines. Keys for a choice of one out of two, three, four or five are provided. With

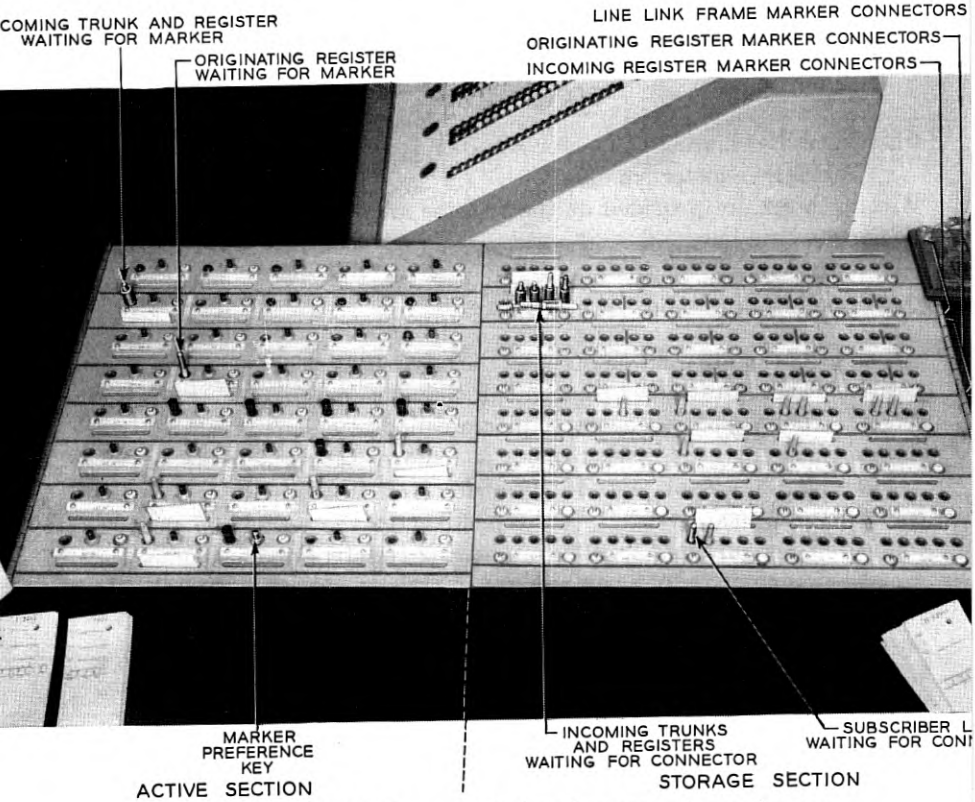


Fig. 18—The gate for control of marker connector preference.

register originations, the operator controls the preference with the assistance of a lamp signal per connector and a written register preference record.

Once in the active section of the gate, a call awaits an idle marker. When correct preference conditions are met, a CONN CLOSED (C) lamp lights, together with an attention lamp, to indicate to the marker operator which is the next call to be served. The marker operator depresses the MKR PREF key in the connector block to lock in a signal at the marker position indicating the marker to be used on this call. The operator then removes the call slip and peg from the gate and inserts them in the correct unit of the marker position. The match ticket portion of the call slip is torn off and passed to the match operator. On originating register controlled calls, the register match ticket will have been attached to the call slip and this also is passed to the match operator.

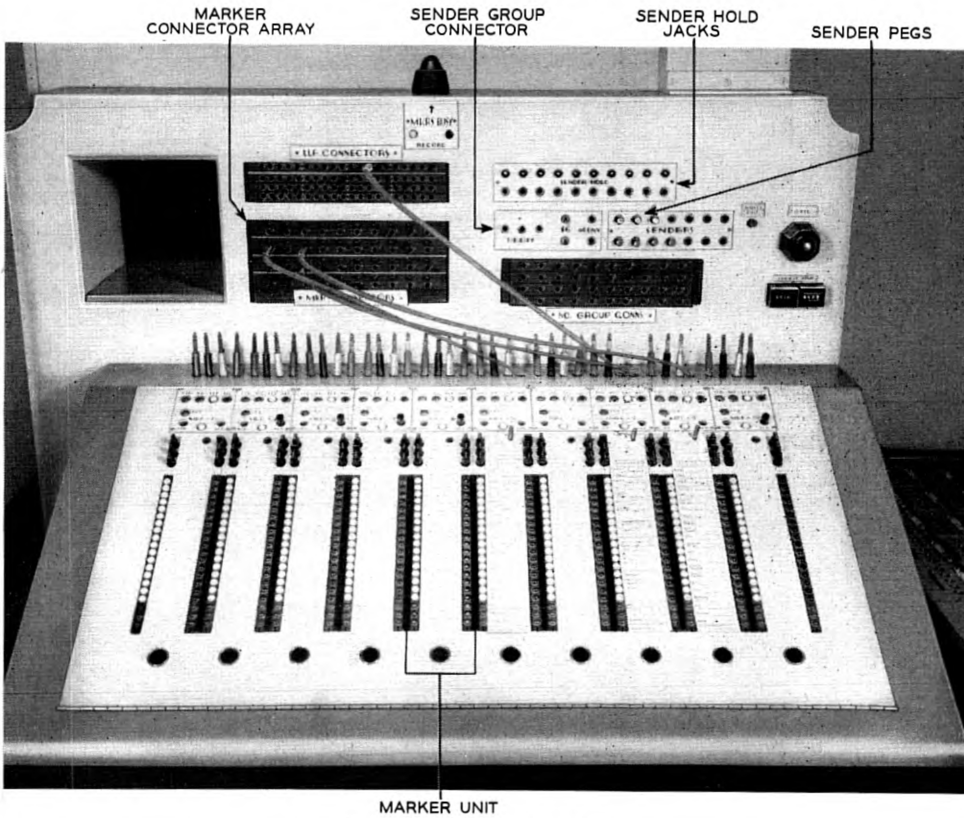


Fig. 19—Marker position.

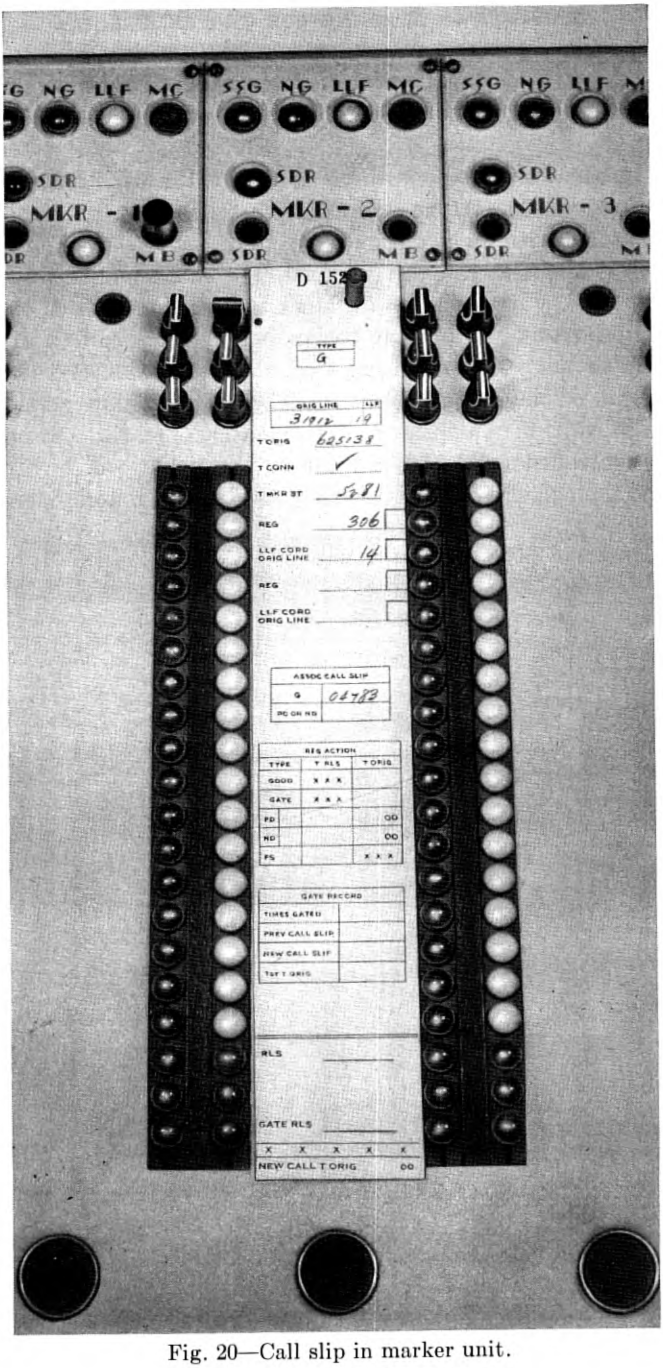


Fig. 20—Call slip in marker unit.

The circuit associated with the gate assigns all calls and markers according to No. 5 preference arrangements and provides the correct gating action. This preference is set up on cross-connection field located within the position and may be changed to correspond to any desired system arrangement.

MARKER POSITION

A photograph of the marker position is shown on Fig. 19. The individual marker units, of which ten are provided, are disposed on the sloping work panel. On the array panel are jack arrays representing the line link frames, the number groups, outgoing sender subgroups and senders, and the marker connectors. Connections to these jacks are made by means of plugs and cords located at the top of each marker unit.

The call slip for dial tone class of call is shown inserted in a marker unit on Fig. 20. At the top of the slip, in the TYPE and ORIG LINE boxes and opposite T ORIG and T CONN, are typical entries as made by the originating operator. The marker operator has also made three time entries.

The heavy dot at the top edge of the call slip indicates which of the six class keys to turn. The class keys condition the marker circuit to handle each type of call correctly. The hole at the top of the slip lines up with a jack and is used to store temporarily the line peg (register or incoming trunk peg in the case of other types of call) as received from the gate; it also serves to locate and hold the call slip in the correct position.

On the lower portion of the slip are spaces corresponding to all marker actions requiring association with other frames or circuits. These spaces line up with lamps on the marker unit which signal when and what action should be taken and indicate whether or not the time should be entered on the call slip. If a delay is encountered at any point, a check is put in the associated delay block to assist in subsequent computations involving the call slip. The left hand row of lamps lights to indicate when a direct action should be taken and a time written down. For example, when the third lamp from the left hand top lights (opposite LLF CORD) the operator places the line link frame cord at the top of the unit in the correct jack in the LLF connector array, and writes down current time in the space opposite the lamp. If the LLF jack is occupied by another marker, the operator plugs the LLF cord in a preference delay jack associated with the line link frame jack and places a check mark in the delay block.

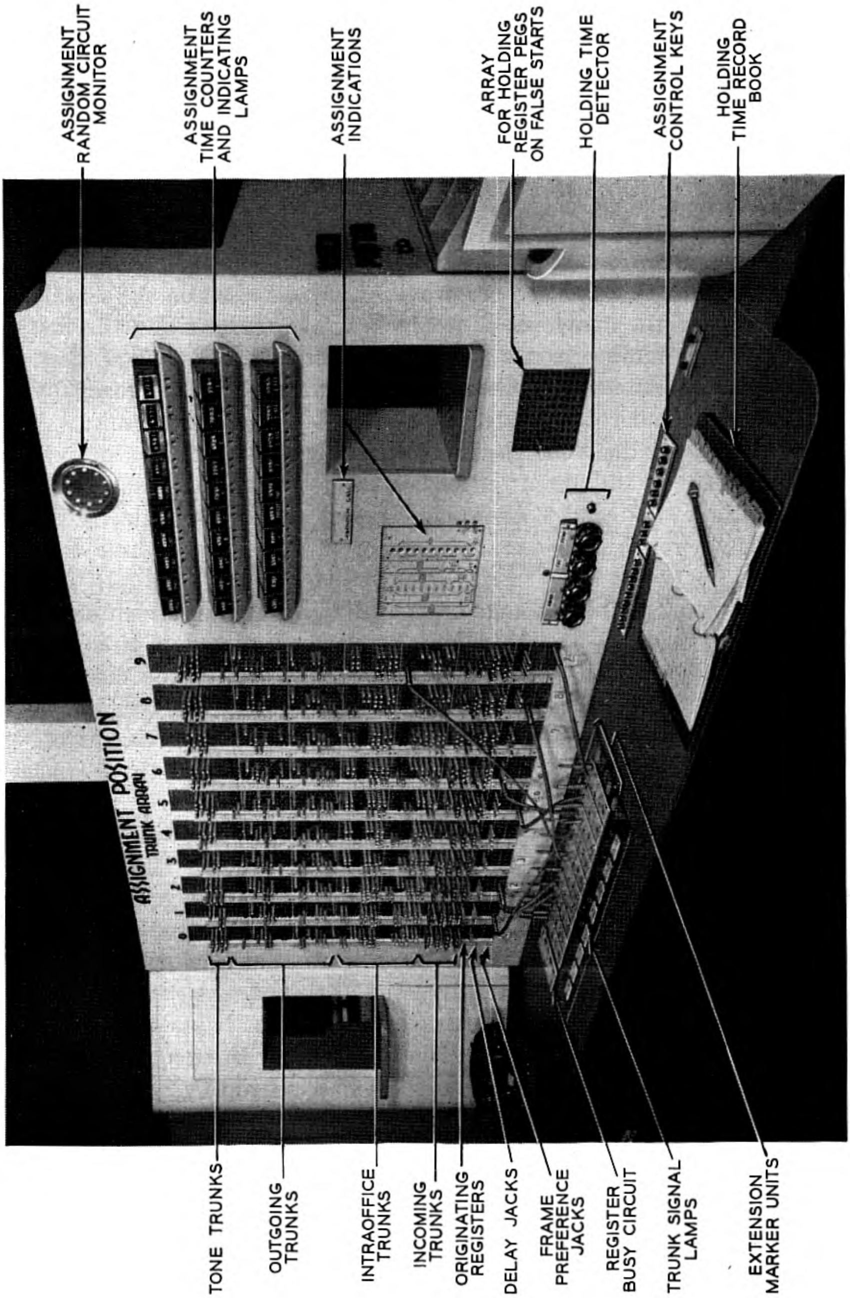


Fig. 21—Assignment position.

As discussed previously, certain marker functions are located at the assignment, originating and match positions. The choice of trunk link frames and trunks, for example, is made at the assignment position. Therefore certain lamps at the marker position light only to signal the writing down of a time. This is true of the second lamp from top left which indicates register seizure. The lamps in the right-hand row are also used in conjunction with actions at other positions. They light only to indicate that a delay should be checked in the corresponding box.

At the time of marker release, both left and right lamps light opposite one of the release categories. The specific type of release determines partially the further disposition of the call.

The START key at the bottom of each marker unit is pressed after completing each action called for by the sequence lamps. Operation of this key puts out the lamp and permits the circuit to advance.

ASSIGNMENT POSITION

The assignment position, shown on Fig. 21, includes the trunk array, the register false start holding array, the traffic assignment equipment, the holding time book and counters, and individuals units representing extensions of the markers. The chief function of the assignment operator is to test and choose trunks on each call, to determine the disposition of certain calls which encounter excessive dial tone or busy conditions, and to ascertain holding times of originating registers and trunks.

The trunk array consists of jacks (holes) representing all the trunks, and jacks and pegs representing the registers. Line pegs are held in the jacks during conversation time to mark the trunks busy. The array is divided into vertical sections representing trunk link frames. The trunk groups are disposed in horizontal rows so that trunk jacks of each group or route occupy the same relative position in each frame section. Thus a trunk-hunting action consists of picking a horizontal level in the array and searching along the level to the first idle trunk as indicated by an empty jack. The array provides more jacks than the total number of trunks so that most small trunk groups can be set up on individual horizontal levels.

The trunk identification is a three or four digit number composed as follows:

R—TLF—SW

where R is route number 0–99
TLF is trunk link frame 0–9
SW is switch number 0–9

The route number in some cases is required only for location purposes since many trunk groups are identified by name. It is assigned to a common group of trunks, maximum 100, or 10 per frame. However, in the case of outgoing trunks, the route number is required to identify the specific trunk group desired. The route number for originating registers is the same as the marker connector number for each register and is unusual in that any route number from 0–9 represents the same register group.

The register jacks occupy the lower trunk level on the array and are supplied with pegs which are used to originate the second half of outgoing and intraoffice calls. During dialing time, the associated line peg occupies the register jack to mark it busy. The next higher level on the array is the group of incoming trunks. These are furnished on a two-jack per trunk basis and are, in effect, multiples of the incoming trunk array at the originating position. The two jacks are required to hold the terminating line peg and the incoming trunk peg. The route number for these trunks is significant only for the function of location since it is not necessary for throwdown purposes to segregate incoming trunks into groups as it is with outgoing trunks.

The intraoffice trunks are divided into an A- and a B-group since more than 20 trunks per frame are required. The No. 5 marker can only test up to 20 trunks per frame at one time. The machine automatically allots calls between the two groups. Two jacks per trunk are required for originating and terminating line pegs. Unlike the incoming trunk jacks each jack of an intraoffice pair corresponds to a different switch location, although only one switch number is used for identification purposes.

The outgoing trunk groups are disposed above the intraoffice trunks. For the most part, these trunks will be in small groups, each with its own route number. When a call is set up to an outgoing trunk, it is necessary for the marker operator to inform the assignment operator of the correct route number. Several of these trunk groups can occupy the same horizontal level to conserve space. Only one jack per trunk is required to hold the originating line peg.

Tone trunks (busy, overflow, partial dial, no dial trunks) are at the top of the array. Only one jack per trunk is required.

Within the trunk link frame, the preference order in which a No. 5 marker tests trunks and the manner in which the order shifts from call to call is rather complex. In order to reduce the load on the assignment operator in trunk hunting, the actual trunk preference is approximated by reversing the direction of hunting within a frame group from call to call on a substantially random basis. Thus, when the operator determines the trunk frame within which she will hunt, she observes the LEFT and RIGHT lamps below the array for the indication as to whether to hunt from left to right or vice versa.

TRAFFIC ASSIGNMENT

One of the important functions of the assignment operator is to determine the assignment of all calls at the time of marker release. For this purpose she is furnished with traffic assignment keys and indicators which appear to the right of her position. An adequate understanding of this feature requires a somwhat detailed explanation.

With the aid of the traffic assignment controls and indicators, shown to the right of Fig. 21, the assignment operator performs the following specific functions:

On Dial Tone Calls: The operator determines whether a call reaching a register should be classified as a good call (successful subscriber dialing), a PD call (partial dial—incomplete subscriber dialing) or an ND call (no dial—no subscriber dialing while connected to a register). A proportion of call slips are originally marked as PD or ND (and also FS or false start) and on these this determination need not be made. If the call is of the PD or ND type, the operator determines whether it should be subtypes PD1, PD2, PD3 or ND1, ND2, ND3. The subtype affects the assumed time until the subscriber abandons the call. If the call is classified as good type, the operator determines which of several dialing times should be used. If the call is classified as PD, ND or FS type, the operator determines when the call is abandoned and whether or not it is routed to a tone trunk.

On Calls Completed to a Subscriber: On calls completed via intraoffice, outgoing or incoming trunks, the operator determines whether the call is answered and which of ten holding times should be assigned for subscriber line and trunk.

On Calls Routed to a Tone Trunk: On calls which are routed to tone trunks or given a tone signal from the register, the operator determines the trunk or register holding time and whether and when the call is re-originated.

In making these determinations, the operator presses keys which cause a lamp to light either beneath a time counter or beside a designation strip. The time counter, set ahead of present time, indicates trunk release time, register return time, etc., while the designation strip classifies the calls. The determining factors include the magnitude of dial tone delay, probability, and whether or not all registers are busy. Where dial tone delay is concerned, it is determined by comparing the originating time of a dial tone call slip with three dial tone delay time counters. These counters give present time minus 1, 2 and 9 seconds respectively so that matching the time of origination against them indicates whether the delay was < 1 sec., 1–2 sec., 2–9 sec. or > 9 sec.

The probability factor is obtained from a circuit which is capable of lighting one lamp out of ten, one out of three, etc. on a random basis when a key is depressed. By means of the circuit, calls can be assigned to various categories in correct proportion in accordance with the best available traffic information. Whether or not all registers are busy is indicated by the all register busy circuit controlled by the assignment operator.

If the assignment for a trunk or a register is that it be held for a period of time and then released, the assignment operator makes use of the holding time book as described later. If the assignment is for a register to return with a bid for a marker, or a new call to return into the system (after encountering busy, for example), the time is noted on the call slip and the latter is passed to the originating operator for subsequent action. For new call return time, a letter designation associated with the signal lamp is also entered on the call slip. The letter is carried forward on the new call slip. If subsequent attempts of this same line meet busy or overflow, the letter designation is used to identify the same category of return time instead of using the random circuit.

When a trunk call is set up, the trunk and one or two lines must be kept out of service for one of several fixed holding times. There are ten different assigned holding times with an equal likelihood of an established call falling into any one of them as determined by the traffic assignment circuit. False start, tone trunk and don't answer connections provide four additional holding times.

The holding time counters provided at the assignment position indicate present time plus a fixed holding time. Thus each counter gives the time at which a connection, set up at present time and assigned to that particular holding time, will release its elements back into service.

Holding time starts for a given call at the time the release lamp lights at the assignment position marker unit associated with that call. At such a time, the assignment operator obtains the one or two line pegs of the call and plugs them in the trunk jack or jacks (identified by the frame connector cord), noting at the same time the trunk number. The operator then depresses a key which causes the traffic assignment circuit to light a lamp under one of the holding time counters, thereby assigning a release time. The trunk number and release time are recorded. At the end of the holding time, the operator removes the line peg or pegs from the trunk jacks and the pegs are returned to their home jacks. Presence of the line pegs in the trunk jacks during holding time marks the trunk busy.

The busy trunk numbers are recorded in a holding time book according to release times. Time units in the system represent millionths of an hour (0.0036 second). Each page of the holding time book covers 10,000 units of time and is divided into one hundred 100-unit blocks. An illustration of one page of the book is shown on Fig. 22.

Since the time unit is one millionth of an hour, each time unit in a one hour series can be represented by a six-digit number. Of these six digits, as far as the long holding times are concerned, the last two are unimportant since 100 units is only 0.36 second. If they are dropped, the first two digits of a four digit holding time release figure give the 10,000 time unit group or page number of the holding time book and the second two digits give the 100-block on the page. Trunk numbers are entered in the book on this basis with the page and block number obtained from the assigned holding time counter.

By this means the release times of all items appear in consecutive order in the holding time book. Holding time entries are always made several 100-blocks beyond the next release time, which eliminates con-

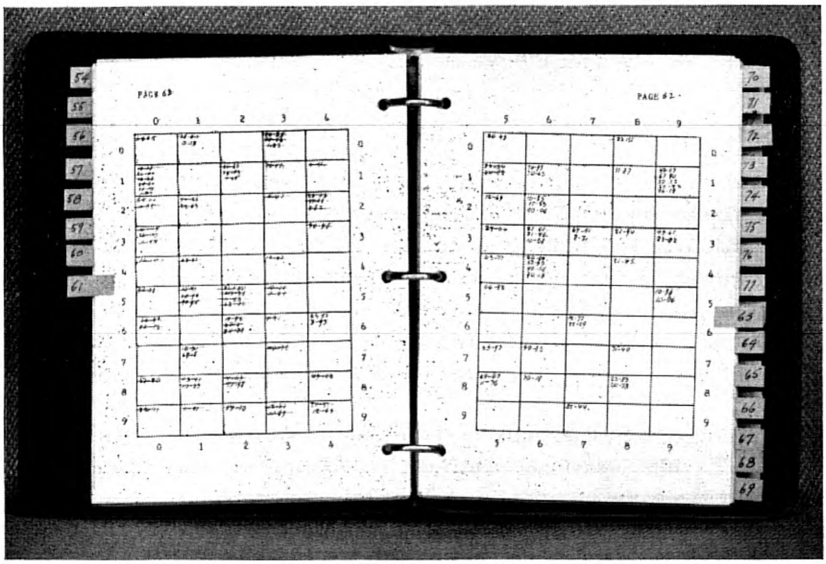


Fig. 22—Holding time book.

fusion. Following the simple numerical order of release times, the operator sets up the next release time on the holding time detector, similar to the time detectors at the originating position, which stops the system and signals when that time arrives. The operator returns the held items listed in that 100-block to normal, crosses out the block and sets up the next following time on the detector.

False start and register overflow release time entries are made in the same holding time book. A jack array is provided to hold the register peg until release time under these conditions.

Since the longest holding time is six minutes (100,000 units), the operator is never concerned with more than 11 pages at one times. With each page clearly labeled by a numbered margin tab, search time to make an entry is reduced to a minimum.

MATCH POSITION

The match operator performs the function of testing and making busy the switching channels through which each connection is set up. Her position includes ten marker units, each consisting of a group of lamps, and a set of files which hold the channel cards by means of which the channel records are kept. The position is shown on Fig. 23.

The No. 5 marker picks out a channel between a subscriber on a line link frame and a trunk on a trunk link frame by matching a line link, a junctor and a trunk link which are capable of being switched together to connect the two end points. A schematic of the system is shown on Fig. 24.

Each horizontal group of subscribers has direct access to a set of ten line links which connect to the ten junctor switches of the frame. Each link of the set can be given a number from 0 to 9 corresponding to the junctor switch on which it terminates.

Verticals on each junctor switch connect to junctors which are distributed over all the trunk link frames in the office. In a ten trunk link frame office, the ten verticals of each line link frame switch are distributed to all ten trunk link frames. This provides a set of ten junctors from each line link frame to each trunk link frame. If there are less than ten trunk link frames in the office, additional sets of junctors, perhaps comprising less than ten junctors per set, are provided between frames. The junctors connect between like-numbered switches and within a set bear the same number (0 to 9) as the switch.

The trunk links are similar to the line links except that twenty links connect from each trunk switch to the ten junctor switches. The twenty links are subdivided into left and right sets of ten which connect to the left and right halves of the junctor switches respectively. Within each set, the links are numbered in accordance with the junctor switches on which they terminate. The junctor switches are split horizontally as shown on Fig. 24 to provide for twenty junctor connections per switch (one from each of twenty line link frames). Thus a set of junctors terminating on the left halves of a junctor switch must be matched with a left set of links.

It can be seen on Fig. 24 that when the three sets of links and junctors capable of connecting a trunk and a subscriber are matched, the two No. 0 links and the No. 0 junctor go together, the two No. 1 links and the No. 1 junctor go together, and so forth. The marker performs the

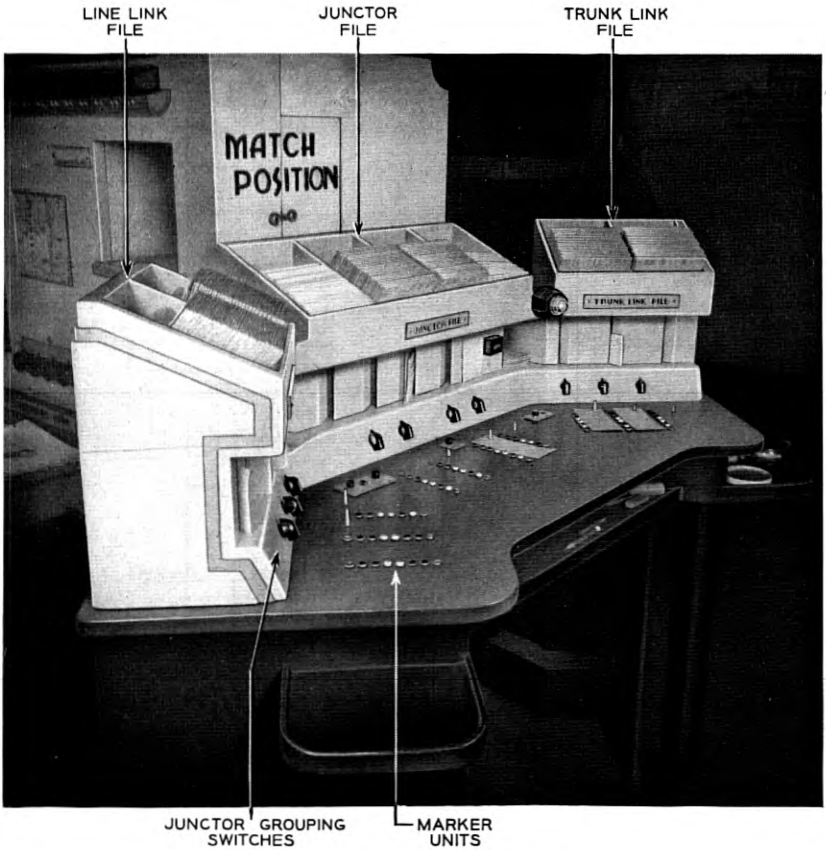


Fig. 23—Match position.

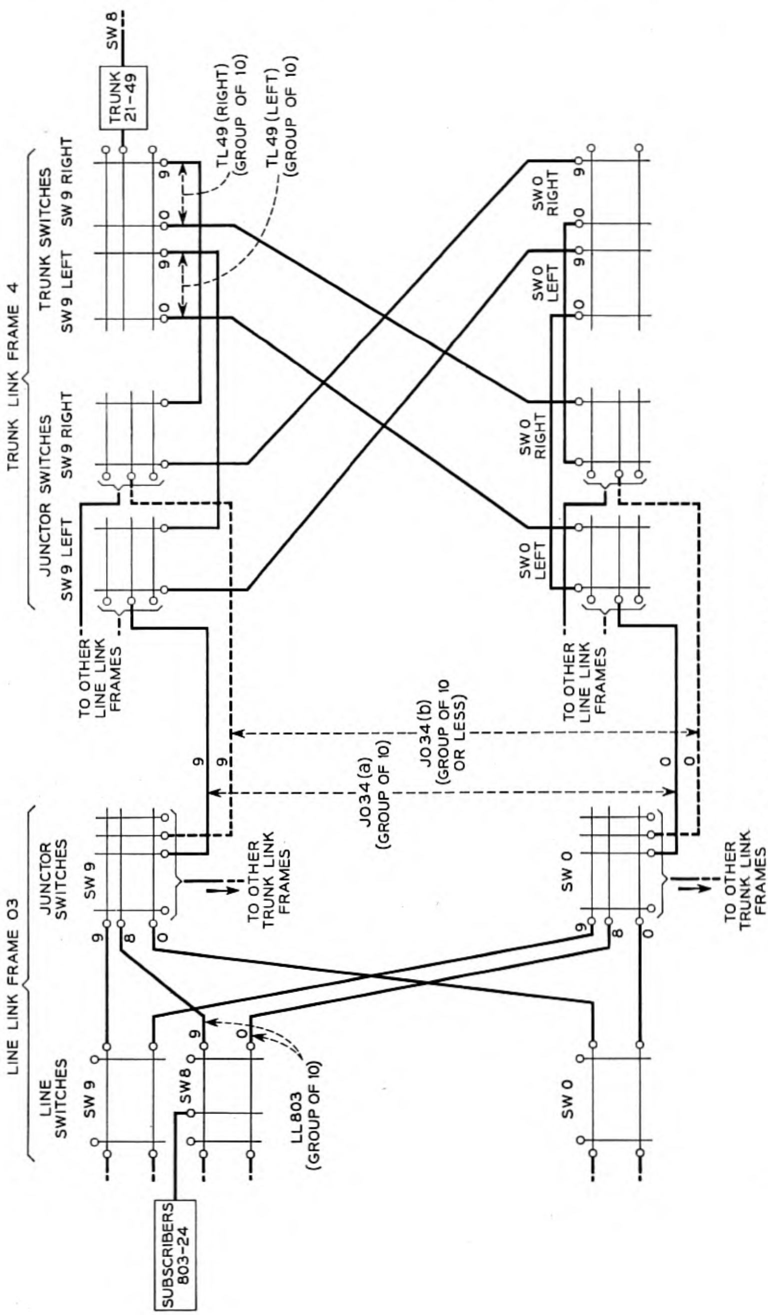


Fig. 24—Channel matching schematic.

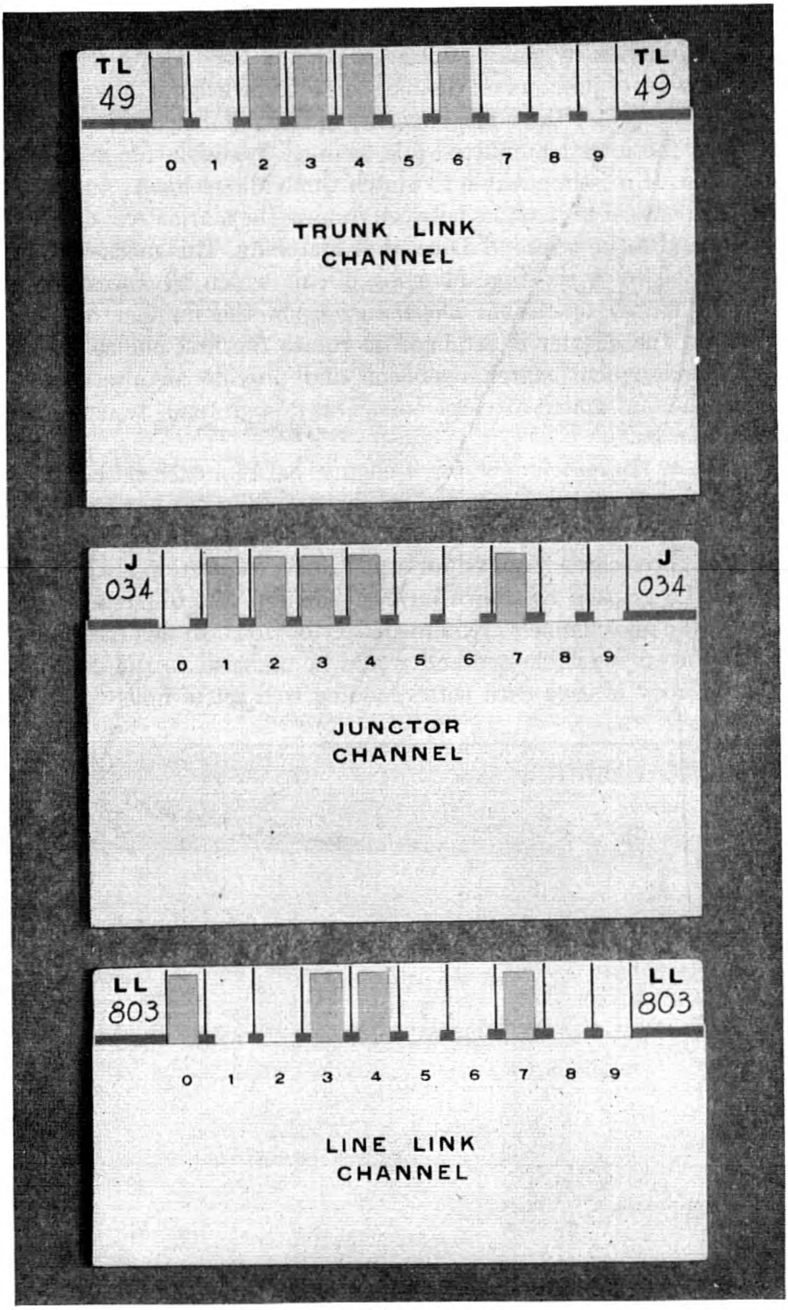


Fig. 25—Channel cards.

matching function by gaining access to a set of line links, a set of trunk links and a set of junctors on the basis of its knowledge of line and trunk location. It tests for three idle elements of like number within these sets and picks the lowest numbered idle channel available for making the connection. If it is impossible to match three idle elements and there is more than one set of junctors between frames, the marker will change the junctors and make a second attempt at matching. This marker function is performed by a six-stage allotting circuit which advances one step for every match operation. Depending upon the number of junctor subgroups, the allotter is arranged to rotate the first choice subgroup on each subsequent match operation and provide an alternate subgroup if the first match attempt fails. This system tends to equalize the use of junctors.

For use by the match operator, a channel card for each set of ten links or junctors is provided. Each card, as shown on Fig. 25, has ten pockets, one per link or junctor element, in which can be inserted a busy tab. When the three cards required for a particular connection are identified, they can be stacked as shown on Fig. 26 (note the difference in card size) and the idle channels are immediately obvious. In this case, channel 5 is the lowest available one and would be assigned to the call.

The identity of each card corresponding to a set of links or junctors

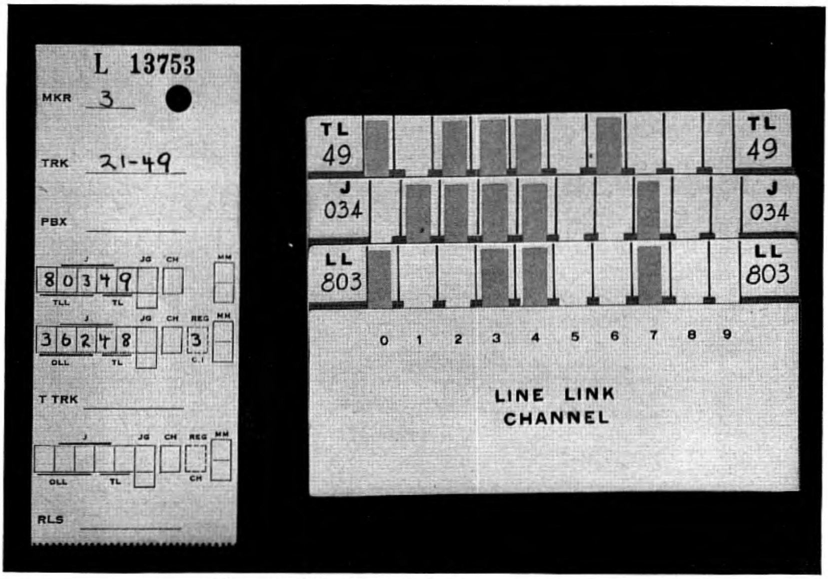


Fig. 26—Channel matching procedure.

is determined by switch and frame numbers. There are ten sets of line links per line link frame and a maximum of twenty frames. The identification of each set is a three digit number made up as follows:

SW—LLF

where SW—switch number (0–9)
LLF—frame number (00–19)
This provides for a total of 200 line link sets or cards.

The number of the line link set is incorporated in the line identification number so that from the latter can be determined immediately the particular line links available to the line. Note on Fig. 24 that line 803–24 must use the LL803 set of links (card shown on Figs. 25 and 26) for any connection.

There are twenty sets of trunk links per trunk link frame (ten left and ten right sets) and a maximum of ten frames. The identification is a two digit number plus a left or right indication. The number is composed as follows:

TLF—SW

where TLF—frame number (0–9)
SW—switch number (0–9)
and the left or right indication is, for convenience, one of two colors. This provides for a total of 200 trunk link sets or cards, 100 of each color. The link identification is included in each trunk number. Thus, on Fig. 24, trunks 21–49 must use trunk link set TL 49, either left or right.

The number and disposition of junctors are determined by the layout of line link and trunk link frames. In a 20 line link-10 trunk link frame office, there is one set of junctors between each pair of frames. For fewer frames there are more sets between frames to a maximum of five for a four-line link, two-trunk link frame office. The junctor sets or cards are identified by the two frame numbers involved, as

LLF—TLF

where LLF—Line Link Frame No. (00–19)
TLF—Trunk Link Frame No. (0–9)
If more than one set of junctors interconnect two frames, letters A to E are added to the base number. For a particular line and trunk, the junctor number is derived from a combination of the line and trunk numbers. For example, the essential parts of line number 803-24 and

trunk number 21-49 combine to give a number

$$8\underline{034}9$$

where the underlined digits represent the set of junctors (see Fig. 24).

In the general case, the junctors of a set may terminate on either the left or right half of a trunk link frame junctor switch. When the channel cards are made out preparatory to a throwdown study, the junctor numbers are assigned to cards of one of two colors (same as trunk link cards) depending upon whether it is left or right connection. Thus, in picking the three channel cards for a match, the choice of trunk link card depends upon, and must be the same as, the color of the junctor card.

Normally the channel cards are kept in filing sections at the match position as shown on Fig. 23. Before the matching operation is physically performed, the operator has available on a match ticket the line-trunk composite number. In the example shown on Fig. 26 this number is

$$80349$$

The digits 803 identify the line link card; the digits 034, the junctor card; and the digits 49, together with the junctor card color, identify the trunk link card. The operator removes the cards from the file, stacks them, and places them in a slot associated with the particular marker until the match signal is received. At that time, the operator picks the lowest numbered free channel in the stack, marks it busy and enters the channel number on the match ticket for record.

Each channel must be released at the same time that the line and trunk with which it is associated are removed from holding. Since release times for lines and trunks are entered on the assignment position holding time detector, this latter detector is used to signal release times to the match operator. The match operator maintains all her established match tickets in release time sequence with the earliest time on top of the pile. The lighting of a signal lamp indicates that the channel identified by the top ticket should be restored to normal.

A channel release condition which the match operator must recognize without a special signal occurs at the marker release time on intraoffice, outgoing, partial dial and no dial calls. At this time the register channel associated with the call must be dismissed. The operator will have received from the marker operator the register channel ticket involved and must restore the channel to normal before answering any new signal.

MASTER CONTROL PANEL

A master control panel for the throwdown machine is supplied on the relay cabinet shown on Fig. 14. This panel provides: means for turning the system on and off at the beginning and end of each day's operations; a centralized alarm indicating system; a bank of marker action lamps which indicate marker status during operation; and a present time counter with a units-to-seconds conversion scale.

The equipment operates on two battery supplies, one known as permanent battery and the other as day battery. Permanent battery is on continuously during a complete throwdown run in order to hold operated certain record relays. The day battery, however, is turned off during idle periods.

The equipment is arranged so that at the end of a working period all operator functions requested by signal lamps during the last working unit of time can be completed before the machine automatically stops. This is controlled by the DAY-NIGHT switch which is turned to the NIGHT position when it is desired to cease operation. The NIGHT lamp lights at this time. When the operators have extinguished the last signal lamp, the DAY POWER switch is turned to OFF.

Certain critical portions of the machine are provided with alarm circuits which automatically stop the machine and light lamps at the control panel. In most cases, the lighting of one or more of these lamps will require troubleshooting. When the trouble has been found, operation of the key associated with the lamp extinguishes the lamp and permits the machine to start again.

RESULTS OF THROWDOWN STUDIES

The throwdown machine has now been in operation for slightly less than four years. During this time 1383 seconds of equivalent central office time, divided among eleven runs, have been accumulated. The machine, of course, has not been in continuous operation, since the time required for preparation of a run and eventual analysis of output data exceeds the actual operating time for the run. In general, the same team of girls has handled both preparation and analysis of data and operation of the machine. Beyond this, there have been periods when no studies were in progress.

A detailed presentation of throwdown results is not properly within the scope of this article. However, a brief resume of the several runs with some mention of their primary objectives and typical results is necessary to conclude this picture of the throwdown machine.

It should be emphasized again that the principal function of a throwdown machine is to provide data, under controlled conditions, which can be used to develop and check a comprehensive theory of operation of a complex system. A secondary function is to study the reaction of the system to specific equipment or circuit arrangements. The No. 5 throwdown machine has been used for both purposes. All runs have furnished masses of statistical data which have been very useful in formulating traffic theories applicable to No. 5 crossbar. Some data, such as those on link matching, have a more general field of application.

All throwdown runs have employed a basic size office of twenty line link frames and ten trunk link frames, loaded by 9000 subscribers. On the first two runs, a traffic level designed to load these frames to their normal capacity was applied to the machine. This level can be called

TABLE III

Run	Duration	Quantity		System Loading	Primary Purpose of Run
		Markers	Orig. Registers		
	Seconds			*per cent*	
I	256	5	67	100	To load up machine and give normal load picture
II	216	5	59	100	To study effect of higher marker occupancy due to reduced number of registers
III	90	5	59	125	To study effect of 25 per cent overload
IVa	86	5	59	110	To establish response of system with original gate preference for comparison with run IVB
IVb	94	5	59	110	To provide data on response of system with reversed gate preference
V	65	6	65	—	This was an intermediate run in which traffic at high level was introduced in order to build up to system equilibrium at the 120 per cent load level
VI	108	6	65	120	To study system at high load level below saturation
VII	288	7	60	120	To study effect of situation where registers are more severe bottleneck than markers
VIIIabc	180	4	35	140	These three runs tested two proposed changes in the gate control circuit against the standard reversed preference arrangement. Identical traffic was used for each run

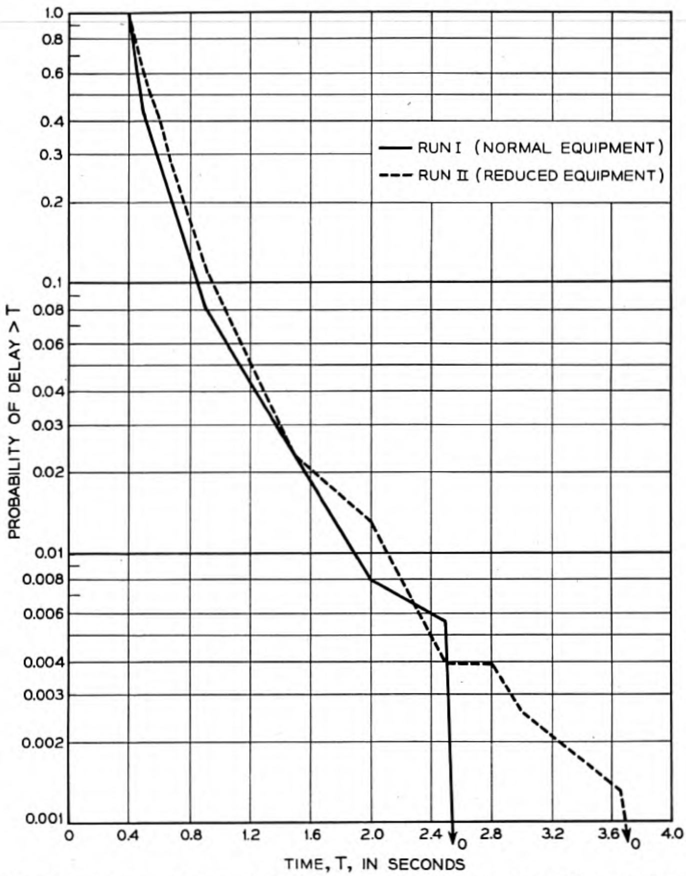


Fig. 27—Dial tone service with different equipment quantities as determined by the throwdown machine.

100 per cent load. In succeeding runs, an overload of varying amounts was utilized. In the several runs, the quantities of registers and markers were varied to obtain basic engineering data.

As the No. 5 system was originally engineered, the preference order in which markers were assigned by the gate control circuit to marker connectors during heavy loads was as follows: (1) line link frame marker connectors; (2) originating register marker connectors; (3) incoming register marker connectors. At a later date, this order was changed to put the line link frame marker connectors last in preference. In the discussion which follows, this arrangement will be known as "reversed preference."

The essential features of the throwdown runs to date are given in Table III.

The curves of Figs. 27 and 28 are representative of the type of data made available by the throwdown machine. Fig. 27 shows the overall dial tone service obtained during the first two runs. The slight degradation of service in Run II is caused by the reduction in number of registers. In both cases, however, service is very good.

Fig. 28 shows the spread of delays met by line link frames in obtaining a marker on dial tone calls during the same two runs. These curves are representative of the distributions of delays encountered at individual stages of handling a call. Other examples might be line link and trunk

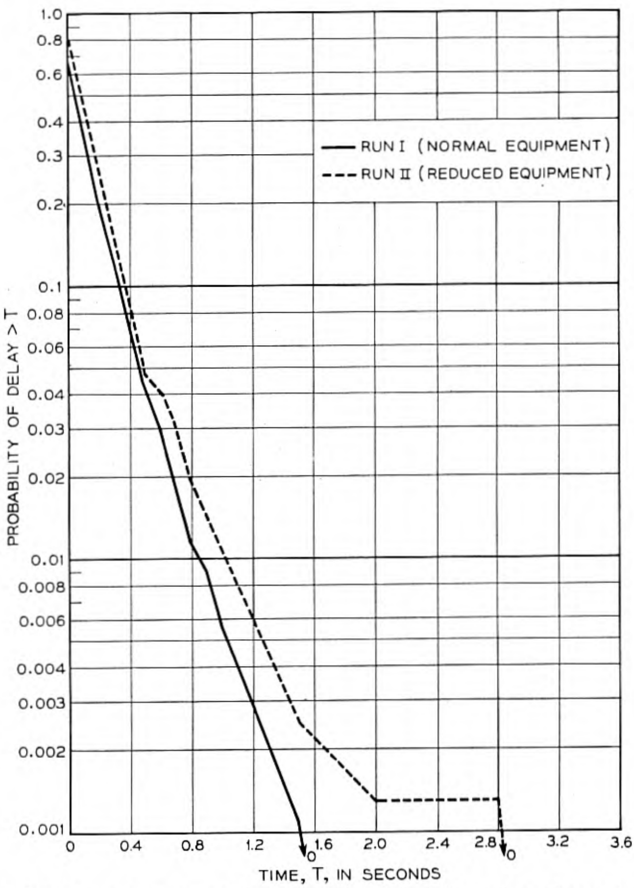


Fig. 28—Marker delays on dial tone calls with different equipment quantities as determined by the throwdown machine.

link frame delays. Taken together, these various delays determine overall grade of service.

An interesting example of the practical utility of the throwdown machine is furnished by the series of events which lead to the introduction of reversed gate preference in the No. 5 crossbar system. The changing quantities of line and register pegs in the gate position provide a graphic visual indication of the dynamic status of the system. After watching the gate for some time in the early overload runs, it was noticed that the register marker connectors were relatively less successful in gaining access to markers than the line link frame marker connectors. During all register busy periods, this reduced the call handling capacity of the system since registers were delayed in becoming available to waiting lines. The effect was compounded by wasting marker time in attempting to set up dial tone calls when no registers were idle.

It was felt that a change in gate preference, placing register marker connectors before line link frame marker connectors, would improve this. The new arrangement was tested and confirmed in throwdown run IV and is now a system standard.

Working Curves for Delayed Exponential Calls Served in Random Order

By ROGER I. WILKINSON

(Manuscript received December 19, 1952)

Working curves of delays for waiting calls served at random are given for a considerable range of loads and group sizes. Exponential holding time calls are assumed originating at random, and served by a simple group of paths. Results of a number of throwdown tests are given to illustrate the effect on call delays of several modes of service, and particularly of service on a random basis. For random service, these results verify the theory recently developed by J. Riordan; perhaps more interestingly they show the effects on delays of certain blends of queued and random service which approximate methods of handling delayed calls in practical use (such as gating and limited storage circuits). The use of random and queued delay theory is illustrated by a number of examples. To remind the reader that these results are not limited to telephony, department store and vehicular traffic problems are included.

A theory for predicting the delays which telephone calls (or other corresponding types of traffic such as vehicular, aircraft, people waiting in line, etc.) having exponentially distributed holding times would encounter when the delayed calls are served in a random order was published in a recent issue of this JOURNAL* by John Riordan. Mr Riordan's mathematical analysis involved a determination of the first several moments of the delay distributions. He then devised a method of combining elementary exponential curves in such a way as to satisfy the moments previously calculated.

Since a limited number of moments were used in the above determinations the curves derived are approximate only, but at the same time they are believed to be good approximations. The critical cases are those of paths carrying very heavy loads, in the occupancy ranges of $\alpha = 0.80$ or higher.

* Bell System Technical Journal, January, 1953, pages 100–119.

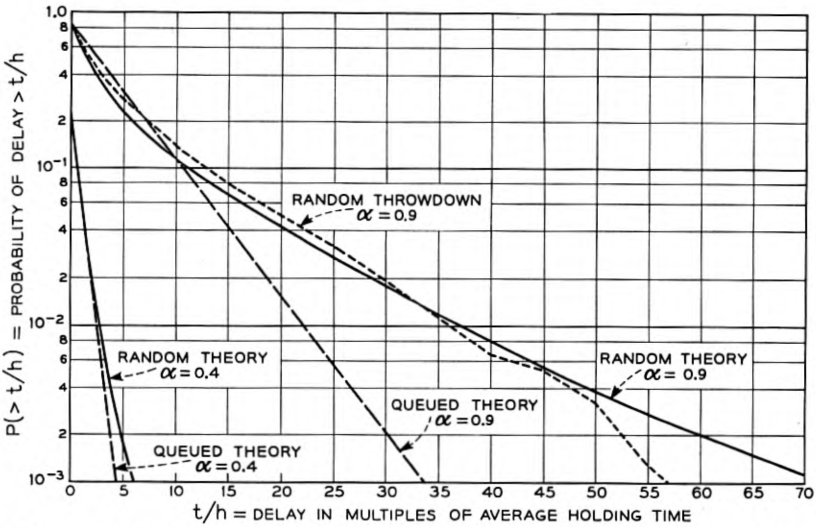


Fig. 1 — Distribution of delays. Theory versus throwdown, delayed calls handled at random, $c = 2$ paths, $\alpha = 0.90$, 3000 throwdown calls.

THROWDOWN CHECKS

Before calculating a field of curves for working purposes it was thought desirable to make at least a modest throwdown test, or traffic simulation, at these high occupancies to observe the agreement of theoretical delays with those determined by a trial in which the theoretical assumptions would be closely followed. This has now been performed at two trunk group sizes, $c = 2$ paths, loaded by approximately $a = 1.8$ erlangs or an occupancy of $\alpha = 0.90$, and $c = 10$ paths at an occupancy of approximately $\alpha = 0.80$.

For these throwdowns, random origination times were obtained through use of Tippett's Random Numbers. An hour was visualized as being composed of 100,000 (or, as in one case, 1 million) consecutive discrete intervals, numbered serially. Choosing 5 (or 6) digit random numbers then provided the start times of the subscribers' bids for service.

Likewise holding times were chosen by random numbers from an exponential universe by dividing it into 100 equal probability segments and assigning each a number from 00 to 99. A central value of holding time was chosen to represent the range of cases within each segment. The last segment, number 99, on the long tail was further subdivided into 100 parts in order to give more definition in the long call lengths which are believed to be critical.

A comparison of the proportion of traffic expected to suffer delays beyond various multiples of the average holding time as given by Riordan's theory for delayed calls served in random order, and by the throwdown results, is given in Figs. 1 and 2. As discussed below, the cases studied are considered to give satisfactory assurance as to the adequacy of the approximations involved in the theory.

The two trunk case based on 3000 calls submitted shows fairly good agreement with the theoretical distribution out to delays as large as 50 multiples of an average holding time which includes more than 99.5

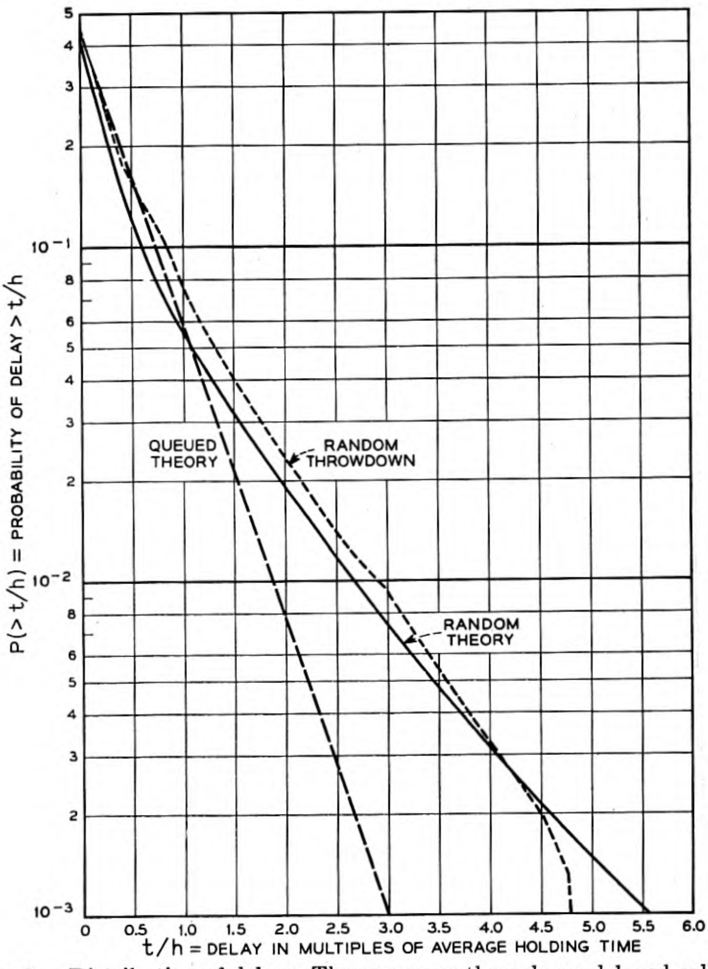


Fig. 2 — Distribution of delays. Theory versus throwdown, delayed calls handled at random, $c = 10$ paths, $\alpha = 0.80$, 1500 throwdown calls.

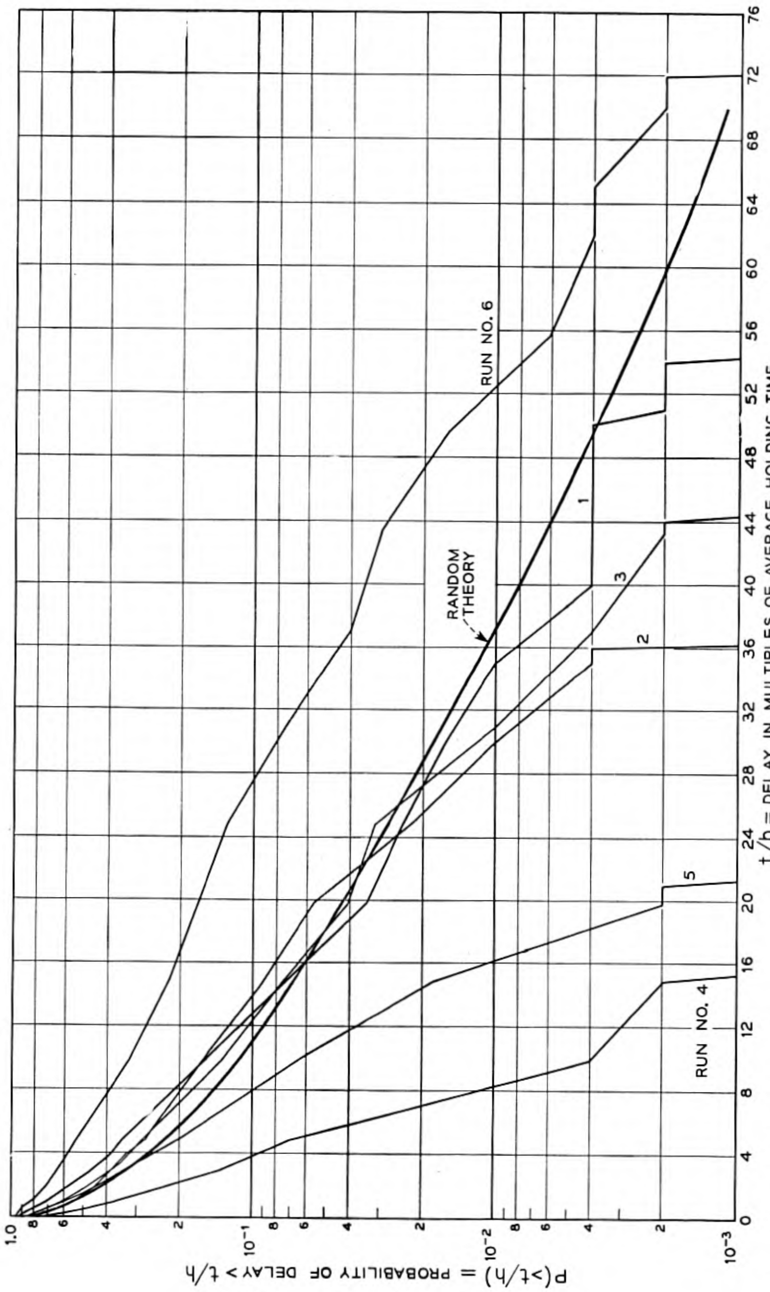


Fig. 3 — Distributions of delays on 2 paths by groups of 500 calls in random delay throwdowns, designed load α = 0.90.

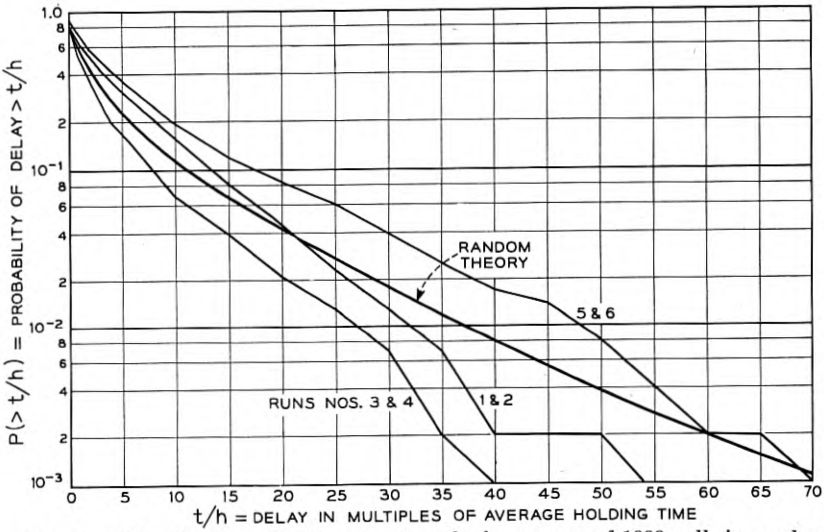


Fig. 4 — Distributions of delays on 2 paths by groups of 1000 calls in random delay throwdowns, designed load $\alpha = 0.90$.

per cent of all the calls delayed. The 10 trunk case based on 1500 calls at 0.8 occupancy also shows good agreement to 99.5 per cent of the calls delayed.

In making throwdown tests of this sort, the criterion for deciding when one has proceeded long enough is rather vague. The usual practice is to summarize the delays at regular intervals and observe at what point it seems likely that making additional tests would not change the results by a sensible amount. For the $c = 2$ trunk case, six runs of 500 calls each produced the several very different broken line curves of Fig. 3 shown superposed on the theoretical delay distribution for $\alpha = 0.90$. Clearly no one of these by itself could be given much weight.

Consecutive runs were paired to form three runs of 1000 calls each, as shown in Fig. 4. As one would expect, their spreads have narrowed appreciably. Combining these three runs yielded the dotted curve of Fig. 1, which, of course, has a correspondingly smaller likelihood of sampling error in it. On the basis of such a succession of narrowing spreads, one can, with some feeling of assurance, estimate within what narrow band about the observed curve the true unknown curve (approachable by many more tests) must lie.

On Figs. 1 and 2 the shapes and positioning of the total throwdown and theoretical curves seldom differ more than 20 per cent on the probability scale down to the $P = 0.005$ probability level. The disparities measured along the delay axis in the higher ranges of the variable, are, of course, considerably less. A comparison of the theoretical and observed proportions of calls delayed, and the average delays on all calls is shown in the following table:

Trunk Group Size, c	Occupancy α	No. Calls in Throwdown	Proportion of Calls Delayed		Average Delay on All Calls in Multiples of Average Hold Time	
			Theory	Throwdown	Theory	Throwdown
2	0.9	3000	0.853	0.855	4.30	4.71
10	0.8	1500	0.409	0.444	0.205	0.254

These differences between theory and observation are well within the variations which would be expected with the lengths of throwdown runs made.

Further reassurance that the traffic submitted in the two throwdown tests originated in a manner reasonably similar to that assumed in the theory was obtained by making "switch counts" at regular intervals during the throwdowns from which frequency distributions, $f(x)$, of the number x of calls simultaneously present were constructed. These are shown in Figs. 5 and 6 for the two throwdown cases. The solid

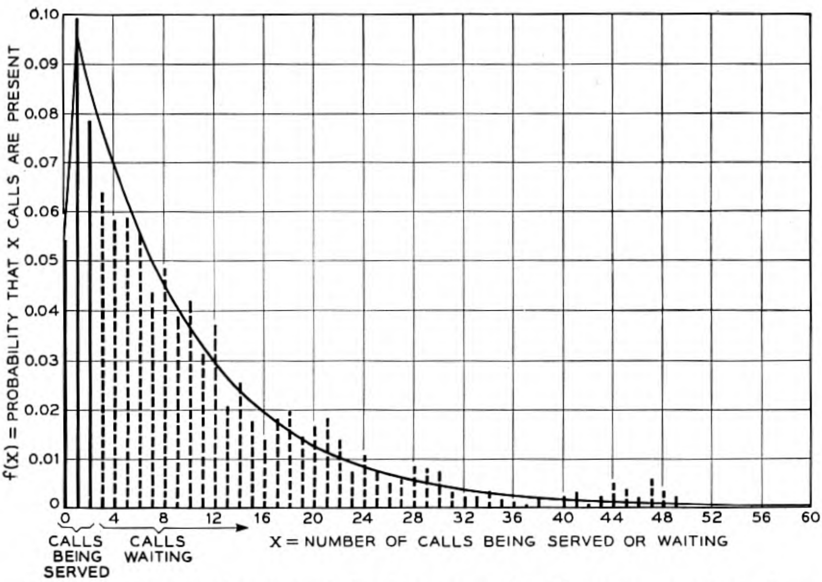


Fig. 5 — Distribution $f(x)$ of simultaneous calls. Theory versus throwdown, $c = 2$ paths, $\alpha = 0.90$, 3000 throwdown calls.

line spikes correspond to observations when all calls in the system were being served, that is $x \leqq c$. The dotted spikes show those proportions of observations when one or more calls were waiting, that is $x > c$. The theoretical values of $f(x)$ are indicated by the smooth curves where they pass over discrete values of x. The theory and observations are seen to be in quite good agreement.

Referring again to the theoretical delays (and the throwdown checks) on Figs. 1 and 2, very much larger delays can obviously be obtained when delayed calls are handled at random than when they are handled in a strict first-come-first-served, or queued, order, the latter distri-

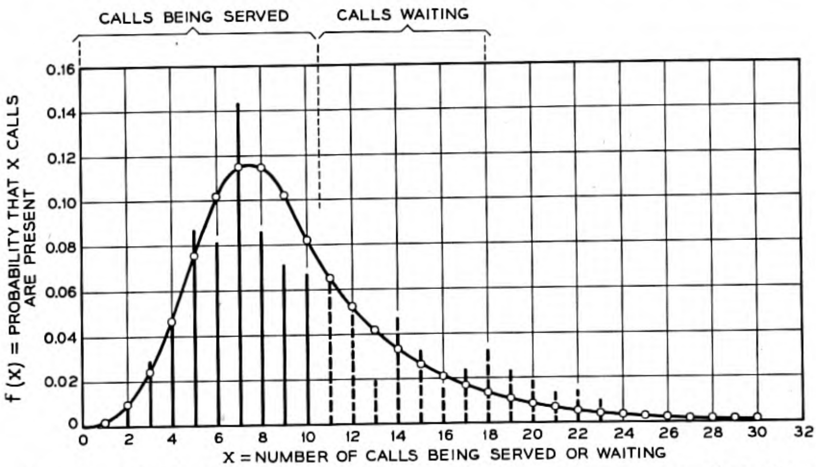


Fig. 6 — Distribution ($f(x)$ of simultaneous calls. Theory versus throwdown, $c = 10$ paths, $\alpha = 0.80$, 1500 throwdown calls.

butions being shown by the straight lines which start at nearly the same ordinates at delay 0 as the random handling curves, and cut down across the lower part of the charts.* Although fewer very short delays occur

* Delay curves for exponentially distrubuted holding time calls in systems where delayed calls are handled in order of arrival, are given by E. C. Molina in "Application of the Theory of Probability to Telephone Trunking Problems," Bell System Technical Journal, Vol. 6, p. 461, July, 1927. They are calculated from the Erlang equation

$$P(>t) = P(>0)e^{-(c-a)t} = \frac{\dfrac{a^c e^{-a}}{c!} \dfrac{c}{c-a}}{\displaystyle\sum_{x=0}^{c-1} \frac{a^x e^{-a}}{x!} + \frac{a^c e^{-a}}{c!} \frac{c}{c-a}} e^{-(c-a)t} \tag{1}$$

where the delay t is expressed in multiples of the average holding time. Values of $P(>0) = C(c, a)$ can be read approximately from Figure 21.

with this method of handling than when a random selection of the waiting calls is followed, the very long delays are markedly reduced, and on this account the queueing procedure is generally preferred. These effects are particularly evident at the higher occupancies. As illustrated in Fig. 1, the "queued" and "random" delay curves at an occupancy of $\alpha = 0.4$ show little difference down to the $P = 0.001$ delay level.

IMPERFECT QUEUEING

Interest has often centered in questions as to what form the delay curves might take in a system in which queueing of the calls is maintained to a limited extent, and beyond which the record of order of arrival would be lost. Such an instance might occur with a team of toll recording operators who were able to keep well in mind the order of arrival of signals up to a certain number waiting, whereupon they would lose track and not regain this ability until the number of waiting calls had again dropped below some small number. Other situations with actual or equivalent limited delay storage arrangements can readily be imagined.

To study a case of limited queueing, a short subsidiary throwdown was next run on the $c = 2$ case, using the 1000 calls of Runs 1 and 2 of Figs. 3 and 4 (which comprised the 1000-call sequence most closely approaching the theoretical distribution). Three rules for delayed call handling were tested:

(1) Delayed calls are served in random order.

(2) Delayed calls are queued (served in order of arrival).

(3) Delayed calls are queued until more than w are waiting at which time their arrival order is lost and they are served at random. When the number waiting again drops below w, newly arriving calls are queued behind those randomized calls still waiting. Note that case 1 corresponds to $w = 0$, and case 2 to $w = \infty$.

The comparative results are shown on Fig. 7, with w given successively values of 0, 8, 20, 25, 30 and ∞. The $w = 0$ curve, of course, is taken directly from Fig. 4 for Runs 1 and 2 combined. Although this curve does not agree particularly well with theory (Curve A), its movement with changes in w is nevertheless instructive. As seen, queueing as far as $w = 8$ waiting calls produced practically no improvement in the delay distributions. (Perhaps with the occurrence of such large numbers of waiting calls, reaching a maximum of 35, one could not expect queueing of so few as 8 to have much effect.) The next selection of $w = 20$, however, still showed only a relatively slight improvement, particularly in

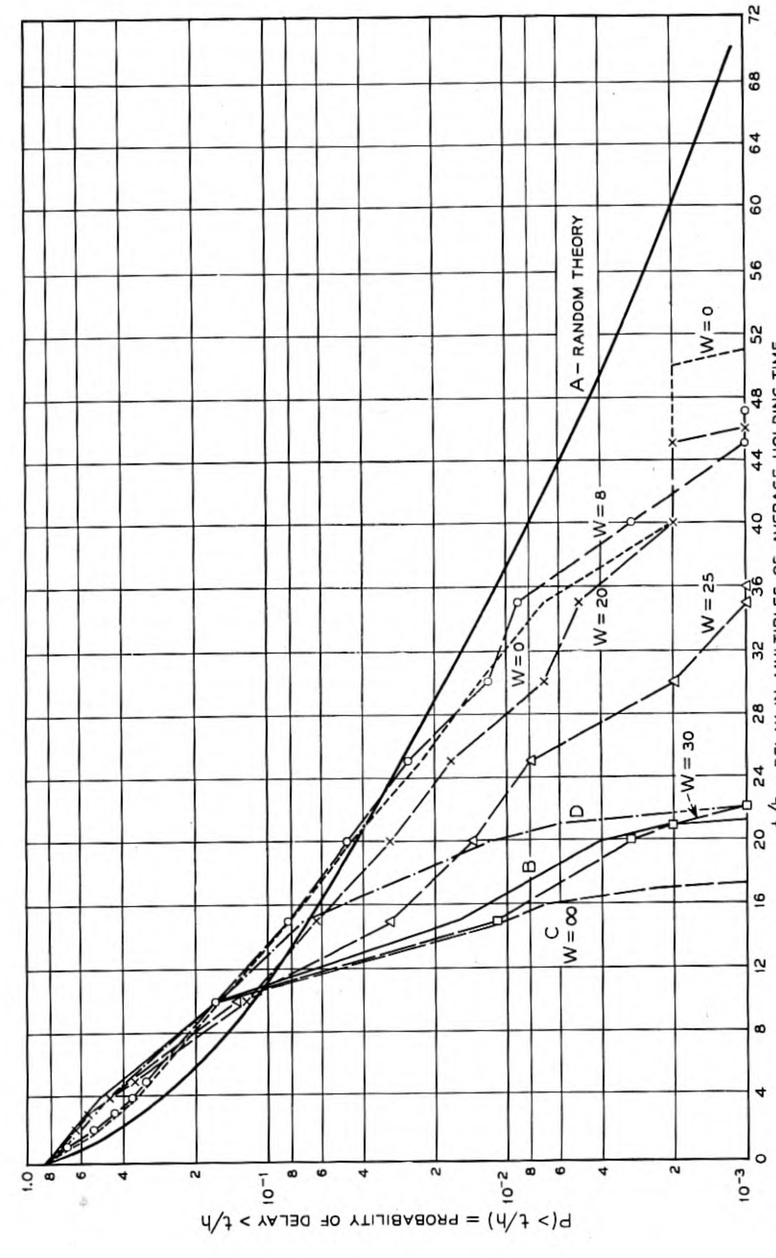


Fig. 7 — Throwdown delay distributions with imperfect queueing, $c = 2$ paths, $\alpha = 0.90$, 1000 throwdown calls.

the long delays occurring on a few of the more unfortunate calls. Even choosing $w = 25$ moved the delay curve hardly more than half way from the completely random to the fully queued curve of $w = \infty$. The $w = 30$ selection shows the accomplishment of nearly fully queued results, the latter being given by curve C ($w = \infty$). Thus one would apparently find little value in instructing a team of two operators working at an occupancy of 90 per cent to try to remember the order of arrival of waiting calls unless they could keep track of an unexpectedly large number.

Electrical storing circuits have long been used to assist the ordering of waiting calls. They have the especial advantage of not becoming confused and losing the order of the calls which have engaged them up to the limit of their storage capacity. In the Bell System two methods of approaching true queueing are in common use. In one method, such as found, for instance, in the No. 3 Information Desk, a number of storage circuits are provided so that as a waiting call is served from the number one storage position, all the others waiting on storage circuits drop down one position. If s such circuits are provided, and more than s calls have been waiting, one of the excess will then be chosen at random to occupy the newly vacated sth storage circuit.

The second method used widely in both local and toll systems is known as gating. In its simplest form a gate opens into a "corral" where the operators or other service media are located. So long as calls simultaneously demanding service do not exceed the number of operators (trunks, markers, etc.) the gate is ineffective. As soon as one call has to wait, the gate closes until that call obtains service, and then admits to the corral all calls which have accumulated on the outside. The gate again closes until all calls within the corral are served; and so on. Thus the calls are admitted in bunches to the corral. Between bunches there is strict queueing but within bunches when they get inside the gate the calls are substantially served at random. As long as the bunches are small the effect of true queueing is approached. In any event a strong safeguard against excessively long delays on a few unlucky calls is introduced. In the Bell System, a variety of gating plans are found such as double gates, gates with additional preferences for certain types of calls, and schemes for placing calls outside the gate again if they cannot be served immediately. Each of these must be studied with its own peculiar characteristics in mind.

To illustrate the effectiveness of the storage circuit type of automatic queueing arrangement, the 1000 calls of Runs 1 and 2, for the two path case, were processed by a throwdown through a two operator, 20 storage

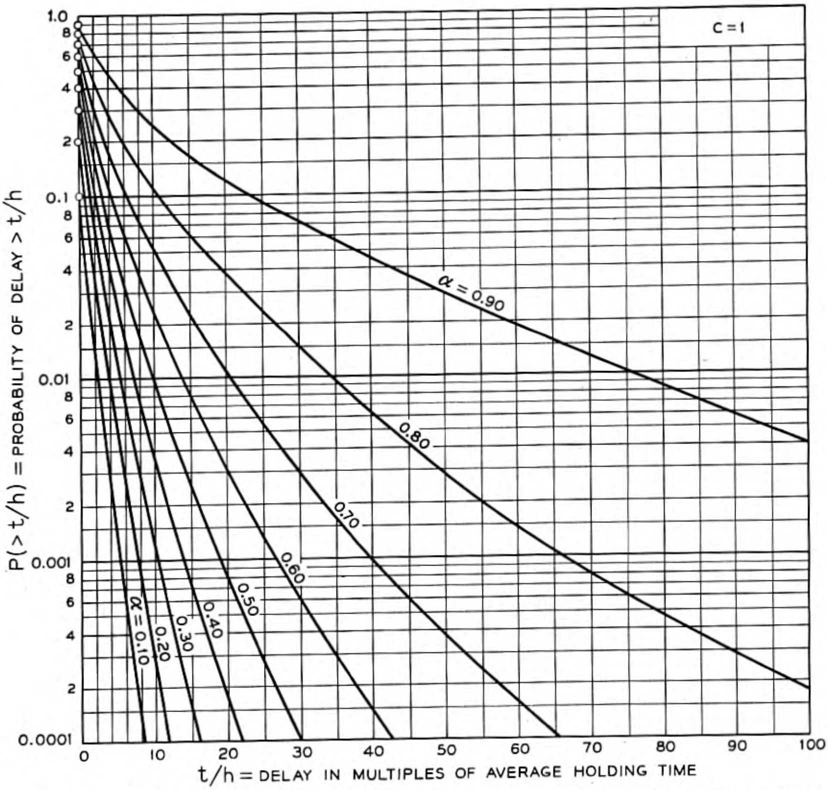


Fig. 8 — Delayed traffic served in random order, exponential holding times, $c = 1$.

circuit system. The resultant delay distribution is shown as Curve B on Fig. 7. (It is appreciated that this hardly represents a tolerable normal operating situation, but rather illustrates what the performance might be under extremely heavy traffic conditions.) The results are very close to those obtained with perfect queueing (Curve C) and show in striking fashion the gains in service to be made in certain delay situations by providing a limited storage apparatus with a memory not subject to confusion during moments of heavy overload.

When the 1000 calls of Runs 1 and 2 are submitted to the 2 paths through a simple gate in order to produce approximate queueing, the resultant delays are shown by Curve D on Fig. 7. Large improvements again occur in reducing the very long delays found with random handling. In fact by use of this simple (and usually relatively inexpensive) gating

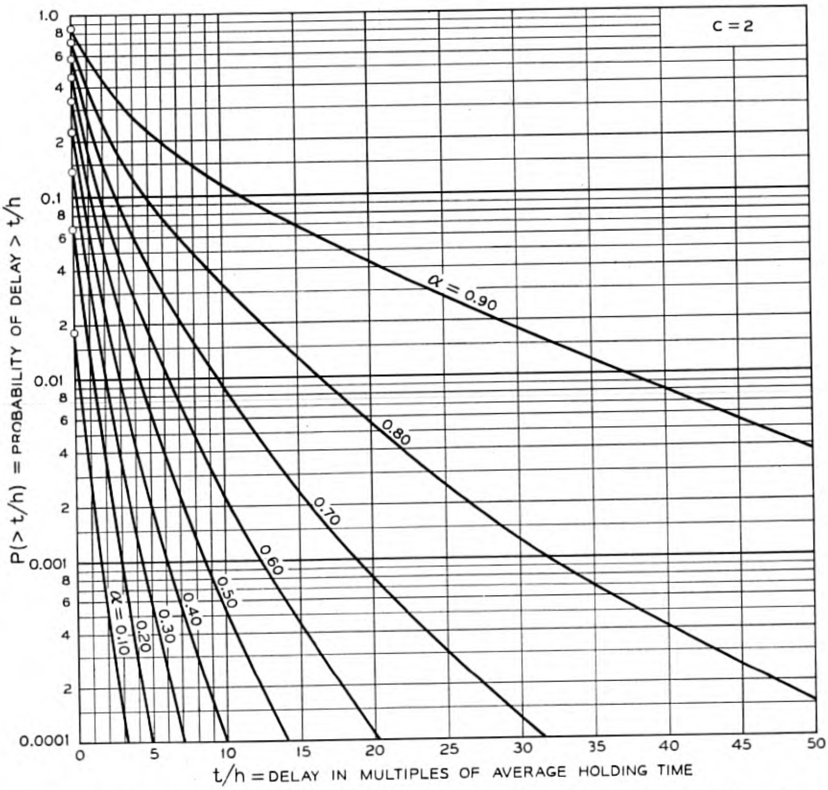


Fig. 9 — Delayed traffic served in random order, exponential holding times, $c = 2$.

scheme, delay results are obtained nearly as good as those realized by the provision of 20 storage circuits (Curve B).

WORKING CURVES

The adequacy of the Riordan theory when delayed exponential calls are served at random is believed to have been established and that it may be used with confidence to solve those practical problems where the underlying assumptions are well satisfied.

For working purposes, curves showing distributions of delays expected for occupancies up to $\alpha = 0.90$ and for group sizes of $c = 1, 2, 3, 4, 5, 6, 8, 10, 20, 50$ and 100, are shown in Figs. 8 to 18. These are plotted in the customary fashion with delay in multiples of average holding time

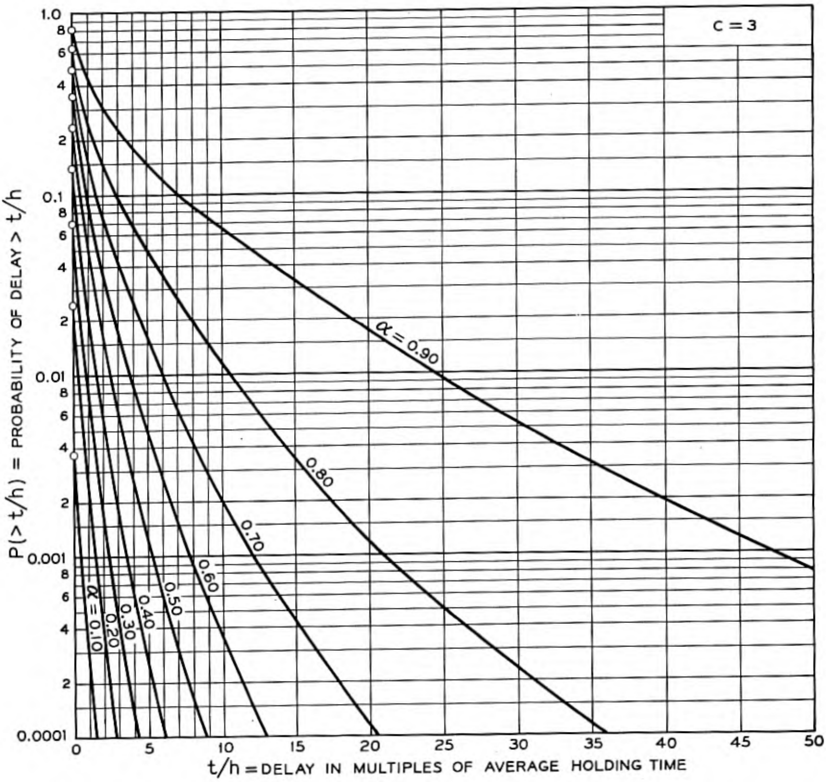


Fig. 10 — Delayed traffic served in random order, exponential holding times, $c = 3$.

as abscissa, and $P(>t/h)$, the probability of a random call meeting a delay greater than t/h, as ordinate.

Estimates of average delays, $\bar{t}$ (which are the same for queued and random service), are also commonly desired, and these are shown in Fig. 19. They are calculated from the equation

$$\bar{t}/h = P(>0)/(c - a) \tag{2}$$

If one wishes instead the average delay, $\bar{\bar{t}}$, *on calls delayed*, it may be obtained from

$$\bar{\bar{t}}/h = \frac{\bar{t}/h}{P(>0)} = \frac{1}{c - a} \tag{3}$$

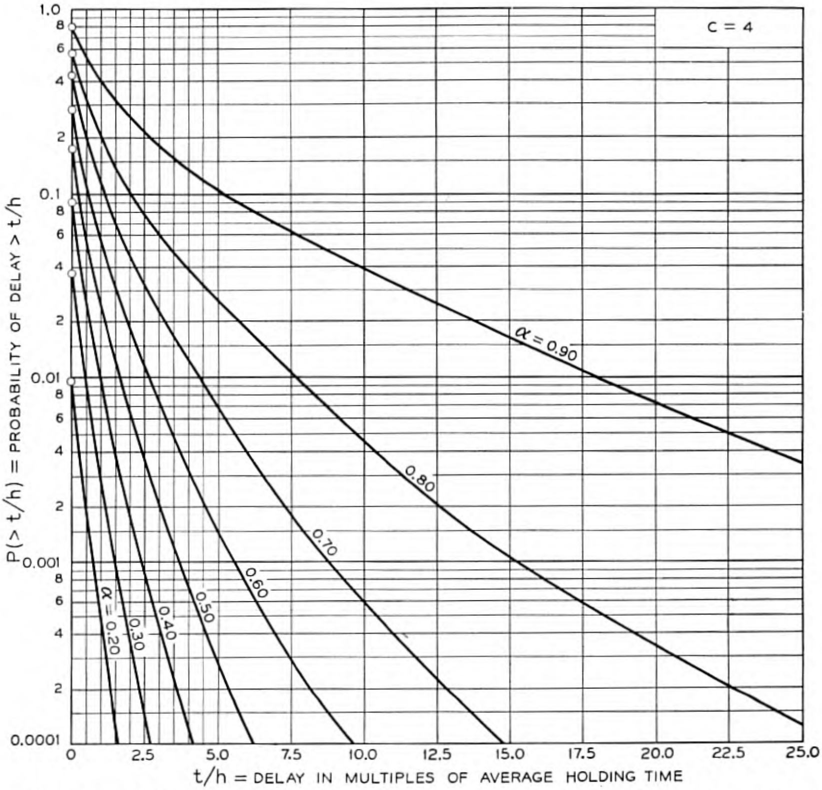


Fig. 11 — Delayed traffic served in random order, exponential holding times, $c = 4$.

ILLUSTRATIVE EXAMPLES

Example No. 1

A 20-trunk toll route carrying exponentially distributed holding time calls with average length of 5 minutes, is loaded with 16.0 erlangs of traffic, and any calls delayed will be served in random order. What per cent of all calls will be delayed? What per cent will be delayed more than 5 minutes? More than 10 minutes?

Solution. Enter Fig. 16 (the $c = 20$ chart) and read on the $\alpha = \frac{16}{20} =$ 0.80 occupancy curve. At $t/h = 0$, the per cent of all calls delayed is found to be 26 per cent. At $t/h = 1$, the calls having delays exceeding 1 holding time, or 5 minutes, are 1.2 per cent, and at $t/h = 2$, the calls with delays exceeding 10 minutes are 0.2 per cent.

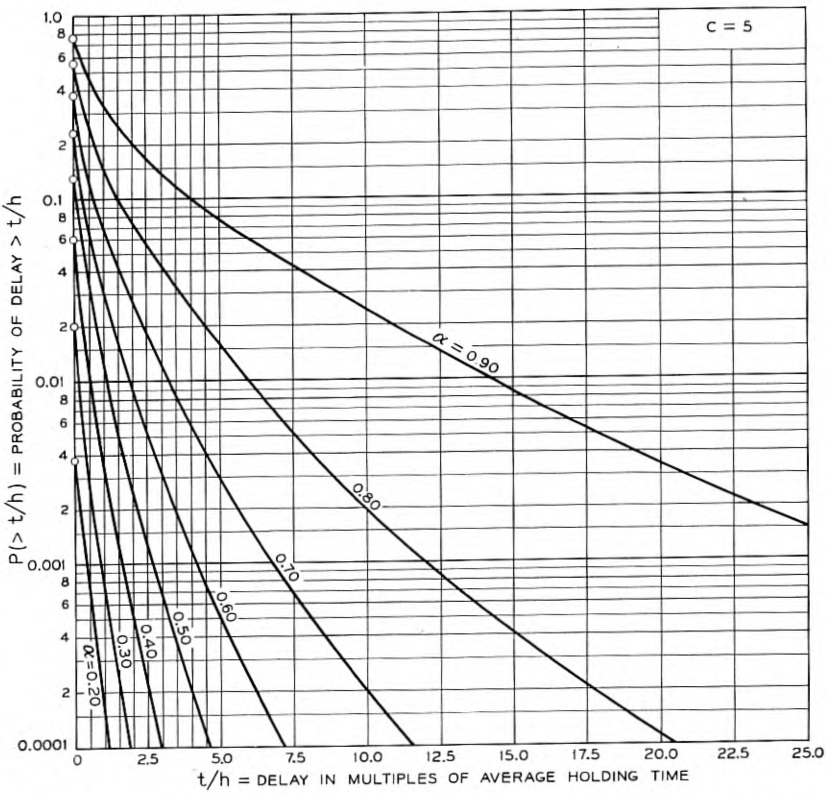


Fig. 12 — Delayed traffic served in random order, exponential holding times, $c = 5$.

Example No. 2

How many operators will be required at a department store telephone order desk to handle 225 calls per hour with an average delay not longer than half a minute, and with no more than 20 per cent of the calls delayed over 1 minute? Assume the average operator work time per call is 100 seconds, and waiting calls are handled in indiscriminate order.

Solution. The load to be carried is $a = (225)(100)/3600 = 6.25$ erlangs. The average delay, $\bar{t}$, is not to exceed $30/100 = 0.3$ holding time. Reading on Fig. 19, opposite an ordinate of 0.3 we select several trial values of trunks (operators) c, versus occupancy α, and form Table I, calculating the last column from the first two:

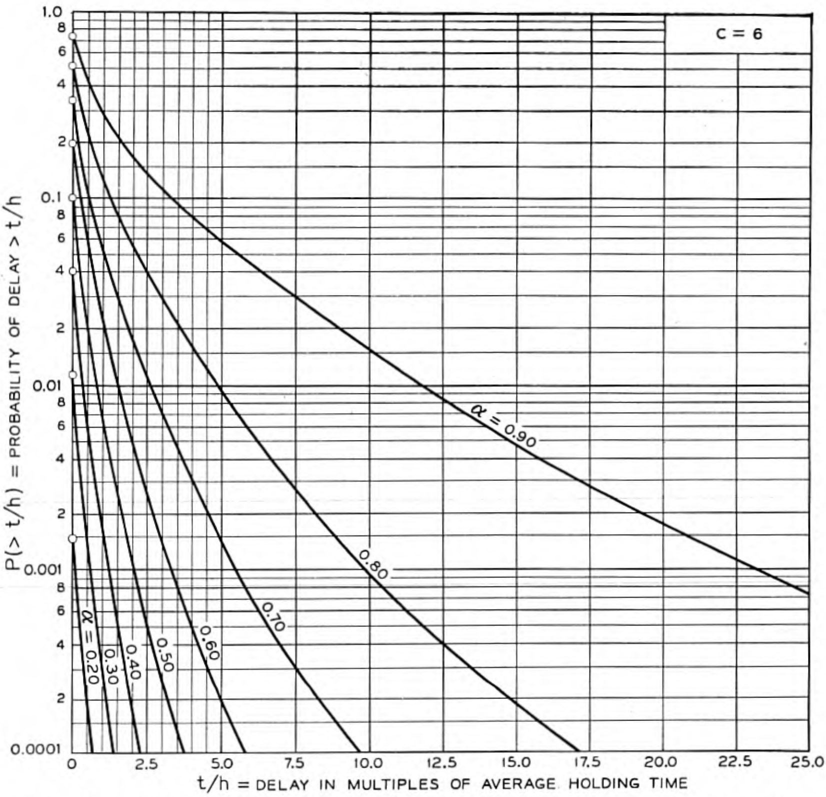


Fig. 13 — Delayed traffic served in random order, exponential holding times, $c = 6$.

To carry the 6.25 erlangs of traffic and meet the average delay requirement we see that 8 operators will be needed. Will 8 operators also fulfill the no more than 20 per cent delay over 1 minute requirement? Enter Fig. 14 (the $c = 8$ chart) with an occupancy of $\alpha = 6.25/8 = 0.78$. The per cent of calls exceeding a delay of $60/100 = 0.6$ holding time is about 12 per cent. A provision of 8 operators satisfies both requirements.

TABLE I

c	α	$a = c\alpha$
7	0.78	5.46
8	0.81	6.48
9	0.82	7.38

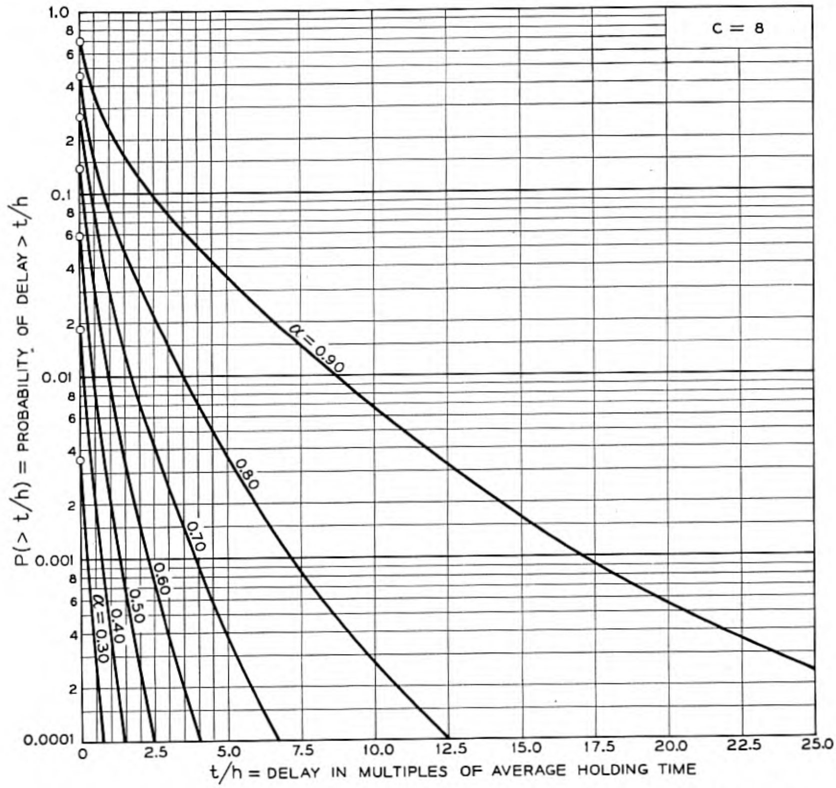


Fig. 14 — Delayed traffic served in random order, exponential holding times, $c = 8$.

Example No. 3

Suppose in Example 2, the second requirement had been that no more than one of 1000 customers should be required to wait over 3 minutes. Would 8 operators then suffice?

Solution. Reading on Fig. 14, with $\alpha = 0.78$ and $t/h = 180/100 = 1.8$, $P(>t/h) = 0.027$. Thus 27 in 1000 calls would be expected to experience delays over 3 minutes, and therefore more than 8 operators will be required. Consulting the $c = 10$ curves of Fig. 15, we find that with $\alpha = 0.625$, and $t/h = 1.8$, $P(>3 \text{ minutes delay}) = 0.0012$ which closely meets the one in a thousand requirement. Ten operators would then be needed; and this would, of course, (from Fig. 19) reduce the average

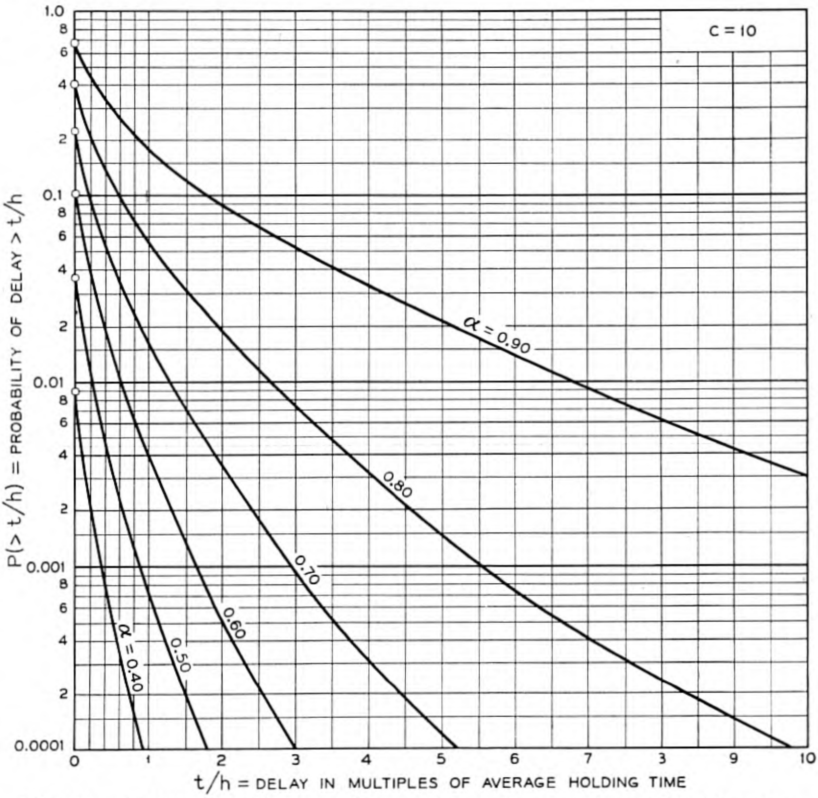


Fig. 15 — Delayed traffic served in random order, exponential holding times, $c = 10$.

delay on all calls to 0.035 (100) = 3.5 seconds, an improvement in this characteristic of 7 to 1 over the 8 operator service.*

Example No. 4

How much improvement in the delay service would be obtained in Examples 2 and 3 by purchasing storing or gating equipment which would substantially insure calls being handled in order of arrival?

Solution. With 8 operators working at an occupancy of 0.78, the pro-

* Had some number of operators been required other than those for which working charts, Figs. 8 to 18, are supplied, intermediate values could be obtained by graphical interpolation, or better still by employing the basic Riordan chart, Fig. 20, combined with $P(>0)$ found on Fig. 21, to obtain delay versus load for any desired number of paths or facilities. This latter process is described in the Appendix.

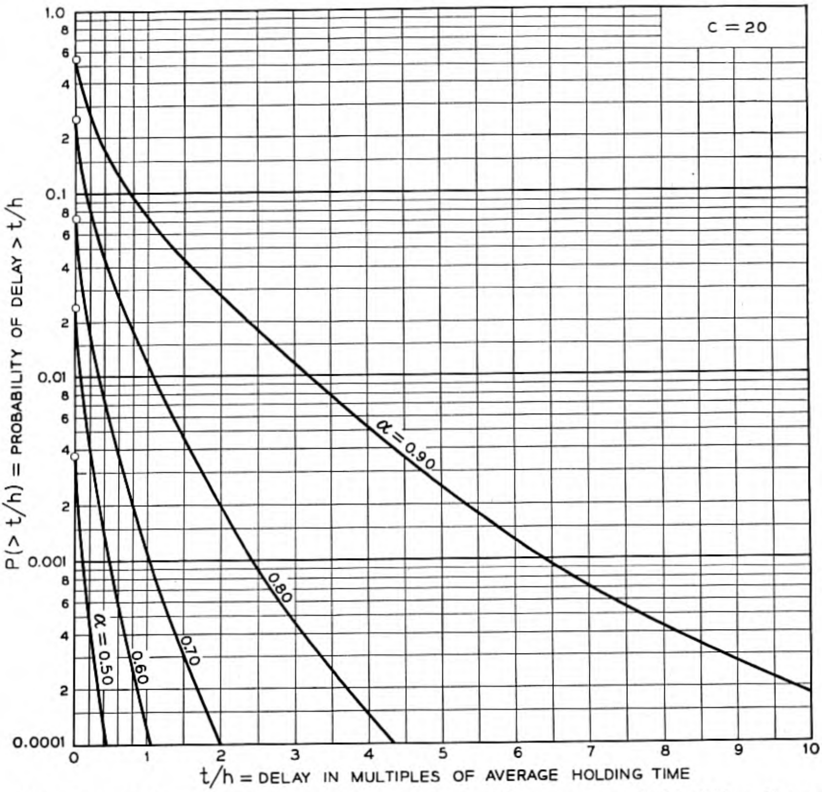


Fig. 16 — Delayed traffic served in random order, exponential holding times, $c = 20$.

portion of calls delayed is found to be $P(>0) = 0.41$ (Fig. 14). The probabilities of exceeding delays of $t/h = 0.6$ and 1.8 holding times are calculated for calls served in order of arrival by equation (1), in the following table:

t (Min.)	t/h	Queued $P(>0) = P(>0)e^{-(c-a)t/h}$	Random Handling
1	0.6	0.143	0.12
3	1.8	0.019	0.027

Comparing the queued and random handling of delayed calls one finds the perhaps unexpected result that with random handling some 2 per cent *fewer* calls are delayed longer than 1 minute than if perfect queueing had been present. This is due to the characteristic shapes of the two types

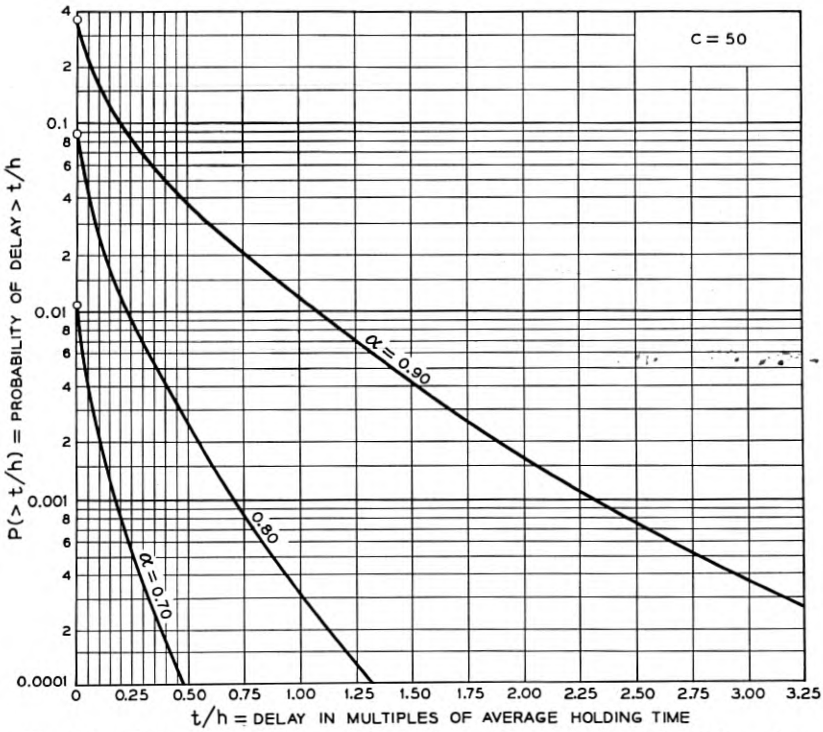


Fig. 17 — Delayed traffic served in random order, exponential holding times, $c = 50$.

of delay distributions, random handling producing more quite short and very long delays than does queueing. When a criterion of service is set at a relatively short delay, one may often expect it to be met more easily by not providing storing or gating circuits. On the other hand a criterion of service based on relatively long delays can nearly always be more readily met by the use of devices insuring partial or total queueing. In the example above the per cent of calls delayed longer than 3 minutes would be cut by a third through the use of queueing devices.

Example No. 5

Automobiles are parked in a large area adjacent to a State Fair grounds. There is one main exit through which two cars can pass at the same time. Upon leaving, drivers pay according to their parking time; and it requires, on the average, 20 seconds to complete the payment. If cars wish to leave during the afternoon busy period at a rate of 5.4 per

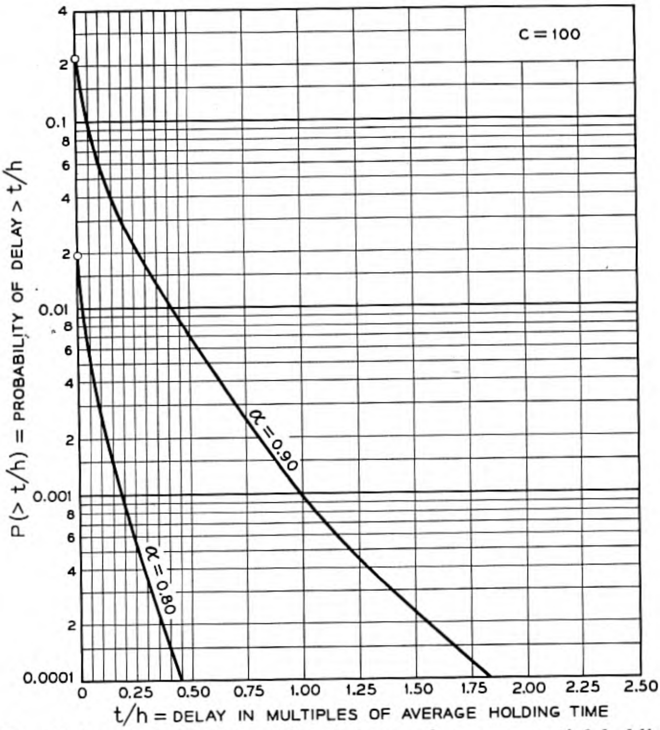


Fig. 18 — Delayed traffic served in random order, exponential holding times, $c = 100$.

minute, what per cent of the cars will be delayed more than 5 minutes? What will be the average delay for all cars?

Solution. Assume there is no traffic supervision and cars converge on the gate from many directions. Service in random order (or worse) among those delayed might then be approximated. Also the distribution of times for calculating and collecting the charge might be roughly exponential. We have then,

$$c = 2 \text{ paths}$$

$$\alpha = (5.4)(20)/(60)(2) = 0.90$$

$$t/h = \frac{5(60)}{20} = 15$$

Enter Fig. 9 at $t/h = 15$, read to the $\alpha = 0.90$ curve, opposite which find $P = 0.069$. Hence 7 per cent of the cars would be expected to have to wait 5 minutes or more. To obtain the average delay for all cars, enter

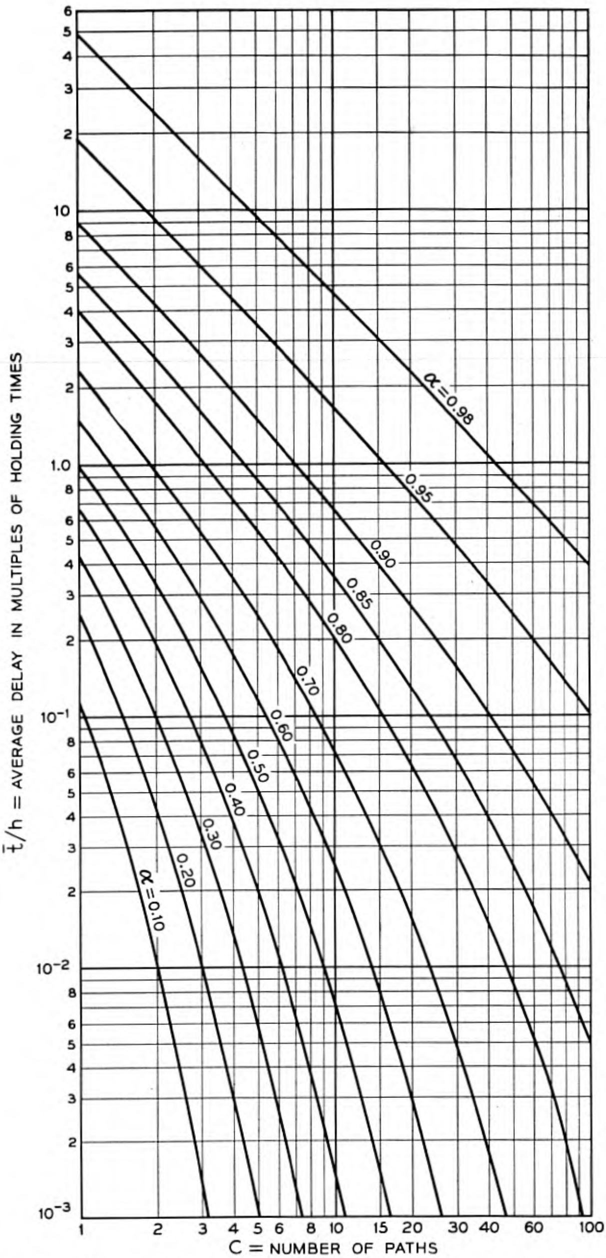


Fig. 19 — Average delay on all calls, exponential holding times.

Fig. 19 with the abscissa of 2 paths, read to the $\alpha = 0.90$ curve and find the average delay = 4.25 average holding times = 85 seconds. Or, one may obtain the same answer by substituting in equation (2),

$$\bar{t} = P(>0)h/(c - a) \doteq (0.85)(20)/(2 - 1.80) = 85 \text{ seconds.}$$

Example No. 6

Suppose in Example 5, an efficient corps of police had been directing traffic toward the exit so that good queueing was maintained. What per cent of the cars would then be delayed more than 5 minutes?

Solution. We may now refer to other published delay curves for queued operation*, or, more generally, calculate the well known equation (1). In the present case we can read the answer from the "queued" curve of Fig. 1 as 4.2 per cent. Thus serving customers in the order of arrival nearly halves the occurrence of very long delays. (Note that the average delay for all cars remains unchanged at 85 seconds.) If a partial queueing were maintained the improvement would be intermediate, perhaps comparable with one of the "limited queueing" distributions shown on Fig. 7.

The author is indebted to Miss C. A. Lennon for constructing the working delay curves, and to Misses C. J. Durnan and J. C. McNulta for performing the throwdown checks.

Appendix

CALCULATION OF DELAY VALUES NOT FOUND ON THE WORKING CURVES OF FIGS. 8–18, FOR DELAYED EXPONENTIAL CALLS SERVED IN RANDOM ORDER

A master chart, Fig. 20, reproduced from Riordan†, gives in condensed form the *proportion* $F(u)$ *of delayed calls delayed* longer than u, where the delay is now expressed in multiples of the h/c ($u = ct/h$), and

c = number of paths (trunks, operators, etc.) provided

h = average holding time

t = delay time

To obtain the probability $P(>t/h)$ of any call being delayed longer than

* E. C. Molina, Ibid.

† Loc. cit.

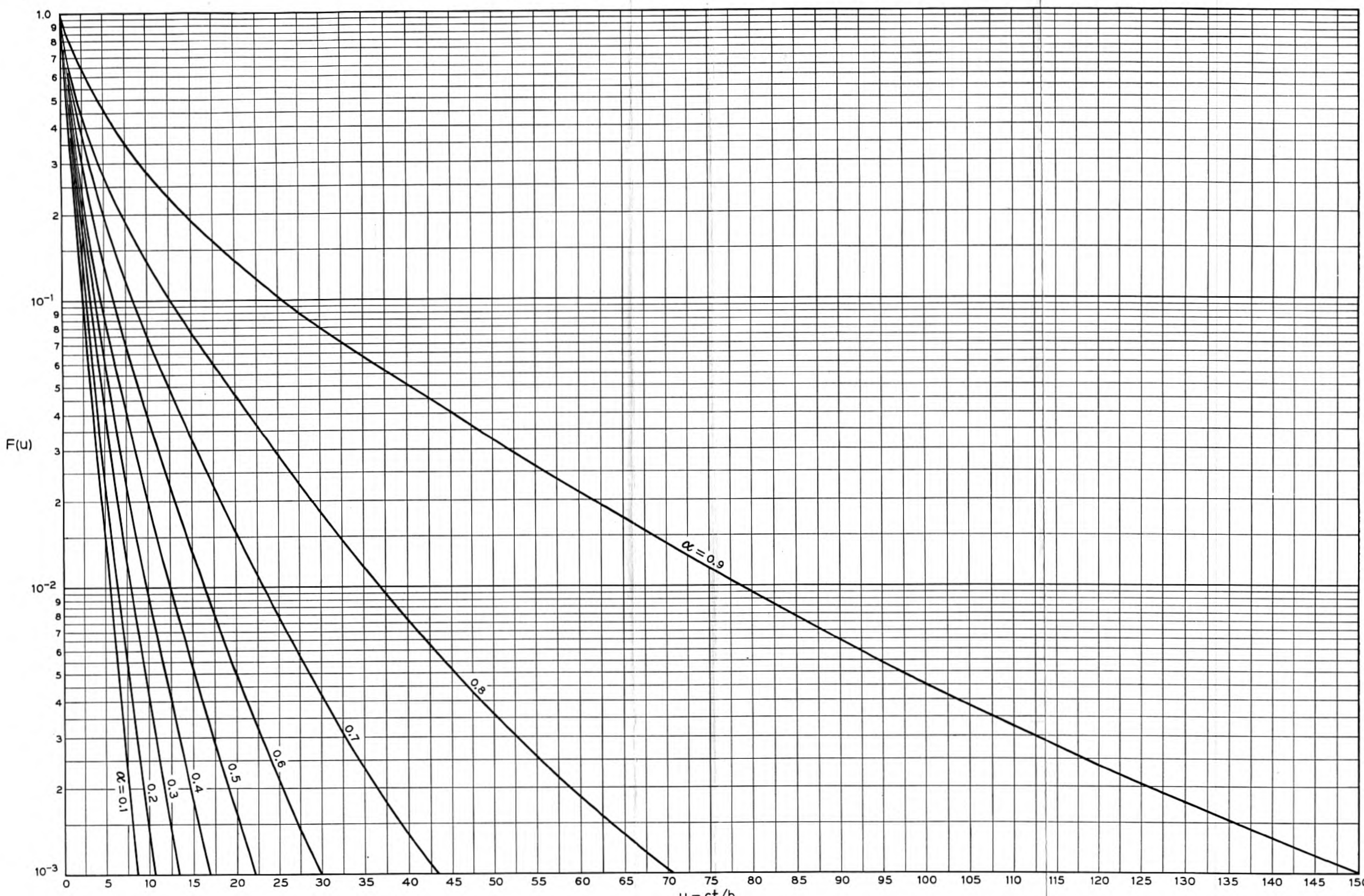


Fig. 20 — General relationship, probability of delays exceeded versus load on paths, with delayed expontial calls served in random order.

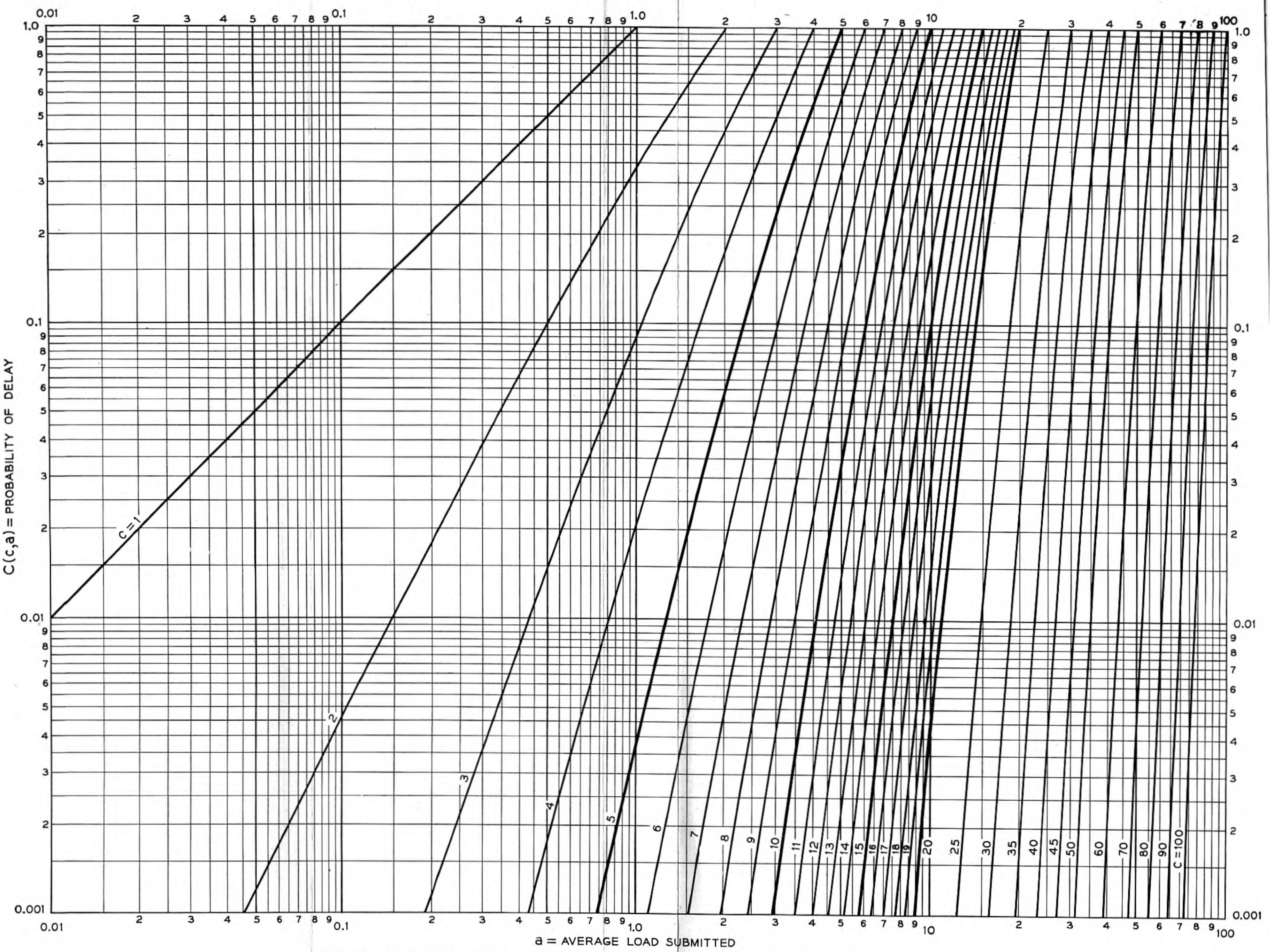


Fig. 21 — Probability of delay for exponential holding time calls handled in a "delayed" basis.

t/h, we have

$$P(>t/h) = P(>0)\ F(u) = C(c,\ a)\ F(u) \tag{4}$$

Values of $P(>0) = C(c, a)$ are given for a wide range of a and c in Fig. 21. The application of equation (4) is quite simple.

Illustration 1. Suppose it is desired to obtain the probability of a call being delayed more than 3 holding times on a 10 trunk group without storage or gating circuits, and which carries $a = 9$ erlangs. Here $t/h = 3.0$, $c = 10$, $\alpha = 0.9$. Then $u = ct/h = 30$, and reading on Fig. 20 with this value of u, and $\alpha = 0.9$, we find $F(u) = 0.080$. Fig. 21 provides $C(c, a) = 0.67$ for $a = 9$ and $c = 10$. Substituting in equation (4),

$$P(>3 \text{ hold times}) = 0.67\ (0.080) = 0.053,$$

which checks the value read directly from the $c = 10$ curves of Fig. 15.

Illustration 2. With an occupancy of $\alpha = 0.65$ on 15 paths what is the probability of meeting a delay greater than one holding time when delayed calls are served in random order? Calculate $u = ct/h = 15$. Enter with this abscissa on Fig. 20, and interpolating between the $\alpha = 0.6$ and 0.7 curves, read $F(u) = 0.022$. Fig. 21 shows for $a = 0.65(15) = 9.75$ and $c = 15$, $C(c, a) = 0.085$. Hence

$$P(>1 \text{ hold time}) = 0.085(0.022) = 0.0019.$$

Magnetic Resonance

PART II — MAGNETIC RESONANCE OF ELECTRONS

By KARL K. DARROW

(Manuscript received December 24, 1952)

Magnetic resonance of electrons is the analogue of magnetic resonance of nuclei, treated in the first part of this article. Though the analogy is close and the fundamental laws are identical, the two topics are remarkably different in detail. Though electrons are the commonest of particles, they display magnetic resonance only in somewhat exceptional cases. In many free atoms and most solid and liquid substances, magnetic resonance is suppressed by what is known as the "anti-parallel coupling" of electrons two by two. The exceptional cases are those of certain free atoms, ferromagnetic substances, and a restricted class of strongly paramagnetic substances; the resonance has also been observed very lately for the conduction electrons in metals. In the cases in which it does occur, resonance is likely to occur at a frequency or frequencies very different from that which the elementary theory predicts. This is sometimes because of the orbital motions of the electrons, oftener mainly because of the electric and magnetic fields existing in solids, and the deviations of the observed cases from the ideal case shed light upon these fields.

The subject of these pages is the magnetic resonance of electrons — "electron resonance" for short. Electrons being everywhere, one might expect it to be found in every substance; but for a fundamental reason it is a rare phenomenon, and this magnifies its interest. Those who search the literature for it under this its proper name will seldom find it, for it is frequently called "paramagnetic resonance" or, in appropriate cases, "ferromagnetic resonance." These are lengthy names which tend to veil the similarities between electron resonance and nuclear resonance, which latter was the theme of Part I of this article (in the January issue of this JOURNAL). I will introduce electron resonance by making use of all these similarities.

Magnetic resonance in general is due directly to the magnetism of subatomic particles: nuclei and electrons. These, apart from the nuclei that are non-magnetic, may be visualized as minuscule barmagnets. The laws of resonance are determined by the fact that in a steady magnetic field, the magnetic moments of these particles may not point in any and every direction: instead, they are constrained to a finite and small number of what are called "permitted orientations." To each of these corresponds a special value of the energy of the little magnet in the field: thus the energy also is constrained to a finite and small number of "permitted" values. These are often called "Zeeman levels" or just "levels"; and the word "level" should be well known to those who are going to delve into the literature.

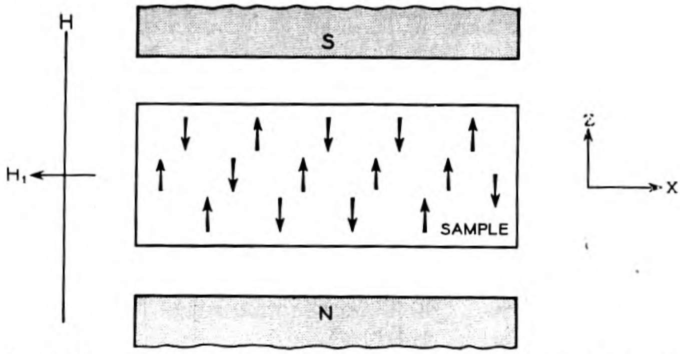


Fig. 1. — Scheme of the apparatus for observing magnetic resonance. The high-frequency circuits are omitted. The arrows within the sample may be taken as portraying the magnetic moments of either protons or electrons: their orientations are as given by the *old* quantum-theory.

Consider two orientations or levels of different energy-values. It will take *work* to turn the tiny magnet from the one of lesser energy to the one of greater energy. Magnetic resonance — and now I ought perhaps to speak specifically of magnetic resonance *absorption* — is such a turning. The agent of the turning and the source of the work is an alternating or oscillating magnetic field. The simplest cases are those in which the particle in question has only two permitted orientations. Many nuclei, among them the proton, belong to this class, and the electron belongs to it also. It is the analogy between proton and electron which I will develop.

Fig. 1 of this part is also Fig. 1 of Part I. The central rectangle depicts the *sample*, which for the study of proton resonance must be hydrogen or a compound thereof. The big arrow on the left represents a big magnetic field, of the order of several thousand gauss, which pervades the sample; it is vertical and its strength is denoted by H. This is the field with respect to which the protons are oriented. These magnetic particles are represented by small arrows within the rectangle, the point of each arrow corresponding to the north pole of the corresponding proton. Slightly more than half of them are pointed in what I call the "up" orientation, which is that of lesser energy. The rest are pointed in the "down" orientation, that of greater energy. Magnetic resonance absorption of protons is the turning of "up" protons into the "down" direction.

The field which does the turning is an oscillating magnetic field with frequency (denoted by ν) in the radio-frequency range. It is horizontal, thus at right angles to the big field. It is produced either in a solenoid (the usual method for nuclear resonance) or in a resonant cavity (the usual scheme for electronic resonance) which encloses the sample but in Fig. 1 is left to the imagination of the reader.

Magnetic resonance occurs when the quantum-energy $h\nu$ of the oscillating field is equal to the work required to turn the proton from the up orientation to the down one:

$$h\nu = \text{work of turning} \tag{1}$$

h standing for Planck's constant. In Part I it was shown that the "work of turning" or energy-difference between the two orientations is equal to $2\mu_p H$: here μ_p stands for the magnetic moment of the proton, soon to be more carefully defined. Thus:

$$h\nu = 2\mu_p H \tag{2}$$

When ν and H are related by this equation one finds *proton resonance absorption*, which manifests itself by a splendid peak in the curve of absorption *versus* H for constant ν or the curve of absorption *versus* ν for constant H. For the frequency 42.6 megacycles the peak is found at $H = 10{,}000$ gauss.

To arrive at the basic formula for *electron resonance* we simply take (2) and substitute into it μ_e, the magnetic moment of the electron, for μ_p:

$$h\nu = 2\mu_e H \tag{3}$$

The magnetic moment of the electron is about 660 times that of the proton. Therefore if one works with such a field strength as brings the proton resonance into the radio frequency range, the electron resonance is to be sought in the microwave range. One might think that now I have said all that there is to be said about electron resonance; but this is only the beginning.

Much was said in Part I about the magnetic resonance of nuclei having more than two permitted orientations. We may seem to be wandering off the course if we revert to these, but this case is very pertinent.

There are nuclei with three, four, . . . up to ten or maybe more allowed orientations. One would expect them to display a multitude of peaks; but there is never more than one. This is for two reasons, which I give after introducing the symbol $(2I + 1)$ for the number of orientations. First, it is impossible to turn a nucleus from any orientation to any other *except* the nearest to the original one. This reduces the number of possible peaks to one fewer than the number of orientations. But second, all of these $2I$ possible peaks are of the same frequency for given H, or at the same field strength for given ν, so that they all coalesce into a single peak.

The formula for this apparent single peak which is strictly $2I$ coincident peaks has been derived in Part I, and this is it:

$$h\nu = (\mu/I)H \tag{4}$$

Now it is necessary to interpret I and μ; and the interpretation is different according as one uses the old quantum theory or the new quantum mechanics. The old quantum theory deals more simply with these problems, and would be preferable if this field could be isolated from all the rest of physics; but the new quantum mechanics is worth the extra trouble that it causes.

In the old quantum theory, there are two definitions of I that reduce to the same thing. First, I is the angular momentum of the nucleus in terms of the unit $h/2\pi$; that is to say, the angular momentum of the nucleus is $Ih/2\pi$. Second, $Ih/2\pi$ is the maximum possible projection, upon the field-direction, of the angular momentum of the nucleus. This is because, among all of the allowed orientations of the nucleus, the one which is most nearly parallel to the field-direction is *exactly* parallel to the field-direction. So it was shown in Fig. 1.

In the new quantum mechanics, the second of these definitions remains valid and the first does not. This is because the orientation which is most nearly parallel to the field-direction is not exactly parallel thereto. It is inclined, in fact, to the field-direction by the angle *arc cos* $I/\sqrt{I(I + I)}$, and the angular momentum of the nucleus is $\sqrt{I(I + 1)}(h/2\pi)$.

Thus there is one definition of I which is valid under both theories, and that is, that I is the maximum possible projection upon the field-direction, of the angular momentum of the nucleus in terms of the unit $h/2\pi$. Similarly it is always correct to say that μ is the maximum possible projection upon the field-direction, of the magnetic moment of the nucleus. But since these phrases are intolerably long, one avoids them by saying that I is the spin and μ the magnetic moment of the nucleus. In this sense, which to the users of quantum mechanics is a distorted one, the words "magnetic moment" shall be used hereafter.

The spin of the proton is 1/2, and so is the spin of the electron. Equation (4) degenerates into (3) for the electron and into (2) for the proton. These we had already; what was then the point of introducing here the general case?

Well, the point is that two, or three, or several electrons may collaborate in what is known as "parallel coupling," though in the new quantum theory it is not quite parallel. They behave as though they formed a rigid unit, of which the spin is the sum of their spins and the magnetic moment is the sum of their magnetic moments. Thus if there are N of these electrons welded together (metaphorically speaking) it comes to the same as though there were a single particle of spin N times 1/2 and magnetic moment N times μ_e. On putting these values of I and μ into equation (4) we find ourselves right back at equation (3), which is that for the individual electron. There is a single peak of magnetic resonance composed of N coinciding peaks, and it is just where the peak for a single electron would be. Thus in the ideal case, N electrons coupled parallel behave just like one electron by itself.

Such a conclusion may seem hardly worth the trouble of arriving at it; but note the stipulation "in the ideal case." This refers to what has been tacitly but obviously assumed till now, to wit, that no force acts upon the electronic magnet except the big field H. But there are also what I will call "local forces," forces due to fields within the sample arising from other particles in the sample. These forces may, and they often do, separate the N peaks which in the ideal case coincide. Often one finds a flock of resonance lines where, or near where, there should be only one; and if this is the explanation (which is not always the case, for there are other causes of "splitting") then the number of lines in the flock is the number of electrons coupled parallel.

This illustrates one of the great contrasts between the electronic resonance and the nuclear. Nuclear resonance is a "textbook phenomenon." The ideal case and the actual case are close together; the deviations due to the local fields are neither trivial nor useless, but they are not large enough to distort the simple laws, and it is quite permissible to leave them out of a first presentation. But the phenomenon of electronic resonance is liable to be distorted almost beyond recognition; and if one were to present only the cases in which the local fields are negligible in effect, one's story would be relatively short and it would be grossly inadequate. But here the physicist, true to the tradition of his science, turns hindrance into help, and analyzes the distortions for the knowledge they are capable of giving about the fields prevailing in the sample. Thus whereas nuclear resonance is largely used for getting light on nuclei, the electronic resonance is largely studied for the information that it yields about the solid state.

Another of the great contrasts is due to what are called the "anti-parallel couplings" between electrons. Generally speaking (and this means: conceding an occasional exception) any type of nucleus of non-zero magnetic moment will display a detectable resonance if there are enough of them in the sample. Were this so with the electron, every substance whatsoever would display electron resonance. Experience shows that electron resonance is rare, usually conspicuous by its absence.

This is because electrons may, and not only may but usually do, pair off with one another in such a manner that the spin of such an "anti-parallel" pair is zero and so is the magnetic moment. There is no resonance for such a pair; and the customary absence of electron resonance signifies that in most solids, all the electrons are joined two by two into antiparallel pairs (this was known before magnetic resonance was first produced). I will call such electrons "compensated"; in this language, the substances in which magnetic resonance is to be sought for are those with uncompensated electrons. Mostly these belong to one or the other of two classes: the ferromagnetic bodies including the anti-ferromagnetic, and the "strongly paramagnetic salts." But there are a few other cases, and among these are those which are closest to the (unattainable) ideal of the perfectly free electron subjected.

THE NEARLY IDEAL CASES

Nearest of all to the ideal case are presumably the atoms which contain uncompensated electrons and are available for study by the molecular-beam method. Outstanding among these is the hydrogen atom, whose single electron must remain uncompensated because there is no other in the atom. About or quite as good are the atoms of sodium, potassium, and the other alkali metals, each of which contains a single uncompensated electron not to speak of several which are compensated. Moreover, these atoms are normally in a "ground state" in which the uncompensated electron has no orbital angular momentum. This hints at a complexity which is not always without influence on electron resonance, and must be mentioned here at the price of a detour.

Going back to ancient theory, let us imagine an electron revolving with frequency f in a circular orbit of radius r. It is equivalent to a current ef running continuously in the circular loop. According to the old theorem of Ampere, its magnetic moment is equal to the area of the circle multiplied by the current-strength; but the current-strength is to be expressed in *electromagnetic* units, so that the magnetic moment μ equals $(e/c)f\pi r^2$. The angular momentum p is mr times the speed of the electron, and therefore equals $2\pi mr^2f$. For the ratio of the two we find:

$$\mu/p = e/2mc \tag{5}$$

This is what has lately been miscalled the "gyromagnetic ratio," a name which was originally applied and ought still to be applied to its reciprocal. It would be good to follow Gorter's suggestion of calling it the "magneto-gyric ratio."

I now state equation (5) in another fashion so as to introduce a symbol which is really a word, and is *the* technical word of this field of physics: it ought to be a word all spelled out, but it is just the letter g.

$$(\mu_{\text{orb}}/p_{\text{orb}}) = g(e/2mc), \qquad g = 1 \tag{6}$$

Thus g is the ratio of magnetic moment to angular momentum given in terms of $e/2mc$ as unit, and *its value for the orbital motion of an electron is one*. Note also that though we have arrived at (6) in a very old-fashioned way, it is one of the results that have stood firm through all the mutations of quantum theory.

The study of what are known as "multiplets" in optical spectra led some thirty years ago to the conclusion that for the spin of the electron the magneto-gyric ratio is such that $g = 2$:

$$(\mu_{\text{spin}}/p_{\text{spin}}) = g(e/2mc), \qquad g = 2 \tag{7}$$

This belief was substantiated by the "Dirac theory," and was not upset until measurements were made of the magnetic resonance of electrons in atoms by the molecular-beam method. The first such measurements were made upon atoms containing uncompensated electrons which had orbital motion as well as spin. I pass them over, and come direct to the most recent experiments on hydrogen atoms in their ground state, where there is no orbital motion of the electron to complicate matters. These are so recent that they came into print as these words were being written.

The hydrogen atom is a good example to take, not only for the reasons that I have given already, but also because it may be compared with the hydrogen molecule H_2. The two electrons of the hydrogen *molecule* compensate one another, and there is no electron resonance. The two nuclei — protons — of the molecule compensate one another in some of the molecules, enter into the parallel coupling in others. There are always some of these last in a beam of hydrogen molecules, and they produce the proton resonance of which so much was said in Part I. The *atoms* produce the electron resonance.

Look now again at equation (4), and remember that p is $Ih/2\pi$ — and remember that p is to be interpreted as the maximum permitted component, along the field-direction, of the angular momentum.

Consider now the experimenter with molecular beams of hydrogen *molecules* and hydrogen *atoms* at his disposal. In a magnetic field of field strength H he finds the proton resonance of the former at frequency ν_p, and ascertains (μ/I) of the proton by putting his data into equation (4):

$$(\mu/I)_p = h\nu_p/H \tag{8}$$

In the same field he finds the electron resonance of the latter at frequency ν_e, and ascertains (μ/I) of the electron similarly:

$$(\mu/I)_e = h\nu_e/H \tag{9}$$

Now he has both values; but the accuracy of both is contingent on the accuracy of the measurement of H, and this is not so good as he desires. However he can dispense with the measurement of H at a price — the price of getting his value of the magnetic moment of the electron in terms of units other than c.g.s. units. This is not a great sacrifice; Nature does not share our affection for c.g.s. units; there are others which are more suitable to the enterprises of the theorist.

If we divide (8) into (9) we get rid of both H and h. This means that if the experimenter measures ν_p and ν_e in one and the same applied field, he can evaluate (μ/I) for the electron in terms of (μ/I) for the proton without bothering about the values of H and h. Since I is the same for both particles, he obtains the ratio of the magnetic moments of electron and proton. The value of this ratio would be precious in itself, even if one had not the faintest idea of the value of either moment in c.g.s. units. It is 658.2288 ± 0.0006.

It is also feasible to get the value of (μ/I) for the electron in terms of the "unit" $eh/4\pi mc$. This entity is so important that it has a name of its own: it is called "the Bohr magneton."

There is also a combination of experiments by which $(\mu/I)_e$ may be evaluated in terms of the unit $(eh/4\pi mc)$. This unit is so important that it has a name of its own: it is called "the Bohr magneton." The reader can easily show for himself that (μ/I) in terms of this unit is none other than the quantity g, of which this is a second definition (not identical with that of g in Part I).

The frequency of the proton-resonance, ν_p, is compared in a special experiment with what is known as the "cyclotron frequency," ν_c, of the electron. A free electron, projected at right angles to a magnetic field H, describes a circle in the plane perpendicular to the field. The frequency with which it makes the tour of this circular orbit is given by the equation:*

$$\nu_c = 2(He/4\pi mc) \tag{10}$$

If this frequency is determined in the same field as has served or is to serve for the location of the proton-resonance, we have:

$$(\mu/I)_p = 2(eh/4\pi mc)(\nu_p/\nu_c) \tag{11}$$

and consequently:

$$(\mu/I)_e = 2(\nu_e/\nu_p)(eh/4\pi mc)(\nu_p/\nu_c) \tag{12}$$

So here is the value of (μ/I) for the electron expressed in terms of the Bohr magneton, determinable by measurements on ratios of frequencies only! At this point the reader may well wonder why I did not eliminate ν_p from (12) by simply dividing it out. The reason is that one group of experimenters has determined (ν_e/ν_p) at one fieldstrength and another group of experimenters has determined (ν_p/ν_c) at another fieldstrength, so that ν_p does not have the same value in the two brackets: this is trivial.

The old belief, as I remarked above, was that (μ/I) for the electron amounted to exactly two Bohr magnetons. But the combination of two experiments which I have just so sketchily described has led to the following result for the electron in the hydrogen atom:

$$(\mu/I)_e = g(eh/4\pi mc), \qquad g = 2.002292 \pm 0.000024 \tag{13a}$$

But is this truly the ideal case? Defining the "ideal case" as that of the free electron, remembering that the electron in the hydrogen atom is bound even though lightly bound, and making what is deemed the appropriate correction, one elevates the foregoing value of g by 35 parts in a million, and obtains:

$$\text{Ideal } (\mu/I)_e = g(eh/4\pi mc), \qquad g = 2.002327 \pm 0.000024 \tag{13b}$$

Thus the old belief was wrong by about one part in a thousand. Be it mentioned in passing that the Dirac theory which led to $g = 2$ has been modified in the meantime by what is known as "quantum electrodynamics", which gives a good account of this result.

* To be derived by equating the force $Hv(e/c)$ exerted by the field upon the electron to the "centrifugal force" mv^2/r; here v stands for the speed of the electron and r for the radius of the circle.

Since I is $\frac{1}{2}$ for the electron (as it is for the proton) the magnetic moment of the free electron is:

$$\mu_e = (\tfrac{1}{2})g(e/4\pi mc) = 1.001146 \pm 0.000012 \text{ Bohr magnetons} \quad (14)$$

This is the value which is 658.2288 times the moment of the proton.

Another case very near to the ideal is afforded by the electrons of such atoms as manganese widely dispersed in a phosphorescent solid. Thus, there exists a measurement of g made upon "zinc sulphide phosphor" containing manganese atoms in a concentration of 0.001 per cent. The value is 2.0024 ± 0.0004. It must be said that the resonance in question is complicated both by fine structure and by hyperfine structure, terms to be explained in following sections. It is therefore necessary to use theory to locate, among the complex of peaks, the frequency which corresponds to the appropriate value of g.

Still another case which is close to the ideal is provided by the "F-centres" in colored crystals, mention of which was made in Part I. An F-centre is a cavity in a crystal lattice occupied by a free electron batting around, as I said in Part I, like a wild animal in a cage. Several physicists have found their resonance, present when the crystal is colored and absent when the crystal is bleached. One, who produced the coloration by neutron-bombardment, located the peak at $g = 2.00$. Others report 1.995 ± 0.001.

Still another case which is close to the ideal is afforded by the conduction electrons in a metal. These are so numerous that one might expect that the electron resonance that they produce must be extremely prominent. Yet the first such peak to be observed has been reported only as these lines are being written! The reasons for its inconspicuous character are two: most of the conduction-electrons are coupled anti-parallel, and the skin-effect confines the oscillating field in a conductor to a very narrow region close up against the surface. The second of these hindrances is overcome by using a colloidal dispersion of the metal, of which the spherules are less than 10^{-3} cm in diameter. Data are available (though not yet all in print) for lithium, sodium and potassium. The values of g are within a few promille of 2.000; the differences between these and the "ideal" value are small but not trivial, and in the case of lithium have been explained.

ELECTRON RESONANCE IN PARAMAGNETIC SOLIDS

There are paramagnetic solids that display the electron resonance. A magnificent illustration is shown in Fig. 2, belonging to an organic substance of which the name and the structural diagram are included in the figure. This is one of the strongest and sharpest electron-resonance peaks on record. The g-value is 2.0064 ± 0.0002; it is therefore *almost* an ideal case, but the difference from the ideal value is sure and significant. It would however be misleading to suggest that such a case is typical.

What are called the "strongly paramagnetic salts" form a group with several features in common. They tend to have long names, and they have complex chemical formulae; crystal lattices or at any rate unit-cells which are non-cubic; and atoms some of which belong either to the rare-earth elements or to the elements of the "first transition group," iron or cobalt for instance. These atoms are likely to have two or more uncompensated electrons in parallel coupling. I now recall what was said about such coupled electrons in the introductory passage.

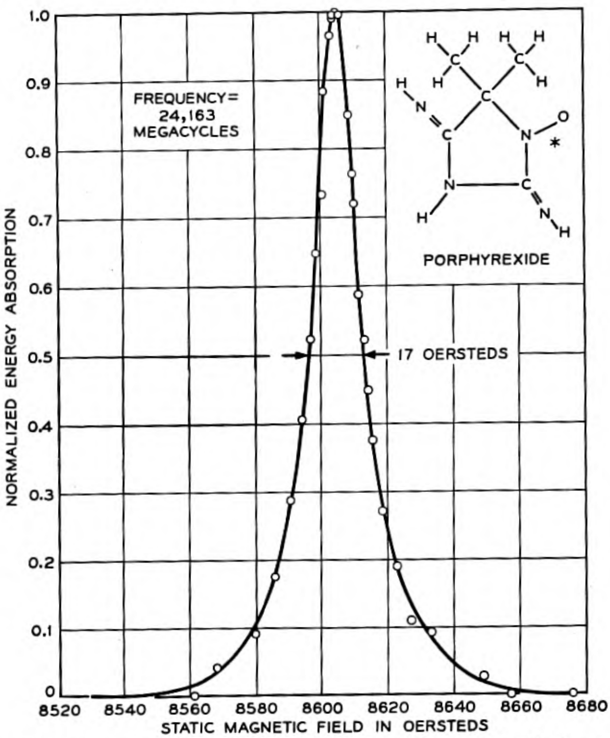


Fig. 2.—Electron resonance of porphyrexide. This is one of the strongest and sharpest peaks of electron-resonance yet observed. The g-value is 2.0064 ± 0.0002, which makes it slightly but significantly different from the ideal case. In the structural diagram, the asterisk signifies a three-electron bond. (A. N. Holden, W. A. Yaeger and F. R. Merritt).

Two or more electrons — N electrons, let me say — may form, in effect, a rigid unit having a total spin $S = N/2$ and a total magnetic moment $N\mu_e$. Such a unit will have $(N + 1)$ allowed orientations in the big magnetic field. These will engender N resonance-peaks. In the ideal case, all of these would have the same frequency $2\mu_e H/h$, and would therefore coalesce into a single peak at the position appropriate to $g = 2.0023$. But in these crystals we are likely to find cases far from ideal, because of the conjoined influence of two factors. These are the presence of *orbital* motions of the electrons, and the presence of a big electric field within the crystal.

Were the atoms in question free, we could allow for the orbital motions. There would be a single resonance-peak, corresponding to a value of g which could be computed by a formula well known and much used in optical spectroscopy. Incidentally, this formula was used in interpreting the earliest molecular-beam experiments (not here described) that were the first to show that g in the ideal case is not exactly equal to 2.

Now, however, we are dealing with resonating electrons that are in a strong electric field, and moreover, an electric field which is usually unsymmetrical. If the asymmetry is sufficiently great, the orbital motions suffer a singular effect. This effect is known as "quenching." It is impossible to explain and difficult even to describe without invoking quantum mechanics. One may say that the orbital angular momentum is no longer constant in time, and the associated magnetic moment almost but not quite disappears.

The spin survives the quenching: but it would not be right to say that the quenching restores the ideal case. The resonance is affected by what have been called the "remains" of the orbital magnetic moment. These have the following consequences:

(a) The N resonance-peaks, which coincide in the ideal case, may be drawn apart. They then form a group of N separate peaks, which is known as a "fine-structure pattern." The number N tells us the number of electrons coupled parallel in the atom, for these two numbers are the same. Often the number of electrons coupled parallel is known from independent evidence, and in such cases it is confirmed by the number of lines in the fine-structure pattern. Sometimes it is not otherwise known, and in such cases it is identified with the number N.

(b) The value of g corresponding to the centre of the fine-structure pattern may be altered considerably from 2.0023, falling as low as 1.35 or rising as high as 6.5. This is as though a part of the orbital magnetic moment were added to or subtracted from the magnetic moment of the spin.

(c) The value of g may depend upon the orientation of the applied magnetic field with respect to the crystal.

(d) The frequency of the resonance-peak or peaks may not be proportional to H. In fact, it may deviate so far from being proportional to H that extrapolation to $H = 0$ will indicate that even in the absence of an applied magnetic field there would be a separation of the levels. Thus the asymmetric electric field within a strongly paramagnetic crystal may by itself produce the effect, which hitherto we have been ascribing entirely to the applied magnetic field. This is called "zero-field splitting."

One sees only too well that the interior of a strongly paramagnetic salt is no place to look for the ideal case, and that resonance in such a salt is a theme for deep study and not for facile interpretation. As a matter of fact, electron resonance in paramagnetic salts is valued for its contribution to our knowledge of the electric fields in these crystals; which is to say, that it is a part of solid-state physics, the details of which lie beyond the scope of this article.

HYPERFINE STRUCTURE OF ELECTRON RESONANCE

One of the most beautiful phenomena in this province of physics — and, I venture to say, not in this province only but in the whole of physics — is the "hyperfine structure" or "hyperfine splitting" of the electronic resonance. Here we see the spin and the magnetic moment of the nucleus collaborating with those of the electron to produce an exquisite and lucid joint effect. It is still the *electronic* resonance, and must never be confused with the nuclear resonance; but the single resonance-peak of the ideal case is split into a group of peaks, the number of which is determined by the spin of the nucleus.

Fig. 3 relates to neodymium — not however to the metal, but to neodymium atoms in a salt of neodymium, diluted with a salt of another metal so that the neodymium atoms may not influence one another through undue proximity. Neodymium is an element with two "odd" isotopes — that is to say, isotopes of odd mass-number — and several "even" isotopes. The even isotopes have non-magnetic nuclei, and so do not perturb the electron resonance. Each of the two odd isotopes has a nucleus of spin 7/2 and non-zero magnetic moment. Such a nucleus will have eight permitted orientations in the big magnetic field. It will produce a local magnetic field in the region of the resonating electrons, and the strength of this field will depend on the orientation. The resonance-frequency depends on the big field compounded with the local field (we met with instances of this rule in the study of nuclear resonance). Therefore there are eight resonance-peaks for the electrons in the atoms of the isotope 143, and eight more for the electrons in the atoms of the isotope 145. This is the key to the remarkable pattern shown in the curve at the bottom of Fig. 3.

In the middle of the pattern is the stump of a tall peak. This is the unperturbed peak due to the electrons in the atoms of even isotopes, those of which the nuclei have no magnetic moment. Whether it is at the position corresponding to $g = 2.00$ will depend on whether the displacement due to electric fields in the crystalline salt of neodymium, with which these data were obtained, is negligible or is not. Then, there are eight much shorter peaks. These are due to the electrons in the atoms

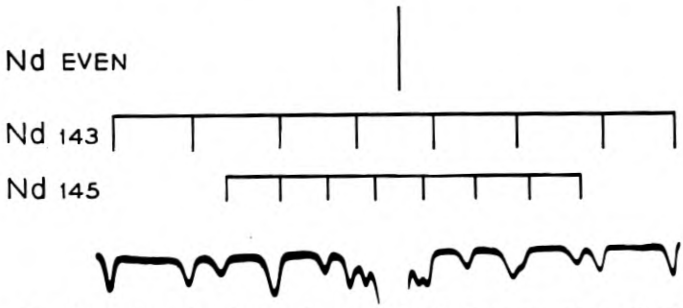


Fig. 3. — Hyperfine-structure pattern of the electron resonance of neodymium in a salt of the metal, showing that the nuclei of each of the odd isotopes of neodymium have eight orientations and therefore a spin of 7/2, and that the even isotopes do not affect the resonance. (Courtesy of B. Bleaney).

of the more abundant of the two odd isotopes. Then, there are eight still shorter peaks (provided we count one which is merged with one of the other group of eight). These are due to the electrons in the atoms of the less abundant of the two odd isotopes. This is beautifully confirmed by the fact that the statures of the two groups of peaks stand to one another in the ratio of the abundances of the two isotopes! Further, the spacings within the two groups stand to one another in the ratio of the magnetic moments of the nuclei of the two isotopes. As for the two combs that stand above the curves, they are markers to identify for the onlooker the members of the two groups of peaks.

Observations on the similar pattern of a (rare) isotope of vanadium — vanadium 50 — have led to the inference that this nucleus possesses a non-zero magnetic moment and a spin equal to 6 (the highest value so far known). This may seem surprising, since I have implied that nuclei of even mass-number have neither spin nor magnetic moment. Vanadium

50 is however a nucleus with an *odd* number of protons and an *odd* number of neutrons. Such nuclei, of which there are only a few stable examples, (in Part I we met with two, the deuteron and N^{14}), are not bound by the usual rule.

FERROMAGNETIC RESONANCE

Ferromagnetic bodies owe their distinctive feature to uncompensated electrons. This suggests that the magnetic resonance of electrons will be discernible in such bodies, and so indeed it is. In this case it is commonly known as "ferromagnetic resonance." However, unless the sample is in the shape of a sphere, the resonance-peak will be found in what appears to be very much the wrong place. This is due to the magnetization of the substance, which produces a remarkable effect upon the location of the resonance. The field strength in the region occupied by the sample, which would be H if the sample were not there, is changed to a very different value; and yet in general it would not be right to take the value of H_i the "internal field strength" and put it in place of H in equation (3). We must understand this effect and make the proper allowance for it before we do anything else with the data (unless, I repeat, we confine ourselves to data obtained with spheres). The effect appears to be beyond the power of "intuition" to conceive, and we must have recourse to the fundamental equations, which describe the precession of the electronic magnets. It will be recalled that in Part I, we looked at nuclear magnetic resonance sometimes as the turning-over of nuclear magnets and sometimes as an outcome of precession. Now we are going to treat the electronic resonance as an outcome of precession.

The fundamental vector equation, which was given in a sort of diluted form as equation (6) of Part I, reads as follows:

$$dp/dt = \mu_e \times H_i \tag{15}$$

Here p and μ_e stand for the angular momentum and the magnetic moment of the electron, and H_i for the field which operates on the electron. We have seen that μ_e/p is written as $ge/2mc$; we denote this quantity by γ; and we give it the minus sign because, for the electron, angular momentum and magnetic moment are antiparallel to one another. Now we have:

$$d\mu_e/dt = -\gamma\mu_e \times H_i \tag{16}$$

This we proceed to write as three scalar equations; but first we replace μ_e by M. This will help to do away with the implication that the magnetic moment varies in magnitude (it is the *direction* that changes with time) and will also convey the plausible suggestion that all of the resonating electrons in the substance are coupled parallel, so that M can signify the magnetization of the substance. We have:

$$\begin{aligned} dM_x/dt &= -\gamma(M_yH_{iz} - M_zH_{iy}) \\ dM_y/dt &= -\gamma(M_zH_{ix} - M_xH_{iz}) \\ dM_z/dt &= -\gamma(M_xH_{iy} - M_yH_{ix}) \end{aligned} \tag{17}$$

Now we are to make the following important substitutions, some of which are approximations.

(1) Presuming that M the magnetization of the substance will not deviate far from the z-direction, we are to write M for M_z.

(2) For H_{iz}, the z-component of the field actually operating upon the electrons, we are to write $(H - N_zM)$. Here H stands as heretofore for the applied field and N_z for the "demagnetizing factor" in the z-direction, which latter is a measure of the strength of the free poles on those surfaces of the sample which face the pole-pieces of the magnet (Fig. 1). Thus $-N_zM$ is the value of the field produced in the substance by these free poles.

(3) For H_{ix} and H_{iy} we are to write $-N_xM_x$ and $-N_yM_y$. This means that whatever applied fields there may be in the x and the y-directions are negligible, and yet the components of magnetization in these directions are not negligible, so that the free poles on the surfaces perpendicular to x and to y respectively are producing the internal fields of which $-N_xM_x$ and $-N_yM_y$ are the strengths.

(4) We are to ignore terms in which the product M_xM_y appears, these being small.

The fourth of these conditions makes dM_z/dt vanish: we are left with only two of the three equations (17), a convenience. Making the substitutions allowed by the first three conditions, we find that the other two assume the forms:

$$\begin{aligned} dM_x/dt &= -\gamma M_y[H - (N_z - N_y)M] \\ dM_y/dt &= -\gamma M_x[-H + (N_z - N_x)M] \end{aligned} \tag{18}$$

Now suppose that M_x and M_y are periodic functions of time, of frequency ν. We write them as $M_x^0 \exp(2\pi i\nu t)$ and $M_y^0 \exp(2\pi i\nu t)$. Substituting into (18), we find:

$$\begin{aligned} 2\pi i\nu M_x^0 + \gamma[H + (N_y - N_z)M]M_y^0 &= 0 \\ -\gamma[H - (N_z - N_x)M]M_x^0 + 2\pi i\nu M_y^0 &= 0 \end{aligned} \tag{19}$$

These two simultaneous equations will be compatible with one another — one might say that they make sense — only if they are ultimately the same equation. The ratio of the coefficients of M_x^0 and M_y^0 in the one must be the same as the ratio of the corresponding coefficients in the other. In words more natural to algebraists, the determinant of the coefficients must vanish. It turns out that this condition determines a specific value of ν, and this value is the resonance-frequency:

$$\nu = (ge/4\pi mc) \sqrt{[H + (N_x - N_z)M][H + (N_y - N_z)M]} \qquad (20)$$

For reasons deriving from the history of celestial mechanics, this procedure is known as "solving the secular equation."

In the most common experimental set-up, the sample is a thin layer parallel to the z-direction — so thin that by comparison with its breadth, the free poles at the surfaces opposite the pole-pieces of the magnet may be regarded as infinitely far away. Under these conditions N_z vanishes, and so does N_x if we lay the x-axis parallel to the surface of the thin layer; but N_y does *not* vanish, it is in fact equal to 4π. Under the radical, the first factor becomes equal to H and the second to $H + 4\pi M$, which latter is by definition the induction B. We have:

$$\nu = (ge/4\pi mc) \sqrt{HB} \qquad (21)$$

Note here that since B depends upon both H and M, one cannot use the formula unless one knows the value of M, which is the magnetization of the substance at saturation. This usually requires knowledge obtained from other experiments; but we shall meet with a case in which, at least "in principle," the value of B may be found from the resonance-experiment itself.

Equation (21) is the commonest formula for the ferromagnetic resonance, for it fits the "geometry" of the original and of most of the subsequent experiments. Yet there are other formulae corresponding to other geometries, and two of these are particularly important.

It is feasible to orient the layer at right angles to the big applied field. For this case we shall do well to turn the axis of z so that it remains parallel to the big field. Now N_x and N_y vanish and N_z becomes 4π, and the formula is this:

$$\nu = (ge/4\pi mc)(H - 4\pi M) \qquad (22)$$

The quantity $(H - 4\pi M)$ is the internal field H_i, the field strength within the magnetized body. This is the special case in which the right result is obtained by going back to equation (4) and putting for H the actual field strength at the scene of the resonating electrons. In other words, this is the special case in which the naive approach does not lead the student astray.

A more singular special case is that of the sphere. In this case N_x and N_y and N_z are all three of them equal — equal to one another but not to zero. Nevertheless the formula is just our old formula (3), the same as though there were no magnetization at all:

$$\nu = (\mu/I)(H/h) = (ge/4\pi mc)H \tag{23}$$

One wonders how long it would have been before anyone set up equations (18) and derived equation (21), if all experiments had been perfomde with spheres.

In the foregoing pages we have derived the resonance frequency by making certain listed approximations in the basic equations (19). Among these approximations was the neglect of the oscillating field, parallel to the axis of x. We arrive at some interesting results by introducing this field into the equations and giving it an arbitrary frequency, while continuing to make all of the other approximations. It shall be denoted by $H_1 \exp(2\pi i\nu t)$; H_1, it may be recalled, was the symbol used in Part I for the amplitude of this field. In this passage ν shall signify any frequency that the experimenter may choose to apply, while the resonance-frequency heretofore called ν shall change its symbol and become ν_0.

On the right-hand side of the second of the equations (19) will now appear, as the reader can show for himself, $-\gamma MH_1$ instead of zero. The two simultaneous equations now make sense for any value of ν, instead of just the value ν_0. On solving them for M_x^0, one finds:

$$M_x^0/H_1 = \frac{M}{H + (N_x - N_z)M} \, \frac{1}{1 - (\nu/\nu_0)^2} \tag{24}$$

The quantity on the left, and hence also the quantity to which it is equated, is the "susceptibility" of the substance with respect to this oscillating field which, be it remembered, is imposed at right angles to the big applied field.

The quantity on the right has the well-known form of an optical dispersion-curve. Suppose the frequency to be increased from zero. The susceptibility rises from a finite and non-zero value at $\nu = 0$ to positive infinity at the resonance-frequency ν_0; here it jumps suddenly to negative infinity, from which value it rises asymptotically to zero as the frequency is increased toward infinity.

In magnetics there are methods of measuring directly, not the susceptibility χ itself but the sum $(1 + 4\pi\chi)$, which is called the "permeability" and is denoted by μ. It is evident that while the susceptibility is rising, with increase of frequency, from its negative-infinite value at ν_0 to its asymptotic value of zero at infinite frequency, the permeability is rising from negative infinity to an asymptotic value which is equal to +1. Somewhere along this range of frequencies it must pass through zero, at a frequency to be denoted by ν_i. For a stratum parallel to both the big applied field and the oscillating field, it is easily shown that ν_i is equal to $(ge/4\pi mc)B$. This offers a way of determining B and consequently M.

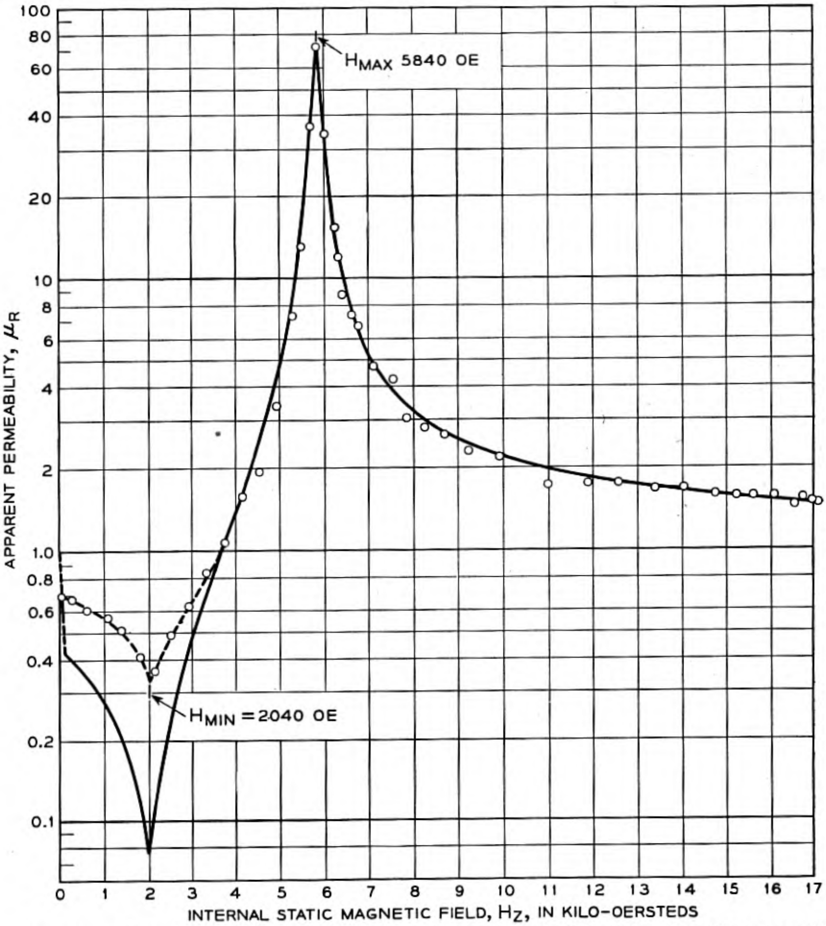


Fig. 4.—Ferromagnetic resonance in Heusler alloy (Cu-Mn-Al). "Apparent" permeability is plotted against H at constant frequency; the resonance-maximum (at the position corresponding to g = 2.02) is vividly shown, as is the minimum mentioned in the text. The solid curve is a theoretical curve based on a specific assumption about damping. (W. A. Yager and F. R. Merritt).

Next suppose that what is plotted against ν is not μ but $|\mu|$, the absolute value of the permeability. The portions of the μ-*vs*-ν curve which were below the horizontal axis now appear inverted and above the horizontal axis. The curve has an upward-pointing peak reaching to infinity at ν_0, and a downward-pointing peak touching the axis with its tip at ν_i.

Such is the general aspect of the curve of Fig. 4, pertaining to a Heusler alloy. There are superficial differences: the curve of Fig. 4 is plotted against H for constant frequency, and the scale along the axis of ordinates is logarithmic. The reader can easily make allowance for these. There is also a fundamental difference: the curve reveals the presence of damping or relaxation, which broadens the peaks and prevents $|\mu|$ from rising to infinity or dropping quite to zero. The continuous curve is derived from a theory which involves a specific assumption about the damping; one sees that it agrees well with the data excepting in a region around the minimum. Curves such as these are likely to be influenced by anisotropy in the ferromagnetic substance, which reversely can be evaluated from the curves.

How about the values of g for ferromagnetic substances? The Heusler alloy to which Fig. 4 pertains has a value of g which, so far as the accuracy of the experiment permits us to judge, may be identical with the ideal value (the most probable value is however 2.01). This is an exception and not the rule. The range of values is rather wide, though apparently not so wide as in the strongly paramagnetic salts. Most of them lie between 2.22 (for cobalt) and 2.01 (for the Heusler alloy aforesaid); but there are instances of values still higher, including one of 3.75 for manganese arsenide. There is also at least one value lower than 2.00; it is presented by gadolinium, a very interesting element. Below its Curie point at 16° absolute, gadolinium shows a resonance-peak of which the breadth interferes with a precise location of its top; the value of g is given as 1.95 to 1.96. Above the Curie point, gadolinium is paramagnetic, but the peak persists and is sharper; the value of g is 1.95 $\pm$ 0.03. I remind the reader that when the experiment is such that formula (21) must be used, a g-value implies an assumption about the value of M the magnetization of the substance at saturation.

I must not close this topic without alluding to something which there is not space to expound. Experiments on the "gyromagnetic effect" — something which has a much longer history than ferromagnetic resonance — lead to values of a quantity which has also been denoted by g. Until a few years ago it was supposed that this quantity must be the same as the g of these pages; but experiment has ruled otherwise, and theory has been successful in at least suggesting a reason. The g of these pages is now called the spectroscopic splitting factor; the other has been set apart as the "gyromagnetic" g, and some people have even taken to writing it as g', which seems rather unfair to the senior g. It seems to be a general rule that when one of the two is greater than 2.00 the other is smaller than 2.00; and in the case of the Heusler alloy, they may well coincide.

ACKNOWLEDGEMENT

Among the many who have helped me with this article I wish to extend especial thanks to Charles Kittel, A. F. Kip, W. D. Knight, M. E. Packard, E. M. Purcell, J. H. Van Vleck and W. A. Yager; and to Messrs. Packard, Purcell and B. Bleaney for prints of the illustrations captioned with their names.

REFERENCES

This makes no pretense of being a bibliography of magnetic resonance: such an enterprise would cover more pages of this Journal than these articles themselves. It is somewhat, but not much, more than a listing of the sources of the data quoted in the articles, with the names which have been omitted in the belief that names tend to slow down exposition. Many relevant papers, with an abundance of footnote references to anterior work, are to be found in what I abbreviate by P.I.C.S.R. ("Proceedings of the International Conference on Spectroscopy at Radiofrequencies, Amsterdam, 1950" separately published and also printed as part of Volume 17 of *Physica*).

The *locus classicus* for the nuclear resonance absorption is the paper of Bloembergen, Purcell and Pound (*Phys. Rev.* **73,** p. 670, 1948); the first publication of this school is by Purcell, Torrey and Pound in *Phys. Rev.* **69,** p. 37, 1946. The *locus classicus* for the precession-theory and the nuclear-induction method is the paper of Bloch (*Phys. Rev.* **70,** p. 460, 1946), followed by the first lengthy description of nuclear-induction measurements by Bloch, Hansen and Packard (*ibid.* p. 464); the first publication of this school is in *Phys. Rev.* **69,** p. 127, 1946. The discovery of nuclear magnetic resonance by the molecular-beam technique was disclosed by Rabi, Zacharias, Millman and Kusch in *Phys. Rev.* **53,** p. 318, 1938, and a more detailed account is given *ibid.* **55,** p. 526, 1939; consult also the paper of Kellogg, Ramsey, Rabi and Zacharias *ibid.* **57,** p. 677 (1940) for the resonances of protons and deuterons in molecular beams of H_2, D_2 and HD.

The reference to the article of G. E. Pake (*Am. Jour. Phys.* **18,** pp. 438–452 and pp. 473–486, 1950) is here repeated to draw attention to this excellent survey of nuclear magnetic resonance and relaxation. Another survey article is that of Rollin, *Reports on Recent Progress in Physics*, 1948–49. The reference for the chemical shift in ethyl alcohol (Fig. 7 of Part I) is Arnold, Dharmatti and Packard, *J. Chem. Phys.* **19,** p. 507, 1951. For the influence of F-centres on nuclear relaxation-time see Hatton and Rollin, *Proc. Roy. Soc.* **199,** p. 231, 1949.

For the final experiments on determination of g for electrons in hydrogen atoms see Koenig, Prodell and Kusch (*Phys. Rev.* **88,** p. 191, 1952); references to earlier

work on other atoms are listed there. For the determination of the proton moment see Gardner, *Phys. Rev.* **83,** p. 996, 1951, and Sommers, Hipple and Thomas, *ibid.* **80,** p. 487, 1950. For the measurement of g with F-centres see Hutchinson and Noble, *ibid.* **87,** p. 1152, 1952, and Tinkham and Kip. *ibid.* **83,** p. 657, 1951, and Schneider and England in P.I.C.S.R.; this last is also the source for the work ion zinc sulfide phosphor. For resonance of conduction-electrons see Griswold, Kip and Kittel, *Phys. Rev.* **88,** p. 951, 1952, and papers yet to be published.

The resonance-peak of Fig. 2 of Part II of this article, for porphyrexide, comes from Holden, Yager and Merritt, *J. Chem. Phys.* **19,** p. 1319, 1951. The literature of resonance in strongly paramagnetic salts is extensive and tough; I refer to the papers in P.I.C.S.R. and the references they give. The basic theory is to be found in the book of J. H. Van Vleck, *Electric and Magnetic Susceptibilities* (Oxford, 1932).

Hyperfine structure of electron resonance was discovered by the late R. P. Penrose (see *Nature,* **163,** pp. 988 and 992, 1949). This field is almost a monopoly of Britain and in particular of Oxford; many of the papers bear the name of Bleaney with or without collaboratores. Fig. 3 of Part II of this article comes from *Proc. Phys. Soc.* **63,** p. 1369, 1950; the statements about vanadium 50 from *ibid.* **65,** p. 952, 1952.

The discoverer of ferromagnetic resonance was J. H. E. Griffiths (see *Nature,* **158,** p. 670, 1946). The precession-theory for ferromagnetic substances is due to Kittel; equation (21) of this article is derived in *Phys. Rev.* **71,** p. 270, 1947; a fuller treatment appears *ibid.* **73,** p. 155, 1948. Survey articles are those of Van Vleck in P.I.C.S.R. and Kittel in *Jour. de Phys.,* **12,** p. 291, 1951. Fig. 4 comes from the paper of Yager and Merritt, in *Phys. Rev.* **73,** p. 318, 1949; a similar curve for supermalloy appears in the preceding paper; references to the g-values of other ferromagnetic substances are given by Yager. The question of "g versus g'" is discussed by Kittel in *Phys. Rev.* **76,** p. 743 (1949).

A Study of Non-Blocking Switching Networks

By CHARLES CLOS

(Manuscript received October 30, 1952)

This paper describes a method of designing arrays of crosspoints for use in telephone switching systems in which it will always be possible to establish a connection from an idle inlet to an idle outlet regardless of the number of calls served by the system.

INTRODUCTION

The impact of recent discoveries and developments in the electronic art is being felt in the telephone switching field. This is evidenced by the fact that many laboratories here and abroad have research and development programs for arriving at economic electronic switching systems. In some of these systems, such as the ECASS System,* the role of the switching crossnet array becomes much more important than in present day commercial telephone systems. In that system the common control equipment is less expensive, whereas the crosspoints which assume some of the control functions are more expensive. The requirements for such a system are that the crosspoints be kept at a minimum and yet be able to permit the establishment of as many simultaneous connections through the system as possible. These are opposing requirements and an economical system must of necessity accept a compromise. In the search for this compromise, a convenient starting point is to study the design of crossnet arrays where it is always possible to establish a connection from an idle inlet to an idle outlet regardless of the amount of traffic on the system. Because a simple square array with N inputs, N outputs and N^2 crosspoints meets this requirement, it can be taken as an upper design limit. Hence, this paper considers non-blocking arrays where less than N^2 crosspoints are required. Specifically, this paper describes for an implicit set of conditions, crossnet arrays of three, five,

* Malthaner, W. A., and H. Earle Vaughan, An Experimental Electronically Controlled Switching System. Bell Sys. Tech. J., 31, pp. 443–468, May, 1952.

etc., switching stages where less than N^2 crosspoints are required. It then deals with conditions for obtaining a minimum number of crosspoints, cases where the N inputs and N outputs can not be uniformly assigned to the switches, switching arrays where the inputs do not equal the outputs, and arrays where some or all of the inputs are also outputs.

SQUARE ARRAY

A simple square array having N inputs and N outputs is shown in Fig. 1. The number of crosspoints equals N^2 and any combination of N or less simultaneous connections can exist without blocking between the inputs and the outputs. The number of switching stages, s, is equal to 1. The number of crosspoints, $C(s)$, is:

$$C(1) = N^2 \tag{1}$$

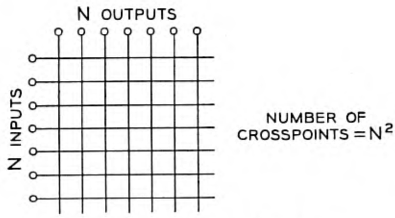


Fig. 1 — Square Array.

THREE-STAGE SWITCHING ARRAY

An array where less than N^2 crosspoints are required is shown in Fig. 2. This array has $N = 36$ inputs and $N = 36$ outputs. There are three switching stages, namely, an input stage (a), an intermediary stage (b), and an output stage (c). In stage (a) there are six 6 x 11 switches; in stage (b) there are eleven 6 x 6 switches; and in stage (c) there are six 6 x 11 switches. In total, there are 1188 crosspoints which are less than the 1296 crosspoints required by equation (1).

Of interest are the derivations of the various quantities and sizes of switches. In stage (a) the number, n, of inputs per switch was assumed to be equal to $N^{1/2}$, thus giving six switches and six inputs per switch. In a similar manner stage (c) was assigned six switches and six outputs per switch. The number of switches required in stage (b) must be sufficient to avoid blocking under the worst set of conditions. The worst case occurs when between a given switch in stage (a) and a given switch in stage (c): (1) five links from the switch in stage (a) to five corresponding switches in stage (b) are busy; (2) five links from the switch in stage (c) are busy to five additional switches in stage (b); and (3) a connection is desired between the given switches. Thus eleven switches are required in stage (b). The remaining requirements, namely, eleven verticals per switch in stages (a) and (c) and six by six switches in stage (b) are then easily derived.

The number of crosspoints required for three stages, where $n = N^{1/2}$, is summarized by the following formula:

$$C(3) = (2N^{1/2} - 1)(3N) \tag{2}$$

$$= 6N^{3/2} - 3N \tag{2a}$$

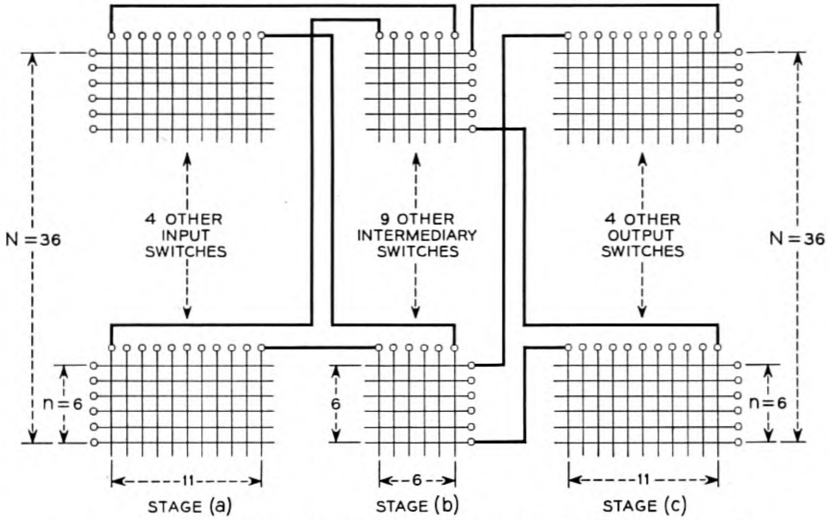
In Table I it may be noted that the number of crosspoints is less than N^2 for all cases of $N \geqq 36$.

PRINCIPLE INVOLVED

The principle involved for determining the number of switches required in the intermediary stage is illustrated in Fig. 3. The figure is for a specific case from which one can generalize for n inputs on a given input switch and m outputs on a given output switch. In the figure it is desired to establish a connection from input B to output H. A sufficient number of intermediary switches are required to permit the $(n - 1)$ inputs other than B on the particular input switch and the $(m - 1)$ outputs other than H on the particular output switch to have connections to separate intermediary switches plus one more switch for the desired connection between B and H. Thus $n + m - 1$ intermediary switches are required.

TABLE I — CROSSPOINTS FOR SEVERAL VALUES OF N

N	Square Array N^2	Three-Stage Array $6N^{3/2} - 3N$
4	16	36
9	81	135
16	256	336
25	625	675
36	1,296	1,188
49	2,401	1,911
64	4,096	2,880
81	6,561	4,131
100	10,000	5,700
...	......	......
1,000	1,000,000	186,737
10,000	100,000,000	5,970,000



NUMBER OF CROSSPOINTS = $6N^{3/2}-3N$ (1188 CROSSPOINTS WHEN N = 36)

Fig. 2 — Three-stage switching array.

FIVE-STAGE SWITCHING ARRAY

A five-stage switching array is illustrated in Fig. 4. The analysis of this array can be made in the following manner. Each input and output switch is assumed to have $n = N^{1/3}$ inputs or outputs, respectively. Connection between a given input switch and a given output switch

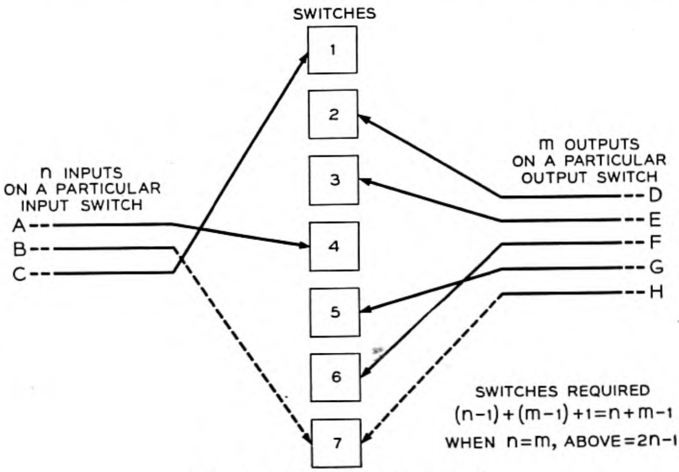


Fig. 3 — Principle involved.

is made via levels, a level consisting of three intermediary switching stages. The number of levels required is $(2N^{1/3} - 1)$. Each level has $N^{2/3}$ inputs and the same number of outputs. The number of crosspoints for a three-stage non-blocking array for $N^{2/3}$ inputs and $N^{2/3}$ outputs can be obtained from equation (2) by substituting $N^{2/3}$ for N in that equation. The total number of crosspoints required for the five-stage array is:

$$C(5) = (2N^{1/3} - 1)^2\, 3N^{2/3} + (2N^{1/3} - 1)\, 2N \tag{3}$$

$$= 16N^{4/3} - 14N + 3N^{2/3} \tag{3a}$$

The number of crosspoints required for several sizes of the five-stage array is given in Table II. The results are compared to the square and three-stage arrays.

SEVEN-STAGE SWITCHING ARRAY

A seven-stage switching array can be analyzed by considering paths requiring five intermediary switching stages as paths via switching aggregates. The number of such aggregates is $(2N^{1/4} - 1)$. Each aggregate has $N^{3/4}$ inputs and a like number of outputs. From equation (3) the crosspoints for each aggregate can be obtained by substituting

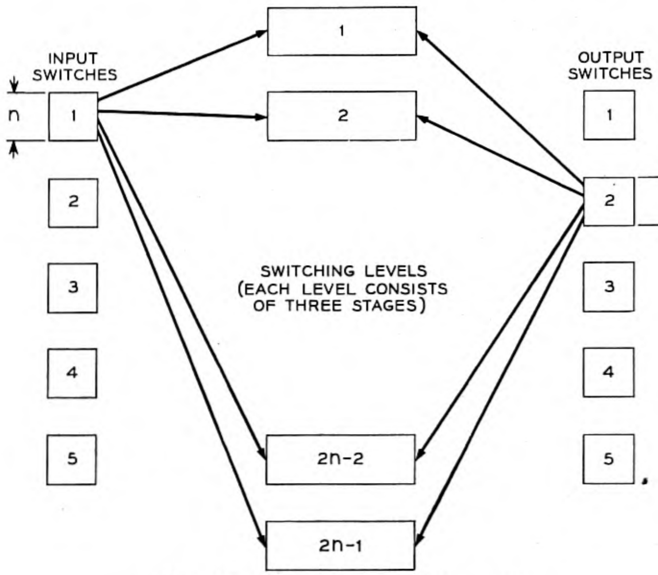


Fig. 4 — Five-stage switching array.

TABLE II — CROSSPOINTS FOR SEVERAL VALUES OF N

N	Square Array	Three-Stage Array	Five-Stage Array
64	4,096	2,880	3,248
729	531,441	115,911	95,013
1,000	1,000,000	186,737	146,300
10,000	100,000,000	5,970,000	3,434,488

$N^{3/4}$ for N in that equation. The total number of crosspoints required for the seven-stage array is:

$$C(7) = (2N^{1/4} - 1)^3\, 3N^{1/2} + (2N^{1/4} - 1)^2\, 2N^{3/4} + (2N^{1/4} - 1)2N \tag{4}$$

$$= 36N^{5/4} - 46N + 20N^{3/4} - 3N^{1/2} \tag{4a}$$

GENERAL MULTI-STAGE SWITCHING ARRAY

Equations (1), (2a), (3a) and (4a) are herewith tabulated as a series of polynomials together with the next polynomial:

$$C(1) = N^2 \tag{1}$$

$$C(3) = 6N^{3/2} - 3N \tag{2a}$$

$$C(5) = 16N^{4/3} - 14N + 3N^{2/3} \tag{3a}$$

$$C(7) = 36N^{5/4} - 46N + 20N^{3/4} - 3N^{1/2} \tag{4a}$$

$$C(9) = 76N^{6/5} - 130N + 86N^{4/5} - 26N^{3/5} + 3N^{2/5} \tag{5}$$

These polynomials can be determined for any number of switching stages from the following formula where s is an odd integer:

$$C(s) = 2\Sigma_{k=2}^{k=\frac{s+1}{2}} N^{\frac{2k}{s+1}} \left(2N^{\frac{2}{s+1}} - 1\right)^{\frac{s+3}{2} - k} + N^{\frac{4}{s+1}} \left(2N^{\frac{2}{s+1}} - 1\right)^{\frac{s-1}{2}} \tag{6}$$

An alternative expression equivalent to equation (6) has been suggested by S. O. Rice and J. Riordan. The recurrence relation used in individually deriving the foregoing polynomials can be used to directly derive the following formula:

$$C(2t + 1) = \frac{n^2(2n - 1)}{n - 1} [(5n - 3)(2n - 1)^{t-1} - 2n^t] \tag{6a}$$

where $s = 2t + 1$

$N = n^{t+1}$

Table III gives comparative numbers of crosspoints for various num-

TABLE III — CROSSPOINTS FOR VARIOUS NUMBERS OF SWITCHING STAGES, s, AND VALUES OF N

N	$s = 1$	$s = 3$	$s = 5$	$s = 7$	$s = 9$
100	10,000	5,700	6,092	7,386	9,121
200	40,000	16,370	16,017	18,898	23,219
500	250,000	65,582	56,685	64,165	78,058
1,000	1,000,000	186,737	146,300	159,904	192,571
2,000	4,000,000	530,656	375,651	395,340	470,292
5,000	25,000,000	2,106,320	1,298,858	1,295,294	1,511,331
10,000	100,000,000	5,970,000	3,308,487	3,159,700	3,625,165

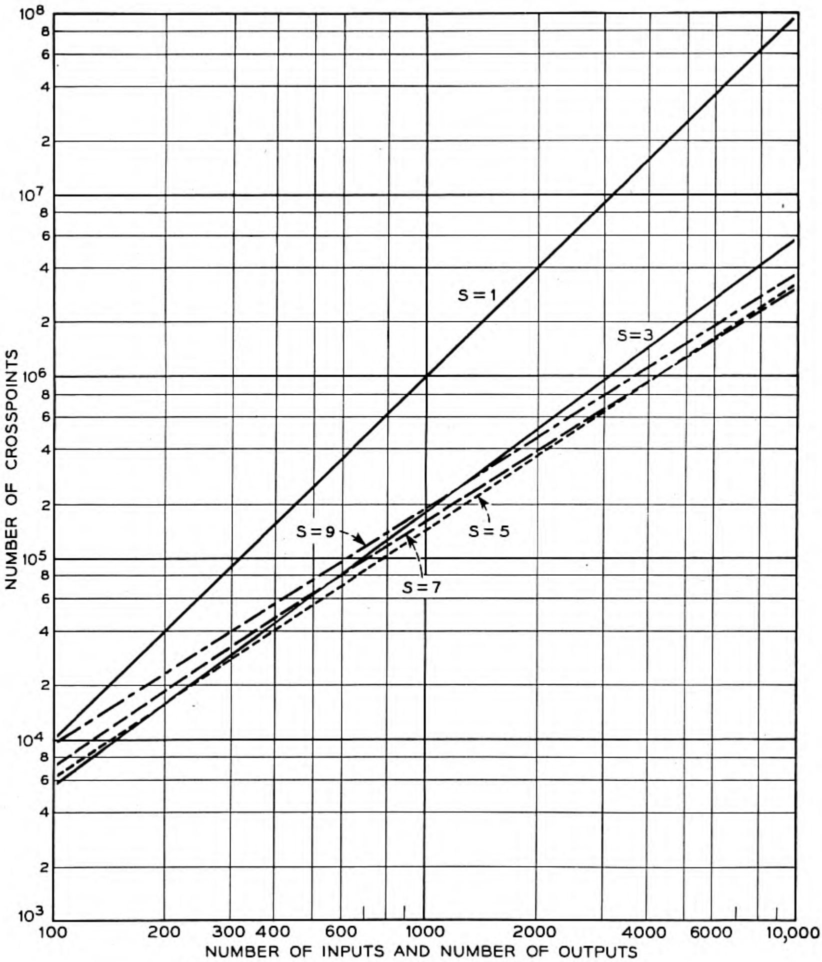


Fig. 5 — Crossoints versus switching stages.

bers of switching stages and sizes of N. The data of Table III are plotted on Figure 5. The series of curves appear to be bounded by an envelope, representing a minumum of crosspoints. The next section dealing with minima indicates that points exist below this envelope.

MOST FAVORABLE SIZE OF INPUT AND OUTPUT SWITCHES IN THE THREE-STAGE ARRAY

The foregoing derivations were for implicit relationships between n and N, namely, n being the $\left(\frac{s+1}{2}\right)th$ root of N. To obtain minimum number of crosspoints a more general relationship is required. For the three stage switching array this is:

$$C(3) = (2n - 1)\left(2N + \frac{N^2}{n^2}\right) \tag{7}$$

When $n = N^{1/2}$ equation (7) reduces to equation (2).

For a given value of N, the minimum number of crosspoints occurs when $dC/dn = 0$ which gives:

$$2n^3 - nN + N = 0 \tag{8}$$

This equation has the following two pairs of integral values:

$$n = 2, \qquad N = 16 \qquad \text{and} \qquad n = 3, \qquad N = 27$$

As N approaches large values equation (8) can be approximated by:

$$N \doteq 2n^2 \tag{9}$$

Graphs of equations (8) and (9) are shown in Fig. 6. In Table IV the numbers of crosspoints are based on the nearest integral values of n for given values of N.

Where comparisons can be made, Table IV indicates fewer crosspoints than does Table I. This fact can be realized in another manner. By eliminating n in equations (7) and (9), the result for large values of N is:

$$C(3) \doteq 4\,(2)^{1/2}N^{3/2} - 4N \tag{10}$$

Equation (10) indicates fewer crosspoints than does equation (2).

MOST FAVORABLE SWITCH SIZES IN THE FIVE-STAGE ARRAY

If n be the number of inputs per input switch and outputs per output switch, and m be the number of inputs per switch in the second stage

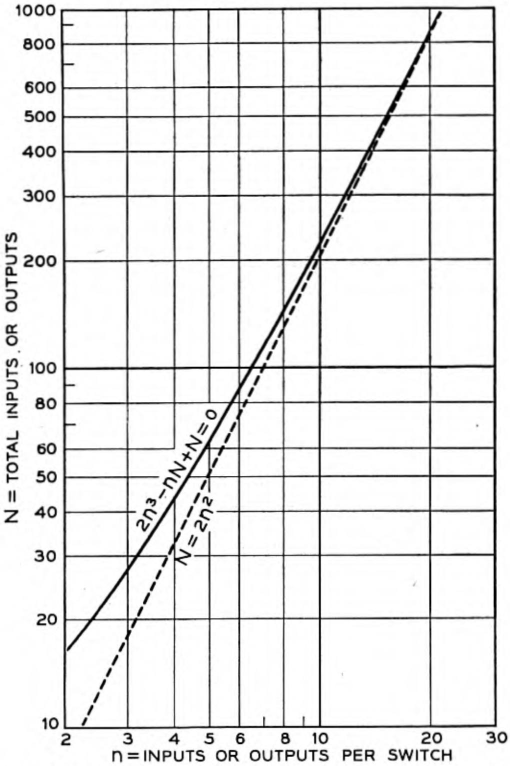


Fig. 6 — Relationship between N and n for minima in crosspoints. Three-stage array.

TABLE IV — CROSSPOINTS FOR SEVERAL VALUES OF N

N	Nearest Integral Value of n	Number of Crosspoints	
		N^2	Equation (7)
16	2	256	288
27	3	729	675
40	4	1,600	1,260
44	4	1,936	1,463
55	5	3,025	2,079
60	5	3,600	2,376
65	5	4,225	2,691
78	6	6,084	3,575
84	6	7,056	4,004
98	7	9,604	5,096
105	7	11,025	5,655

and outputs per switch in the fourth stage, then the following equation gives the total number of crosspoints:

$$C(5) = (2n - 1)\left[2N + (2m - 1)\left(\frac{2N}{n} + \frac{N^2}{n^2m^2}\right)\right] \tag{11}$$

The partial derivative of this equation with respect to m when set equal to zero yields:

$$n = \frac{N(m - 1)}{2m^3} \tag{12}$$

The partial derivative of this equation with respect to n when set equal to zero yields the following equation:

$$N = \frac{nm^2(2n^2 + 2m - 1)}{(2m - 1)(n - 1)} \tag{13}$$

Equations (12) and (13) can be solved for n and m in terms of given values of N. For example for $N = 240$, we obtain $n = 6.81$ and $m = 3.56$.

SEARCH FOR THE SMALLEST N FOR A GIVEN n FOR THE THREE-STAGE ARRAY

For a given value of n, equation (7) furnishes a means for locating that size of three-stage switching array which has N^2 or fewer crosspoints. This can be done by setting equation (7) equal to N^2:

$$N^2 = (2n - 1)\left(2N + \frac{N^2}{n^2}\right) \tag{14}$$

and solving for N in terms of n. The solution is:

$$N \geqq \frac{2n^2(2n - 1)}{(n - 1)^2} \tag{15}$$

Minimum values of N for given values of n are listed in Table V. This table also lists the next highest N exactly divisible by n. From this table it appears that when $N = 24$, we have the smallest switching array for which it may be possible to have less than N^2 crosspoints. However for $N = 25$, as shown in Table I, equation (2) gives more than N^2 crosspoints. The problem is one of finding an array for $N = 25$ with fewer than N^2 crosspoints. For this and all cases beyond, the next section indicates that it is profitable to consider situations where N is not exactly divisible by n.

CASES IN THE THREE-STAGE SWITCHING ARRAY WHERE $N \equiv r(\text{MOD } n)$

Table I indicated that for $N = 25$ and $n = 5$ a total of 675 crosspoints were required. A square array requires only 625. Fig. 7 shows a layout of switches where $N = 25$ and $n = 3$. In this case one input is left over when 25 inputs are divided into threes. The lone input requires three paths to the intermediary switches. This is in accordance with Fig. 3. The lone output also requires three paths to the intermediary switches. Also from Fig. 3, the lone input to the lone output requires only one path. Hence there must be one switch capable of connecting the lone input to the lone output. The number of crosspoints required is 615 which is less than the 625 required by the square array. This scheme can be extended to any case where $N = kn + r$, where the remainder, r, is an integer greater than zero but less than n. The formula for the number of crosspoints where k input and k output switches of size n and one input and output switch of size r are used is:

$$C = 2(2n - 1)(N - r) + 2(n + r - 1)r + (n - r)\left(\frac{N - r}{n}\right)^2 + (n + r - 1)\left(\frac{N - r}{n} + 1\right)^2 - n + r \tag{16}$$

I. G. Wilson has pointed out that for a lone input the crosspoints in the intermediary switches can be used to isolate its possible connections hence no crosspoints are required in the input stage. This likewise applies for a lone output. With this modification the array in Fig. 7 requires six fewer crosspoints. For this case, when $r = 1$, the number of crosspoints is:

$$C = 2(2n - 1)(N - 1) + (n - 1)\left(\frac{N - 1}{n}\right)^2 + n\left(\frac{N - 1}{n} + 1\right)^2 - n + 1 \tag{16a}$$

TABLE V — MINIMUM VALUES OF N FOR GIVEN VALUES OF n

n	N per Equation 15	$N \equiv 0 \pmod{n}$
2	24	24
3	22.5	24
4	24.9	28
5	28.1	30
6	31.7	36

J. Riordan has found a more efficient arrangement for cases where $N = kn + r$. In place of using k switches of size n and one switch of size r, he proposes that $(k + 1 - n + r)$ switches of size n and $(n - r)$ switches of size $n - 1$ be used. For this case the number of crosspoints is:

$$C = 2(2n - 1)(k + 1 - n + r)n + 2(2n - 2)(n - r)(n - 1) + (2n - 3)(k + 1)^2 + 2(k + 1)(k + 1 - n + r) \tag{17}$$

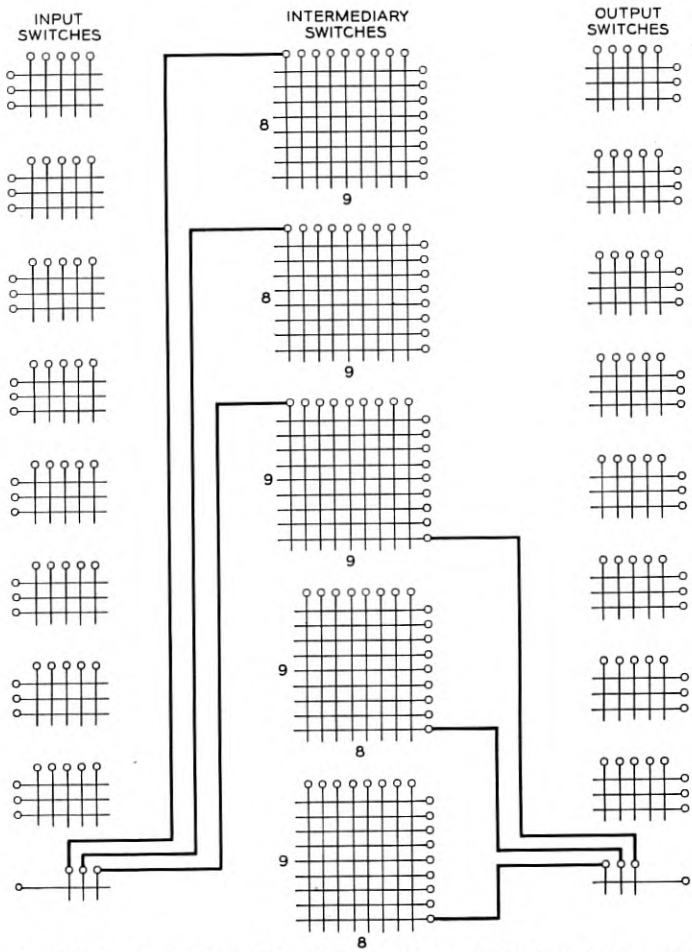


Fig. 7 — Three-stage array. $25 \equiv 1 \pmod 3$. An equivalent arrangement is to provide two 8 x 8 and three 9 x 9 intermediary switches. Two of the 9 x 9 switches need only 80 crosspoints.

TABLE VI — CROSSPOINTS FOR VARIOUS VALUES OF N AND n

N	Square Array	Three-Stage Array			
		$n = 2$	$n = 3$	$n = 4$	$n = 5$
23	529*	540	530	556	—
24	576	576	560*	588	—
25	625	625	609*	633	—
26	—	663	643*	667	—
27	—	716	675*	701	—
28	—	756	730*	735	—
29	—	813	766*	788	—
30	—	855	800*	824	864
31	—	—	861	860*	911
32	—	—	899	896*	951
33	—	—	935*	957	991
34	—	—	1002	995*	1031
35	—	—	1042	1033*	1071
36	—	—	1080	1071*	1128
37	—	—	1153	1140*	1170
38	—	—	1195	1180*	1212
39	—	—	1235	1220*	1254
40	—	—	1314	1260*	1296
	$n = 4$	$n = 5$	$n = 6$	$n = 7$	$n = 8$
50	1819	1800*	1879	—	—
60	2415	2376*	2420	—	—
70	3164	3024*	3056	—	—
80	3920	3744*	3764	—	—
90	—	4536	4455*	4499	—
100	—	5400	5291*	5315	—
110	—	—	6199	6100	6156
120	—	—	7040	7044	6975*
130	—	—	8076	7923*	7947
140	—	—	—	8840*	8860
150	—	—	—	9968	9811*
160	—	—	—	10979	10800*

* Minimum values.

Equation (17) is identical to equation (16) when $r = n - 1$. There are two cases, namely, when $n = 2$ and $n = 3$ where equation (16a) gives fewer crosspoints than does equation (17).

SEARCH FOR THE MINIMUM NUMBER OF CROSSPOINTS BETWEEN $N = 23$ AND $N = 160$

The equations of the preceding sections furnish a means for searching for minimum crossnet arrays. Table VI shows the results of such a search up to $N = 160$. Results are indicated in unit steps from $N = 23$ to $N = 40$ and for every tenth interval thereafter. At $N = 161$, a five-stage array requires the fewest crosspoints.

Table VI was computed by the use of finite differences. The equations were:

$$C[(k+1)n] - C(kn) = (2n-1)(2n+2k+1) \tag{18}$$

$$C(kn+r+1) - C(kn+r) = 2(k+3n-1) \tag{19}$$

$$C(kn+1) - C(kn) = 2kn+1 \tag{19a}$$

Equation (18) was derived from equation (7) with N being replaced by $(k+1)n$ and by kn as required. Equation (19) was derived from equation (17) with r being replaced by $r+1$ as required. This equation applies for all values of n greater than 3 and for the particular case of $n = 3$ and $r = 2$. Equation (19a) was derived from Equations (16a) and (7) and is for the particular case of $r = 1$, when $n = 2$ and $n = 3$.

SEARCH FOR THE MINIMUM NUMBER OF CROSSPOINTS FOR $N = 240$

For a case where N is large enough to require five-switching stages, the search for the minimum number of crosspoints should be based on equations (12) and (13) and on the use of Table VI. The method is suggested by means of Table VII. The data in a previous section indicate

TABLE VII — CROSSPOINTS FOR $N = 240$ AND VARIOUS VALUES OF n

Input and Output Stages				Intermediary Stages				Total Crosspoints
n	No. of Switches	Size of Switches	Cross-points	No. of Levels	Inputs and Outputs	m	Cross-points	
2	120	2 x 3	1,440	3	120 x 120*	8	20,925	22,365
3	80	3 x 5	2,400	5	80 x 80*	5	18,720	21,120
4	60	4 x 7	3,360	7	60 x 60*	5	16,632	19,992
5	48	5 x 9	4,320	9	48 x 48	4	15,120	19,440
6	40	6 x 11	5,280	11	40 x 40*	4	13,860	19,140
7	30	7 x 13	5,460	2	30 x 35	3	1,826	19,369
	5	6 x 12	720	11	35 x 35*	4	11,363	
8	30	8 x 15	7,200	15	30 x 30*	3	12,000	19,200
9	24	9 x 17	7,344	2	24 x 27	3	1,230	19,467
	3	8 x 16	768	15	27 x 27*	3	10,125	
10	24	10 x 19	9,120	19	24 x 24*	3	10,640	19,760
11	20	11 x 21	9,240	2	20 x 22	—	880	20,116
	2	10 x 20	800	19	22 x 22	—	9,196	
12	20	12 x 23	11,040	23	20 x 20	—	9,200	20,240

Crosspoints per equation (3) five-stage array	20,596
Crosspoints per equation (2) three-stage array	21,624
Crosspoints per equation (1) square array	57,600

* See Table VI for minimum number of crosspoints.

that a minimum should occur for $N = 240$, when $n = 6.81$ and $m = 3.56$. In Table VII the minimum occurs when $n = 6$ and $m = 4$. It fails to occur at $n = 7$ because 240 is not exactly divisible by 7. Except for this situation, the minimum would have occurred as predicted.

RECTANGULAR ARRAY

Referring to Fig. 1, if there were N inputs and M outputs, a simple rectangular array would result which would be capable of sustaining up to N or M, whichever is the lesser, simultaneous connections without blocking. The number of crosspoints is:

$$C(1) = NM \tag{20}$$

N INPUTS AND M OUTPUTS IN A THREE-STAGE ARRAY

For the case of a three-stage switching array with N inputs and M outputs, let there be n inputs per input switch and m outputs per output switch. A particular input to be able to connect without blocking under the worst set of conditions to a particular output will require $(n - 1) + (m - 1) + 1$ available paths. Thus by providing for that many intermediary switches, a non-blocking switching array is obtained. The number of crosspoints is:

$$C(3) = (n + m - 1)\left[N + M + \frac{NM}{nm}\right] \tag{21}$$

Differentiating this equation first with respect to n and then to m yields two partial differential equations whose solution indicates that a minimum is reached when $n = m$. Replacing m by n in equation (21), the equation for the number of crosspoints becomes:

$$C(3) = (2n - 1)\left[N + M + \frac{NM}{n^2}\right] \tag{22}$$

Solving for the minimum number of crosspoints gives the following expression:

$$n^3 - \frac{NM}{N + M}n + \frac{NM}{N + M} = 0 \tag{23}$$

When $N = M$ this equation reduces to equation (8).

The three-way relationships of n, N and M are shown in Fig. 8.

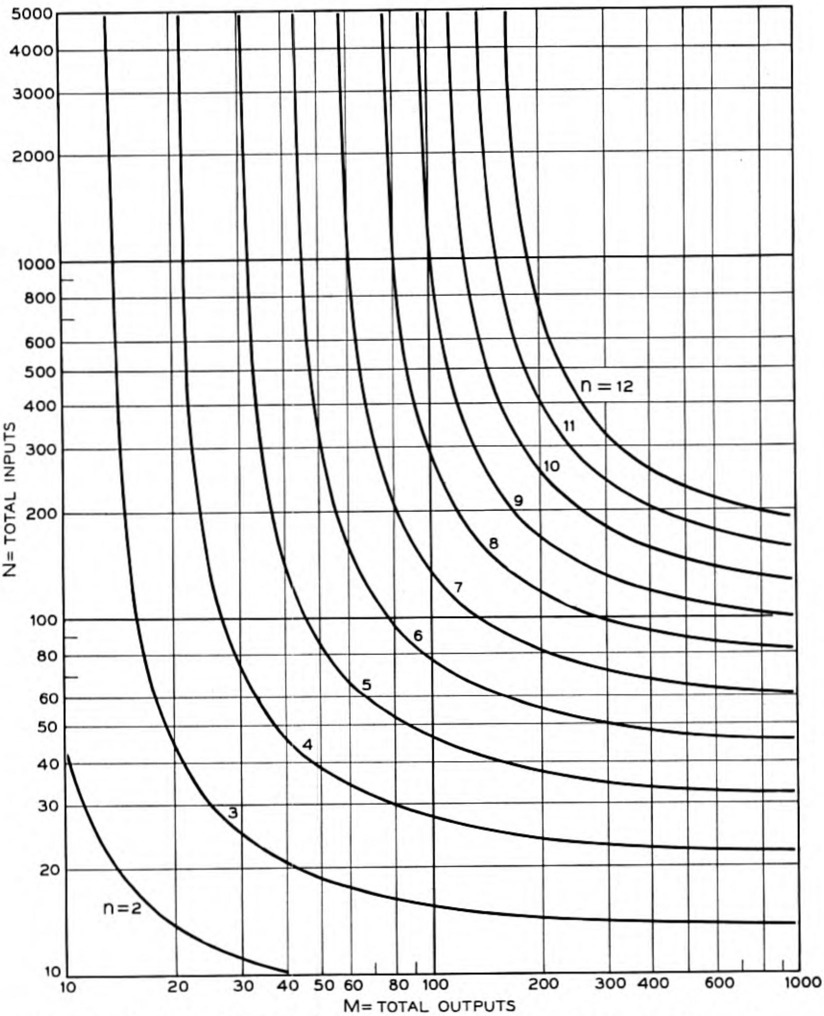


Fig. 8 — Relationship of n to N inputs and M outputs for a minimum in crosspoints in a three stage array.

TRIANGULAR ARRAY

If a case exists where all inputs are also the outputs, then an arrangement such as is shown in Fig. 9 can be used. The crosspoints in the intermediary switches permit connections between all switches on the left hand side. For connections between two trunks on the same switch it is assumed that one of the links to an intermediary switch can be used to establish the connection but without affecting any of the crosspoints on the intermediary switch. The number of crosspoints for this case is:

$$C = (2n - 1)\left(T + \frac{T^2}{2n^2} - \frac{T}{2n}\right) \tag{24}$$

where T = number of two-way trunks.

By differentiation, conditions for obtaining minimum numbers of crosspoints can be determined. The arrangement can also be extended to cases where extra switching stages are required.

ONE-WAY INCOMING, ONE-WAY OUTGOING AND TWO-WAY TRUNKS

A combination of the triangular array of Fig. 9 and of unequal inputs and outputs is shown in Fig. 10. In this figure, one-way incoming,

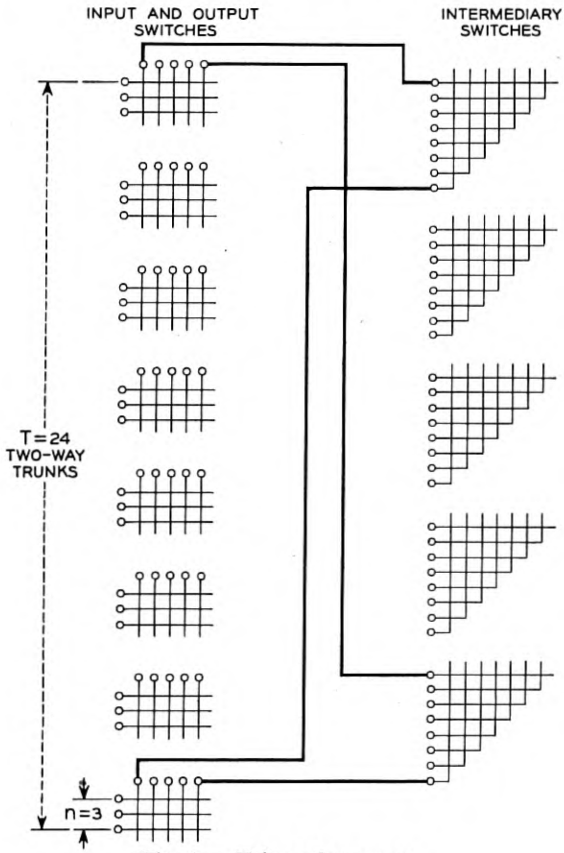


Fig. 9 — Triangular array.

one-way outgoing and two-way trunks can be freely interconnected without blocking. The number of crosspoints for this case is:

$$C = (2n - 1)\left[N + T + M + \frac{NT}{n^2} + \frac{MT}{n^2} + \frac{T^2}{2n^2} - \frac{T}{2n}\right] \quad (25)$$

The comments concerning the triangular array also apply for this case.

COMPARISON WITH EXISTING NETWORKS

Few existing crossnet arrays are non-blocking. An example is the four-wire intertoll trunk concentrating system. In one of its standard sizes 4,000 crosspoints are required for 100 incoming trunks and 40 outgoing intertoll trunks. From Fig. 8, for $N = 100$ and $M = 40$ it may be noted that the nearest integral value for n is 5. By substituting this value in equation (22), a non-blocking three-stage switching array of 2,700 cross-

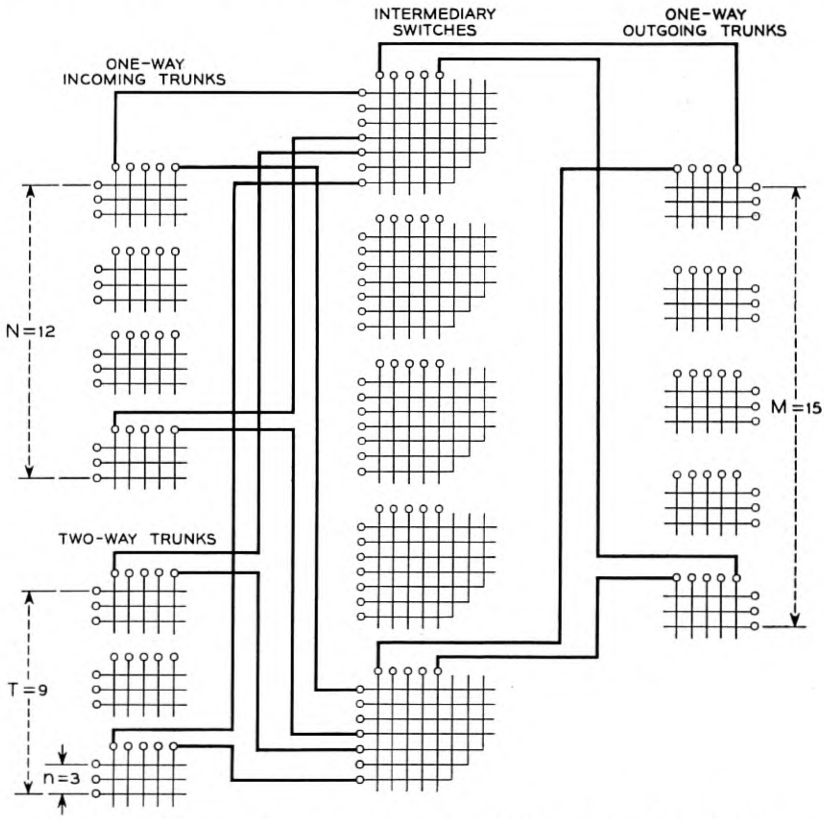


Fig. 10 — One-way incoming, one-way outgoing and two-way trunks.

points is found which could be used for the concentrating switch. In this case the new approach to the switching network problem may prove to be of value.

Comparisons with existing arrays having blocking are likely to be unfavorable because the grades of service are not the same. For instance, a No. 1 crossbar district-to-office layout of 1,000 district junctors and 1,000 trunks requires 80,000 crosspoints. This layout can handle 708 erlangs with a blocking loss of 0.0030. The minimum number of crosspoints with a non-blocking array is slightly less than 138,000. This, however, can handle 1,000 erlangs without blocking. By introducing blocking into the design methods described in this paper, a more favorable comparison with existing arrays having blocking can be made. This can be done by omitting certain of the paths. If done to an array requiring 1,000 inputs and 1,000 outputs a layout can be obtained requiring 79,900 crosspoints with a blocking loss of 0.0022 for a load of 708 erlangs. For this example, at least, it appears that the new design methods may prove to be valuable especially for use in the development of electronic switching systems where the control mechanism may not be dependent upon the particular switching array used.

CONCLUSION

In present day commercial telephone systems the use of non-blocking switching networks is rare. This may be due to the large number of crosspoints required. With the design methods described herein, a wider use of non-blocking networks may occur in future developments. For the usual case of networks with blocking, new systems have generally been designed by an indirect process. Several types and sizes of switching arrays are studied until the most economical one for a given level of blocking is found. With the new design methods, a straightforward approach is possible. Fig. 5 indicates that a region of minimum values exists. By first designing a non-blocking system with a reasonable number of switching stages and then omitting certain of the paths, the designer can arrive at a network with a given level of blocking and be very close to a minimum in crosspoints. The possibility of the adoption of this direct design method is important.

ACKNOWLEDGEMENT

In addition to those specifically mentioned in this paper, the author is also indebted to E. B. Ferrell, B. D. Holbrook, C. A. Lovell and E. F. Moore for suggestions and encouragement in the preparation of this paper.

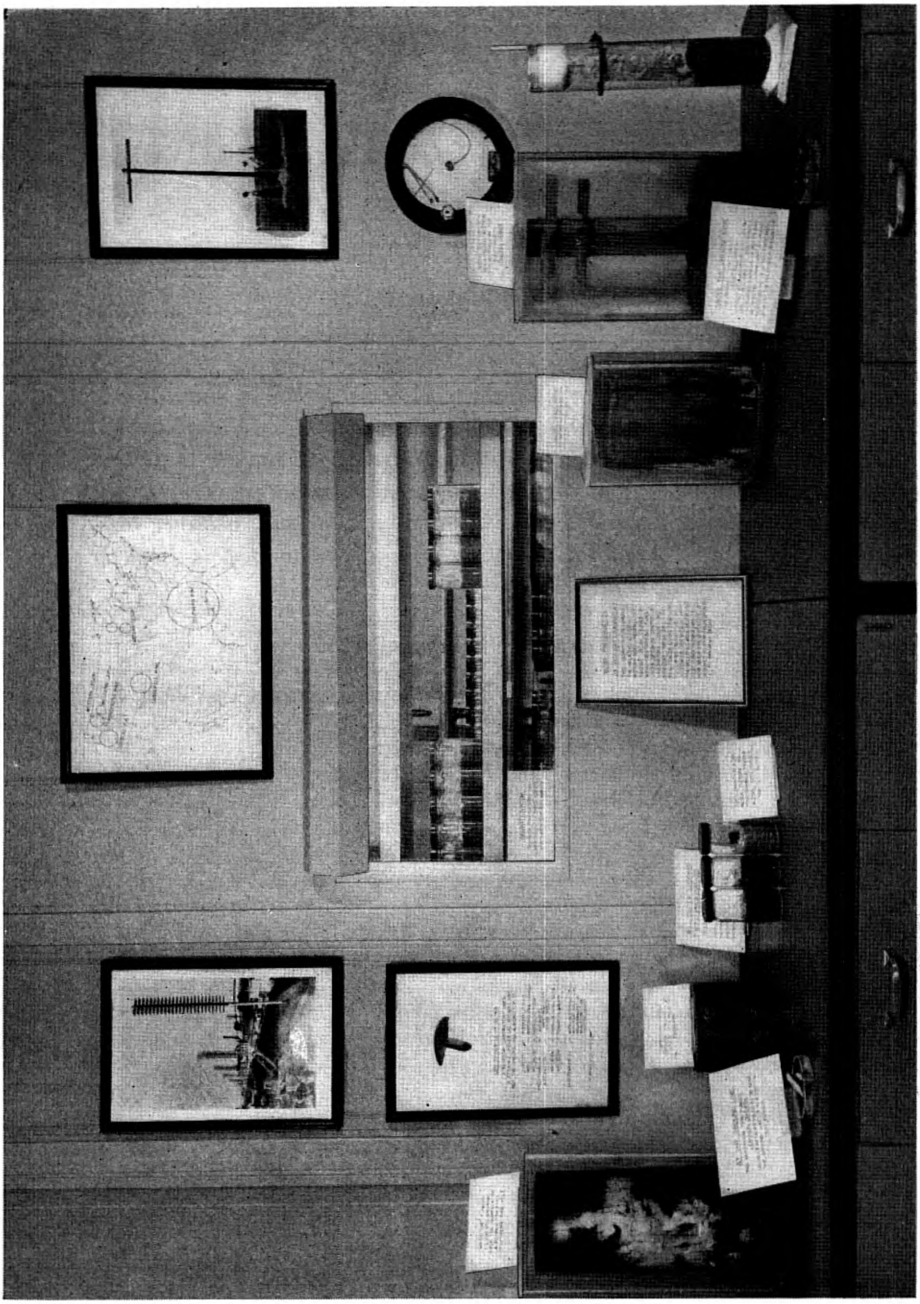
The Evaluation of Wood Preservatives Part II

By REGINALD H. COLLEY

(Manuscript received September 22, 1952)

This paper offers a review and interpretation of laboratory and field experiments aimed at determining the necessary protective threshold quantities of wood preservatives. It details the procedure followed in the soil-block tests at Bell Telephone Laboratories, Incorporated. Discussion of specific criticisms of the techniques involved and replies to these criticisms are included. The paper also presents for the first time a correlation of the results obtained from soil-block culture tests, outdoor exposure tests on stakes and on pole-diameter posts as well as pole line experience.

TABLE OF CONTENTS — Part II



The "Bioassay Corner" at Bell Telephone Laboratories, Murray Hill. Centered in the

EVALUATION BY TREATED ¾-INCH SOUTHERN PINE SAPWOOD STAKES IN TEST PLOTS

Rating the Condition of the Stakes

One of the general and unavoidable difficulties in experiments involving exposure of small specimens in test plots is arriving at a measure of the inspector's judgment of the condition of the individual specimens at each inspection period. Bell Laboratories' investigators have used a system of numbers beginning with 10 as the highest, and running down in single steps to 0, to define the various gradations of destruction shown in the specimens as they pass from perfectly sound to the state of "failed" units. This system has been considered by some as slightly cumbersome; but it is a truly effective method of depreciation rating in a continuous series of inspections. In such a series any minor errors of judgment in one season can be corrected in the next. Slow depreciation can be recorded in the upper ratings until progressive destruction becomes clearly evident. Most observers of test plot experiments on treated wood specimens use a series of five numbers for five condition categories, about as follows:

10.0 Sound—no decay.
7.5 Surface soft—suspicious of decay.
5.0 Slight—positive decay.
2.5 Severe—deep decay.
0.0 Failed—almost complete loss of strength.

Some, like Rennerfelt[88] for example, use this system upside down, with 0 for no decay, and 10 for failure. The writer has proposed[85] the use of a new 5-number depreciation system, with the same definitions, based on the logarithms of the above 5 figures, and rounded off to 10, 9, 7, 4 and 0 respectively. This simplifies the Bell Telephone Laboratories' 11 division, 10–0, system while retaining the advantage of slow depreciation at first; and at the same time it avoids the sudden, and in the writer's opinion, unjustified drop from 10 to 7.5 in the arithmetic series for suspicious-of-decay specimens. In the following presentation and interpretation of the behavior of some of the ¾-inch stakes in the Gulfport test plot the recorded per cent condition of the stakes at any one inspection period has been translated into terms of the proposed 5-number log base system.

The stakes were carefully sawed and planed units, ¾-inch square in cross-section.[12, 69] Before treatment the stakes were selected so that they would represent the normal distribution of density in the material available. Table IX shows the analyses of the four different creosotes used to treat the stakes in respective 4- and 8-pound groups. Both empty-cell and full-cell treatments were employed. The full-cell treatments were made with toluene as a diluent in order to provide more uniform and lower controlled retentions in the treated specimens. The empty cell specimens were sorted after treatment to retain the middle group of retentions, with a view to eliminating as far as practicable some of the factors in the empty-cell treatment variation. All of the stakes were treated between March 11 and March 26, 1941, and they were all placed in the plot in the approximate period from April 8 to April 22, 1941. The distribution of retentions in the 8-pound stakes set out in the Gulfport plot are shown in Table X, along with data on average retention, standard deviation, and coefficient of variability.

Table IX — Analyses, Water-Free Basis, of Four Creosotes Used in Treating ¾-inch Southern Pine Sapwood Stakes

1941 series; Gulfport test plot

Creosote BTL No.	5283	5286B	5286A	5285A
Specific gravity, 38/15.5°C	1.055	1.053	1.068	1.111
Distillation, per cent, cumulative				
To 210°C	2.4	4.6	5.1	0.5
210–235	13.2	22.4	20.3	4.2
235–270	38.1	49.3	41.1	18.8
270–300	51.1	60.5	50.0	28.7
300–315	58.4	65.6	53.5	33.1
315–355	81.1	80.7	67.3	53.2
Residue	18.8	18.6	32.5	46.7
Total	99.9	99.3	99.8	99.9
Sulph. res., gm/100 ml	3.4	1.6	0.7	0.7
Tar acids, gm/100 ml	7.8	5.7	4.0	4.0

Table X — Distribution of Retentions,* lb/cu ft at Treatment, of Four Creosotes, 8 lb Empty-Cell (EC) and Full-Cell (FC) Groups

¾-inch southern pine sapwood stakes; 1941 series; Gulfport test plot.

Creosote, BTL No.	5283		5286B		5286A		5285A	
lb/cu ft	EC	FC	EC	FC	EC	FC	EC	FC
6.4	—	—	2	—	2	—	2	—
6.5	—	—	—	—	—	2	—	—
6.6	—	—	4	—	6	2	2	—
6.7	—	—	—	—	—	2	—	—
6.8	—	—	—	2	—	—	4	—
6.9	—	—	—	—	—	2	—	—
7.0	4	2	2	—	—	—	2	—
7.1	—	1	—	—	—	2	—	2
7.2	6	3	8	—	4	4	—	2
7.3	—	2	—	2	—	2	—	6
7.4	6	4	4	—	2	4	6	2
7.5	—	—	—	4	—	2	—	2
7.6	4	2	4	—	8	6	2	3
7.7	—	2	—	2	—	—	—	3
7.8	—	4	2	4	4	—	4	4
7.9	—	6	—	6	—	—	—	7
8.0	2	3	2	—	4	—	6	2
8.1	—	1	—	4	—	—	—	2
8.2	—	2	—	4	—	2	—	1
8.3	2	2	4	2	—	2	—	5
8.4	—	—	—	2	—	—	—	2
8.5	4	—	—	—	4	4	6	—
8.6	—	1	—	—	—	—	—	1
8.7	—	—	—	2	4	2	—	—
8.8	—	1	2	—	—	—	—	—
8.9	2	4	—	2	2	2	2	6
9.1	2	—	—	2	—	—	2	—
9.3	—	—	4	2	—	—	2	—
9.4	—	1	—	—	—	—	—	—
9.5	8	—	—	—	—	—	—	—
9.6	—	1	—	—	—	—	—	—
9.7	—	1	2	—	2	—	—	—
9.8	—	1	—	—	—	—	—	—
12.5	—	—	2	—	—	—	—	—
n								
	40	44	42	40	42	40	40	50
Average Retention lb/cu ft								
	8.12	8.02	7.97	8.07	7.76	7.59	7.79	7.91
Standard Deviation								
	0.913	0.723	1.288	0.514	0.839	0.694	0.758	0.523
Coefficient of Variation								
	11.24	9.01	16.16	6.37	10.81	9.14	9.73	6.61

* All retentions were calculated from weights before and after treatment. The full cell (FC) treatments were made with toluene-creosote solutions.

TABLE XI — INSPECTION RATINGS

¾-inch southern pine sapwood stakes in test 6 and 7 years; 1941 series; Gulfport test plot.

Creosote BTL No.	Treatment	n	Average retention at treatment lb/cu ft	Test period years	Number and per cent of specimens rated					Average per cent condition
					10	9	7	4	0	
5283	EC	40	8.12	6	29 72.5	8 20.0	1 2.5	— —	2 5.0	92.2
				7	17 42.5	17 42.5	2 5.0	1 2.5	3 7.5	85.6
	FC	44	8.02	6	16 36.4	22 50.0	1 2.3	3 6.8	2 4.6	85.7
				7	2 4.6	19 43.2	14 31.8	3 6.8	6 13.6	68.4
5286B	EC	42	7.97	6	25 59.5	10 23.8	4 9.5	1 2.4	2 4.8	88.6
				7	9 21.4	14 33.3	10 23.8	4 9.5	5 11.9	72.0
	FC	40	8.07	6	10 25.0	27 67.5	1 2.5	1 2.5	1 2.5	88.5
				7	1 2.5	19 47.5	18 45.0	— —	2 5.0	76.8
5286A	EC	42	7.76	6	25 59.5	15 35.7	1 2.4	1 2.4	— —	96.5
				7	8 19.0	32 76.2	— —	— —	2 4.8	87.6
	FC	40	7.59	6	10 25.0	27 67.5	1 2.5	1 2.5	1 2.5	89.0
				7	8 20.0	20 50.0	8 20.0	1 2.5	3 7.5	80.0
5285A	EC	40	7.79	6	36 90.0	4 10.0	— —	— —	— —	99.0
				7	29 72.5	9 22.5	— —	2 5.0	— —	95.0
5285A	FC	50	7.91	6	35 70.0	12 24.0	— —	2 4.0	1 2.0	93.2
				7	17 34.0	23 46.0	4 8.0	2 4.0	4 8.0	82.6

Depreciation Curves for ¾ Inch Stakes

Under the conditions at Gulfport the depreciation curves for the creosotes — particularly the low residue oils — show an increased downward pitch at the 6th to 7th year of exposure. The relative proportion of the stakes rated respectively at 10, 9, 7, 4 and 0 at the 6- and 7-year inspections, are shown by number and per cent in Table XI. The change within the one year interval is particularly striking in the 10 and 9 columns. The rating and the distribution of retentions of creosote 5283, empty-cell treatment, at 6 and 7 years, respectively, are shown in Figs.

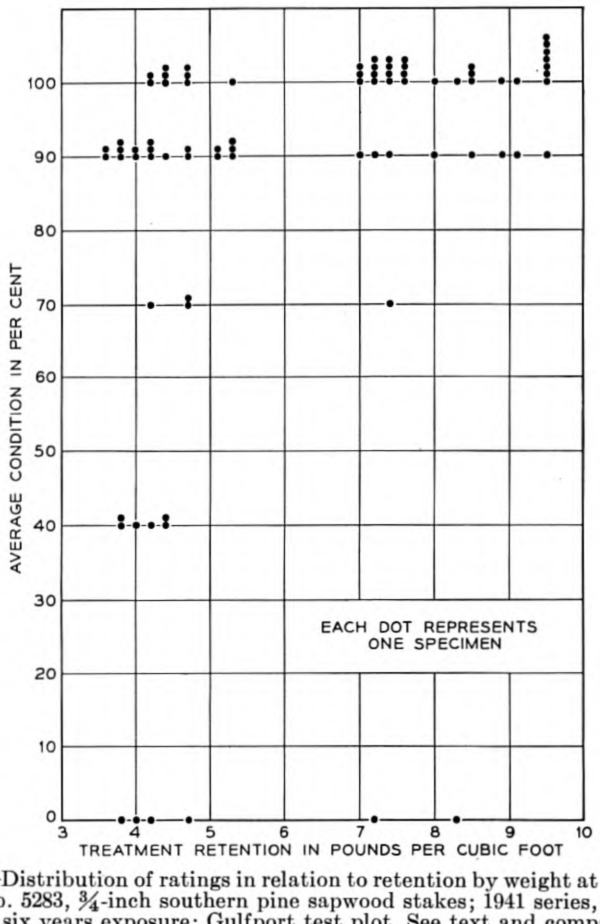


Fig. 15—Distribution of ratings in relation to retention by weight at treatment; creosote No. 5283, ¾-inch southern pine sapwood stakes; 1941 series, empty-cell treatment; six years exposure; Gulfport test plot. See text and companion Figs. 16, 17 and 18.

15 and 16; and similar data for the full cell toluene dilute treatments are shown in Figs. 17 and 18, respectively. In spite of the attempted care in selection of the specimens and in treatment of the empty cell 8-pound group, the lot is heavily weighted (see Table X) toward the high retention end, e.g., by the 12 stakes in the 8.9, 9.1 and 9.5 groups. The presence of these relatively heavily treated stakes may explain in part the position of the depreciation curves for the 8-pound treatments shown in Fig. 19. On the other hand the empty cell 8-pound (nominal) stakes treated with

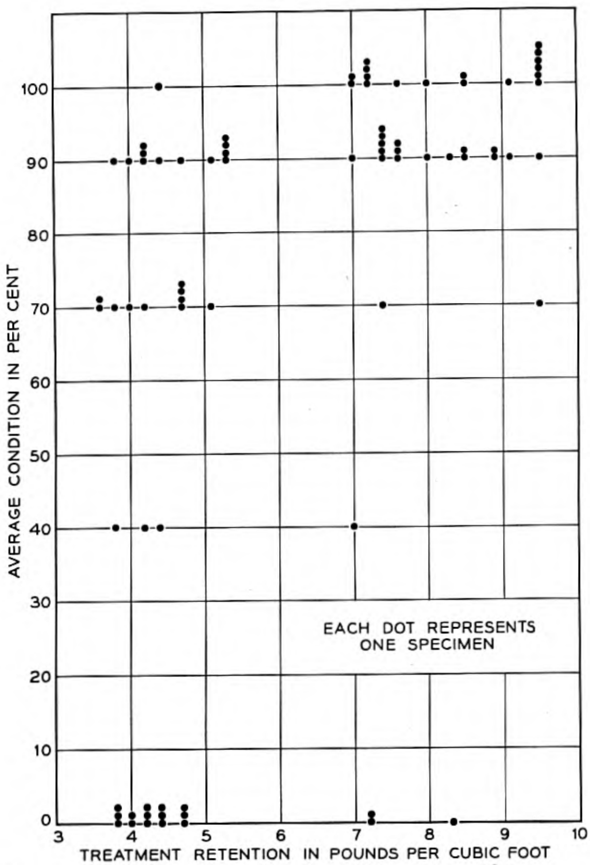


Fig. 16—Distribution of ratings in relation to retention by weight at treatment; creosote No. 5283, empty-cell treatment, seven years exposure.

creosote 5286B at 9.3, 9.7 and 12.5 pounds (see Table X) apparently have not operated to increase the average life of the group treated with this oil as much as appears to be the case in the group treated with creosote 5283. The difference in behavior at the 6–7 year interval of the stakes that were treated by a full cell process with treating solutions made by dissolving the creosote in toluene is even more marked than it is in the empty cell groups.

The average per cent condition of the stakes over the 9-year test period up to 1950 is shown in Table XII. Data are included on the number of stakes in each lot, and on the average treatment retention in

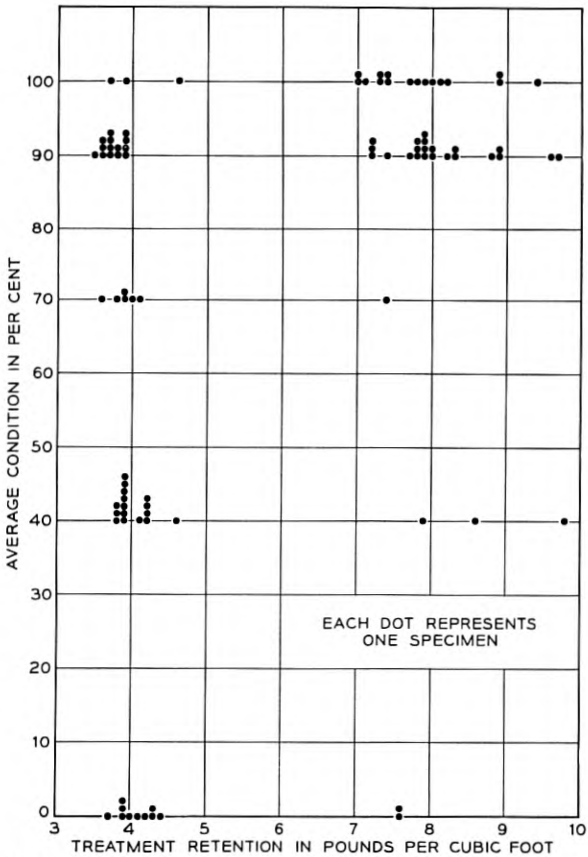


Fig. 17—Distribution of ratings in relation to retention by weight at treatment; creosote No. 5283, full-cell treatment (toluene dilution), six years exposure.

pounds per cubic foot for the respective groups. Two groups of stakes treated with greensalt K[74, 79] at 0.57 and 1.17 pounds per cubic foot, respectively, are included in the table for comparison.

Depreciation curves for the 4- and 8-pound groups for the four creosotes are shown in Figs. 19, 20, 21 and 22. A depreciation curve for greensalt K specimens treated with 1.17 pound per cubic foot is included in Fig. 19.

Estimating Threshold Retentions and Average Life

In presenting the following discussion of a theoretical approach to the estimation of threshold retentions and average life no attempt has been

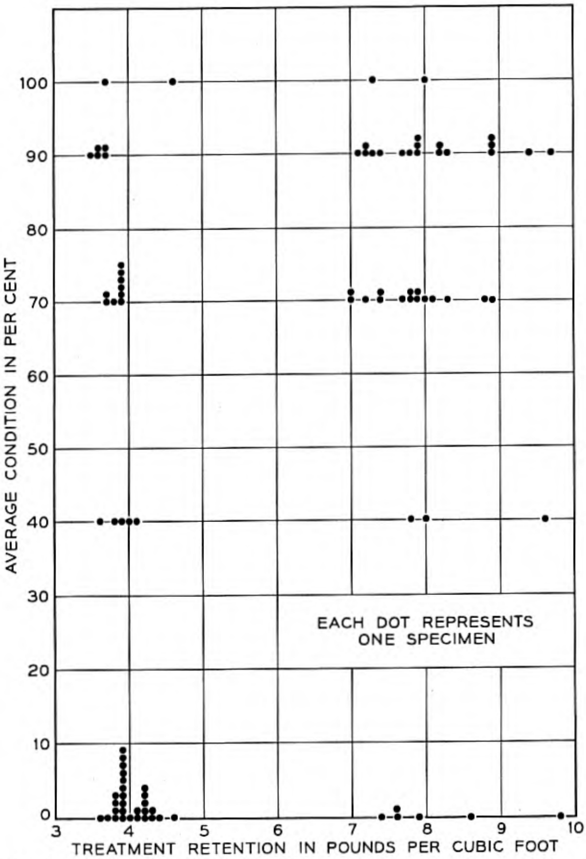


Fig. 18—Distribution of ratings in relation to retention by weight at treatment; creosote No. 5283, full-cell treatment (toluene dilution), seven years exposure.

made to separate the possible effects of the attack by different fungi or combinations of such fungi. The basic data are the figures reported by the inspectors of the stakes; and data on the actual organisms involved are very difficult — if not impossible — to obtain at the time of inspection. For the present purpose then, any differences in rate of decay by different organisms or in different parts of the test plot are all blanketed under the per cent condition averages.

TABLE XII — AVERAGE PER CENT CONDITION OF ¾-INCH SOUTHERN PINE SAPWOOD STAKES TREATED WITH FOUR CREOSOTES, AND WITH GREENSALT K

9 years in test; 1951 series; Gulfport test plot.

Creosote BTL No.	Treatment	n	Average retention at treatment lb/cu ft	Years in test					
				1	3	6	7	8	9
				Average per cent condition					
5283	EC	40	4.38	100	97	74	50	32	13
	EC	40	8.12	100	99	92	86	74	55
	FC	49	3.93	99	86	54	30	16	8
	FC	44	8.02	99	96	86	68	54	31
5286B	EC	42	4.23	100	93	67	40	21	12
	EC	42	7.97	100	100	89	72	54	40
	FC	40	4.03	100	86	56	29	10	1
	FC	40	8.07	100	97	89	77	47	32
5286A	EC	42	3.93	100	96	76	57	42	28
	EC	42	7.76	100	100	97	88	82	72
	FC	40	4.08	100	93	84	64	44	16
	FC	40	7.59	100	96	89	80	72	53
5285A	EC	40	4.15	100	99	91	75	62	53
	EC	40	7.79	100	100	99	95	92	90
	FC	51	3.92	100	94	78	61	46	24
	FC	50	7.91	100	97	93	83	76	62
Greensalt K									
	FC	99	0.57*	100	98	84	80	72	67
	FC	100	1.17	100	99	96	88	86	79

* lb/cu ft of dry salt.

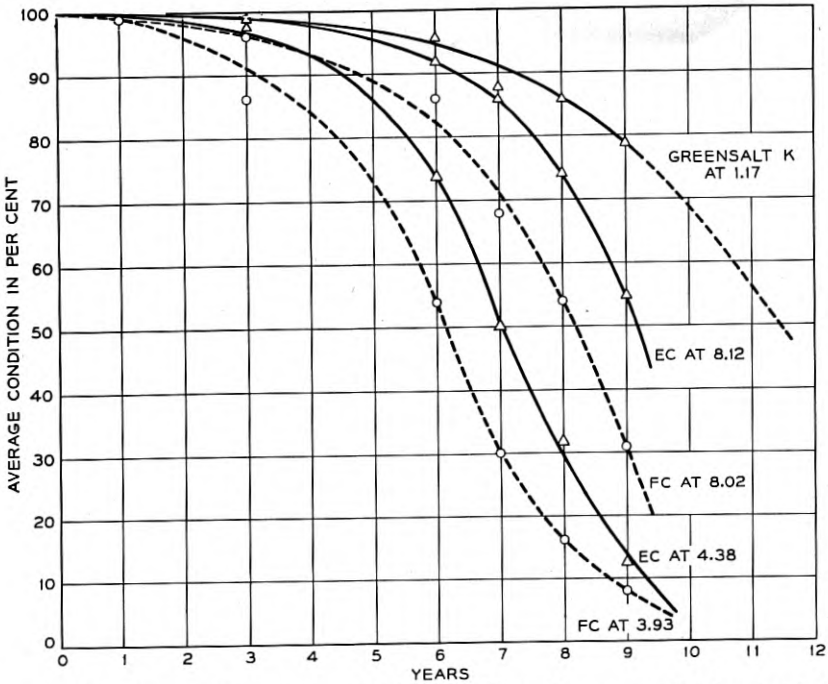


Fig. 19—Depreciation curves for ¾-inch southern pine sapwood stakes treated with creosote, BTL No. 5283, empty-cell and full-cell (toluene dilution) processes, and with greensalt K; Gulfport test plot. See text, Tables XII–XIII, and companion Figs. 20, 21 and 22.

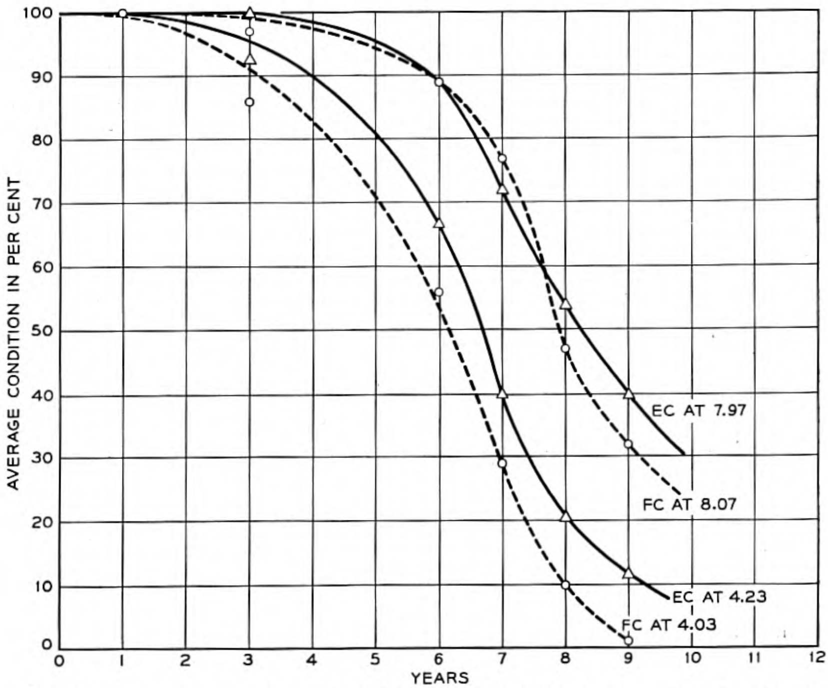


Fig. 20—Depreciation curves for ¾-inch southern pine sapwood stakes treated with creosote, BTL No. 5286B.

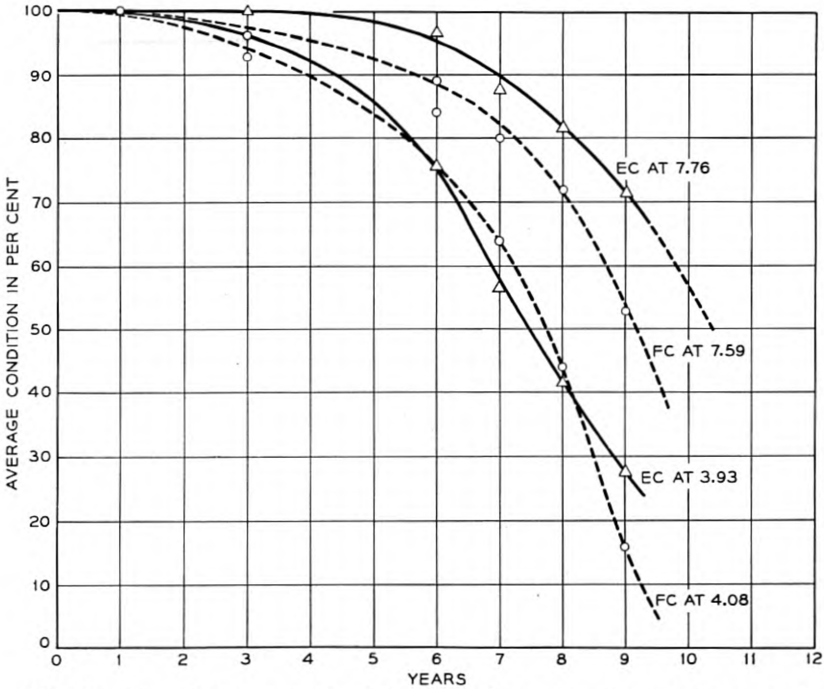


Fig. 21—Depreciation curves for ¾-inch southern pine sapwood stakes treated with creosote, BTL 5286A.

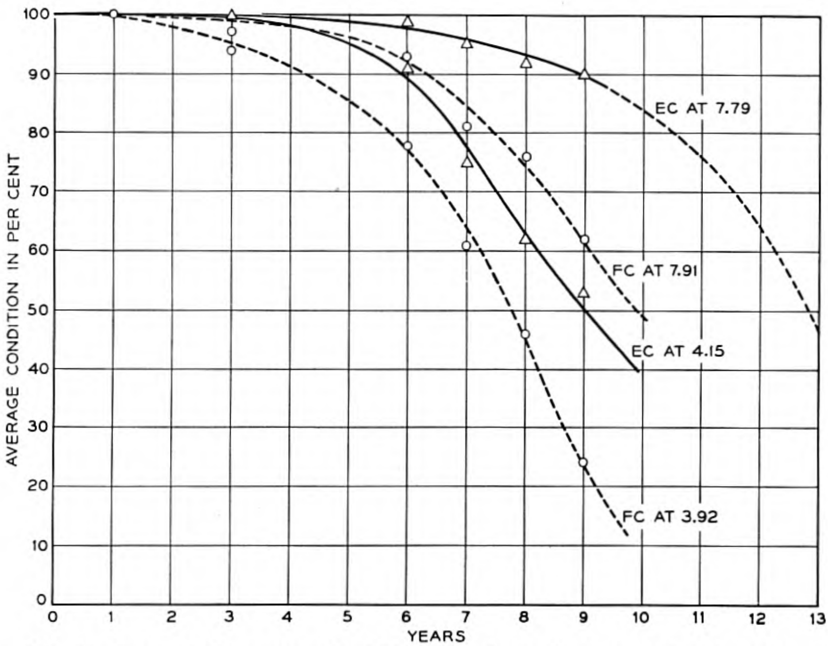


Fig. 22—Depreciation curves for ¾-inch southern pine sapwood stakes treated with creosote, BTL-5285A.

The straight lines in Figs. 23, 24, 25 and 26 are drawn through the average points for per cent condition and retention of the respective 4- and 8-pound groups, and projected to intercept the 100 per cent condition line. The empty-cell data are represented by the solid line and the full-cell data by dashes. The threshold concentrations necessary to prevent all decay are estimated from the intersection of the gradient lines and the 100 per cent condition line. This method assumes that the relation of condition to treatment retention is linear at or near the threshold, providing the average condition points through which the lines are drawn are established by the logarithmic based rating system described. The method also resembles in a way the procedure of the Madison investigators who have used the intersection of straight lines drawn through operational losses and decay weight losses in estimating the threshold retention in creosoted blocks.[39] As in the case of the latter the method would probably be more precise if one had more points for average condition at average retention nearer the thresholds. At any event the system represented by Figs. 23–26 seems to be about the only one that indicates probable thresholds for these particular creosotes and these particular sets of data.

Table XIII — Estimated Threshold Retention and Average Life

¾-inch southern pine sapwood stakes; (see Tables X–XII); 1941 series; Gulfport test plot.

Creosote No.	Treatment	Years in test			Average life-years*	
		3	6	7	"4 lb"	"8 lb'
		Estimated thresholds lb/cu ft*				
5283	EC	9.7	9.7	9.6	7.0	9.2
	FC	9.6	9.8	11.4	6.1	8.2
5286B	EC	8.0	9.8	11.3	6.7	8.3
	FC	9.2	9.4	10.0	6.2	7.9
5286A	EC	7.7	8.3	9.3	7.5	10.4
	FC	12.2	Indet.	11.9	7.8	9.2
5285A	EC	7.8	8.3	8.7	9.0	12.8
	FC	11.8	9.7	11.1	7.8	9.9
Greensalt K	FC (0.57)	—			10.5	
	FC (1.17)	—	1.4	2.1		11.4

* See text for additional data on method used in estimating the threshold retentions at treatment and the average life figures.

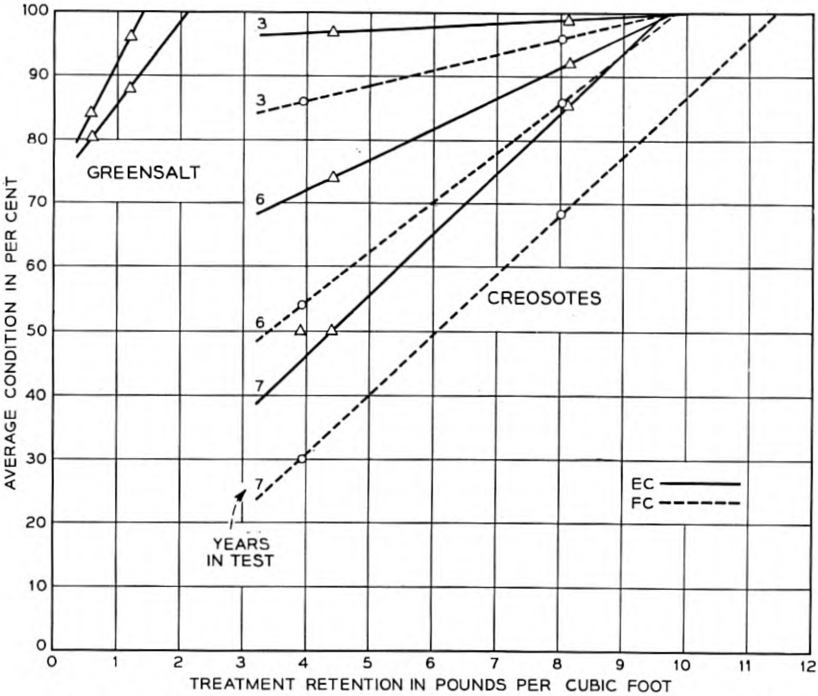


Fig. 23—Theoretical lines for estimating threshold retention for creosote No. 5283, in ¾-inch southern pine sapwood stakes, empty-cell and full-cell (toluene dilution) treatments, and for greensalt K; Gulfport test plot. See text, Tables XI and XIII, and companion Figs. 24, 25 and 26.

The tendency for the 7-year lines to fall off to the right shows the effect of increasing decay in the 8-pound group. In the higher residue creosotes the earlier decay of the toluene dilute 8-pound treated specimens — that is, specimens that were treated below the threshold retention — tends to pull the lines so far down as to spoil their usefulness as tools for estimating thresholds. Obviously the slopes of the lines will be influenced by the depreciation rating of the 4-pound as well as the 8-pound groups. Furthermore it would appear that the utility of the specimens treated with toluene-creosote solutions for estimating thresholds does not extend much beyond the 6th year of exposure under conditions such as those prevailing at the Gulfport plot.

The estimated thresholds at the 3-, 6- and 7-year inspection periods, and the estimated average life values for the different groups are summarized in Table XIII. The average life is estimated from the intersection of the depreciation curves and the 50 per cent condition lines. There

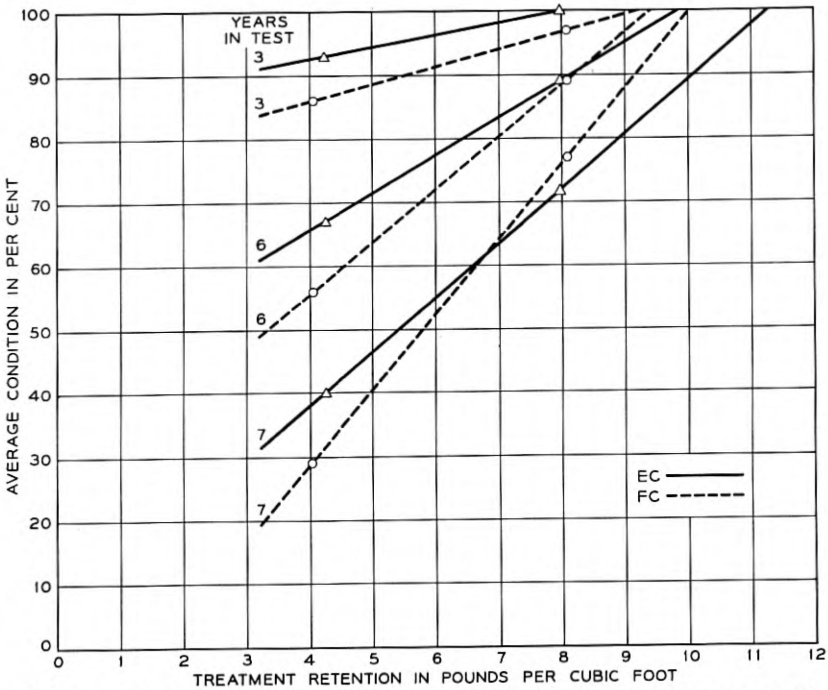


Fig. 24—Theoretical lines for estimating threshold retention for creosote No. 5286B; ¾-inch southern pine sapwood stakes.

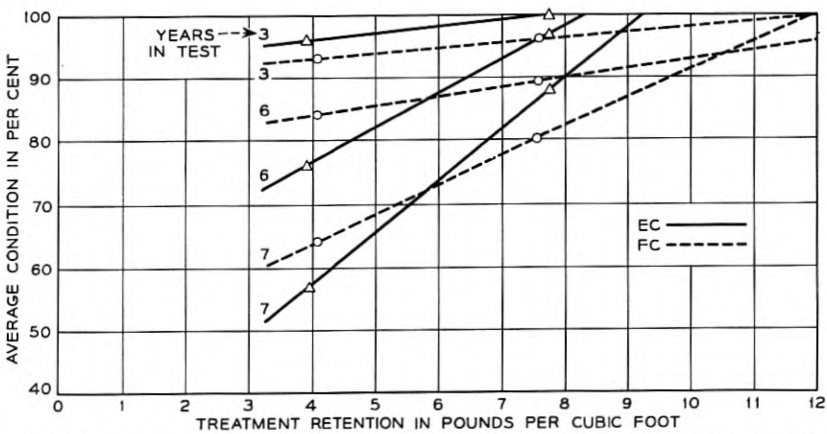


Fig. 25—Theoretical lines for estimating threshold retention for creosote No. 5286A; ¾-inch southern pine sapwood stakes.

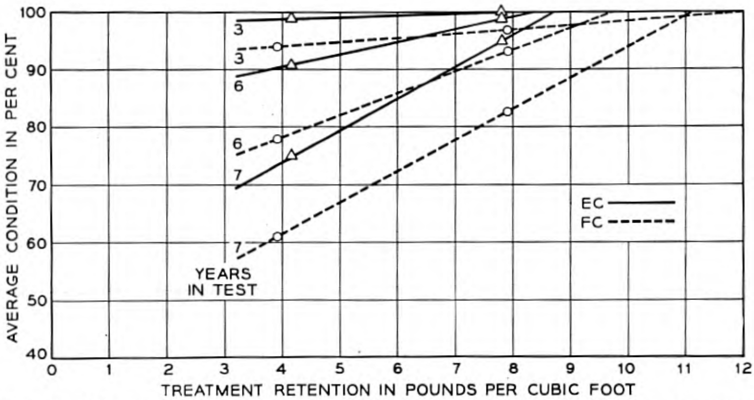


Fig. 26—Theoretical lines for estimating threshold retention for creosote No. 5285A; ¾-inch southern pine sapwood stakes.

will inevitably be some difference of opinion as to which level to use. In the case of ¾-inch treated stakes it is quite evident that the preservative is no longer functioning effectively if the stakes have reached a decay rating of 7 or less. In such small specimens it is questionable whether any purpose is served by allowing them to stay in the ground under the conditions at the Gulfport test plot until they practically fall over by being completely destroyed at the ground line.

Anyone who has worked with small test plot specimens will appreciate the many difficulties in the way of establishing standard procedure for determining the "failed" point or the end point of specimen life. On somewhat larger stakes Rennerfelt[88] has used a strength testing apparatus. To test the fitness of small poles in line some Associated Operating Companies have used a spring scale dynamometer which is slipped onto the base of a pike pole. In actual utility plant experience it is obvious that it is impossible to wait for the complete decay of the wood unit. Elaborate tables have been worked out as guides for pole line inspectors to let them know how far decay can go under given load conditions before a pole has to be removed from line. Generally speaking, such removal must occur at some period well in advance of the time that the specimen would have rotted clear through at the ground line. In the writer's opinion it might be preferable to estimate the average life for ¾-inch stakes from the point where the depreciation curve passes downward through the 60 per cent condition line, leaving *all* the units in any given series in test until that time, except of course the stakes that may have actually rotted off earlier.

Study of Table XIII indicates clearly that both threshold and average life estimates for any given set of small specimens will vary, depending *on the time at which the estimates are made.* Taking the 6-year inspection data as perhaps representing the best figures from which to make threshold and average life estimates of this kind, it appears that in the case of all of the four creosotes something more than 8 pounds of creosote per cubic foot was necessary to protect the ¾-inch stake specimens against decay. There seems to be no material difference in the performance of creosotes 5283 and 5286B as far as the estimated thresholds are concerned. For both the empty-cell and full-cell treatments it appears to be somewhere in the neighborhood of 9.5 pounds or above. In the case of the higher residue oils 5286A and 5285A there appears to be a significant difference between ratings obtained from the empty-cell specimens and from the toluene dilute full-cell specimens. Estimates of average life from the 8-pound empty cell specimens appear to be significantly higher than estimates from the 8-pound full cell toluene-creosote specimens, except in the case of oil No. 5286B (Table XIII). The estimated thresholds for the full-cell toluene-creosote specimens lie within the same general magnitude as the retention at treatment thresholds found in the soil-block tests. (Cf. Tables XIII and XXXV).

How far one is justified in comparing straight 8-pound empty-cell treatments and 8-pound full-cell toluene-creosote treatments in ¾-inch stakes is still not clear. It certainly cannot be done without taking into account the much greater variability in the retentions in empty cell stakes and the different but difficult to describe variations in distribution of the creosote from outside to inside, particularly in the case of empty cell treatments with high residue oil. Unless the empty-cell stakes are carefully selected within limited variations from the average retention, it has been found that the empty-cell coefficient of variability for retention in an 8-pound treatment, for example, may run as high as 35–40 per cent, against a coefficient of 8–10 per cent only for companion full-cell treatments with toluene-creosote solutions. The comparisons among the latter treatments appear to be more rational and fairer; and they may give a truer picture of effective threshold requirements.

The analysis of the stake test data here presented is intended merely to illustrate one set of procedures that may be used in the interpretation of small stake tests. The four sets of data are part of a much larger group of results that are being worked up for publication. Among the latter there are numerous lots indicating that truly protective thresholds of creosote for ¾-inch stakes lie somewhere between 10 and 12 pounds per cubic foot, which is not far out of line with the estimates given above. Final analysis and publication will either confirm or modify such estimates.

EVALUATION BY TREATED POLE-DIAMETER POSTS IN TEST PLOTS

Reference was made in the introduction to the fact that Bell Telephone Laboratories' experience over the last quarter of a century in the evaluation of preservatives by the use of pole-diameter posts in the Gulfport test plot was reviewed by Lumsden[76] in April, 1952, before the American Wood-Preservers' Association. A supplementary analysis and interpretation of some of the same evidence is attempted in the following paragraphs.

Data for 185 of the posts, the time of placement and the number of years in test, and the general condition of the posts as of the 1950 inspection, are shown in Table XIV. The data for the individual posts of these

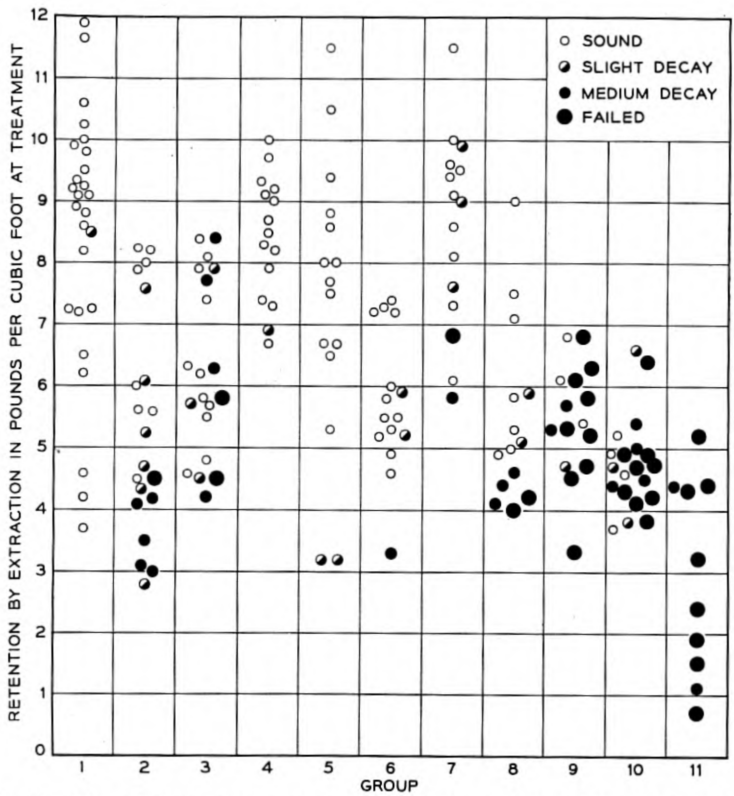


Fig. 27—The relation of creosote retention in lb/cu ft at treatment, by toluene extraction, and the rated condition of pole-diameter southern pine test posts, by post groups; 1950 inspection; Gulfport test plot. See text and Table XIV.

TABLE XIV — CREOSOTED SOUTHERN PINE TEST POSTS

Groups, years in test, and condition at 1950 inspection; Gulfport test plot.

Group*	Year placed	Years in test	n	Sound	Decaying		
					Slight	Medium	Failed
1	1931	19	26	25	1	—	—
2	1931	19	20	8	6	5	1
3	1931	19	20	11	3	4	2
4	1936	14	15	14	1	—	—
5	1935	15	15	13	2	—	—
6	1935	15	15	12	2	1	—
7	1935	15	15	10	3	1	1
8	1935	15	14	7	2	3	2
9	1935	15	15	3	1	2	9
10	1936	14	20	4	3	4	9
11	1935	15	10	—	—	2	8
Totals.			185	107	24	22	32

* For analyses of creosotes, see Table II in Lumsden's report[76], Am. Wood Preservers' Assoc., Proc., 1952.

TABLE XV — CREOSOTED SOUTHERN PINE TEST POSTS

Relation between average treatment retentions by zones, by toluene extraction, and condition of the posts at the 1950 inspection, Gulfport test plot.

Condition	Group	n	Average retention at treatment lb/cu ft					
			Zones, inches				Remainder of treated sapwood	Whole cross section
			Outer ¼	Next ¼	Next ½	Next 1		
Sound	5–10	64	13.1	9.9	8.5	6.4	5.2	7.4
Decaying	5	2	9.7	7.3	4.9	—	3.2	4.2
	6	3	8.8	5.5	4.8	—	4.2	4.8
	7	4	13.1	11.6	11.8	11.1	6.5	8.1
	8	5	10.9	8.7	7.4	5.0	2.9	5.8
	9	3	12.7	9.2	7.0	2.9	2.9	5.2
	10	7	10.1	5.2	3.9	3.6	3.5	4.3
	Total	24	10.9	7.7	6.5	4.3	3.9	5.4
Failed	7	1	13.7	9.2	11.7	—	4.7	6.7
	8	2	9.2	7.4	6.8	—	3.6	5.1
	9	9	13.3	10.0	6.2	3.7	3.6	5.6
	10	9	9.2	4.9	4.0	3.2	3.3	4.2
	Total	21	11.1	7.4	5.6	3.5	3.5	5.0
	Overall	109	12.3	8.9	7.5	5.5	4.6	6.5

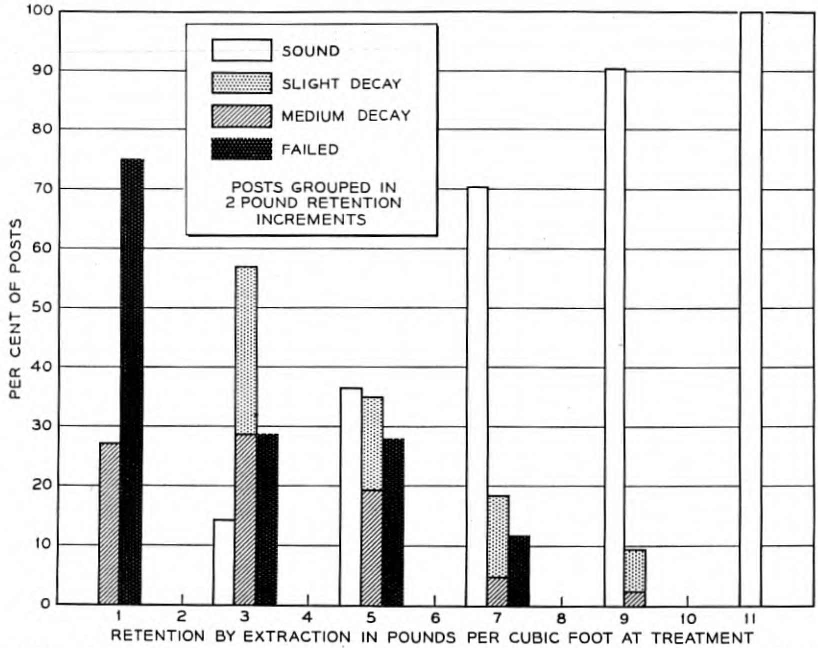


Fig. 28—The relation of creosote retention in lb/cu ft at treatment by toluene extraction, and the rated condition of pole-diameter test posts, by retention groups; 1950 inspection; Gulfport test plot. See text and Table XVI.

eleven groups are shown graphically in Fig. 27. The relation of retention by extraction at the time of treatment to the sound, decaying and failed specimens is clearly evident. All of the failures and the very great majority of the remaining decaying poles are below the 7-pound retention line. Only two cases of medium decay occur above the 7.5 pound line. It should be borne in mind particularly at this point that these retention levels represent the retentions found by extraction in the whole cross sections of the posts *as soon as practicable after* the posts were treated. The significance of these over-all retentions and of the calculated retentions in the outer one-inch layer of the posts will be discussed later. Since the data were calculated in terms of oven-dry weight and volume of the extracted wood the values may be a little high.

For Groups 5 to 10 inclusive data are available on the retention in zones, that is, in the outer quarter inch, the next quarter inch, the next half inch, the next one inch, and the remainder of the treated sapwood. These data were obtained by appropriate cutting of the samples into zones and pooling for extraction the parts that came from the same zones, either in sectors cut from discs or in boring samples. A summary of the distribution of the retentions will be found in Table XV.

If the overall retention data for post groups 1–11 are distributed in 2-pound retention lots the evidence falls into the categories represented by Table XVI and Fig. 28. The inference is clear. One would expect a satisfactory service life if the treatment retention by extraction, based on the whole cross section, was at or above 8 pounds per cubic foot. This is perhaps over-simplification. The variations shown in Tables XIV and XV in the behavior of the posts in Groups 7 and 9, and possibly to a certain extent in Group 8, disturb what looks like reasonably straight reasoning.

In order to extend the reasoning and provide another method of interpretation, let it be assumed that the posts average 8 inches in diameter at the ground line, which is close to their actual size. Calculation of the pounds per cubic foot in the outer inch of such average posts is shown in Table XVII. The influence of posts in groups 7 and 9 on the averages is still evident. If there are differences in the efficiency of the

Table XVI — Creosoted Southern Pine Test Posts

Condition at 1950 inspection by retention groups; 14–19 years in test; Gulfport test plot.

Retention* lb/cu ft at treatment	n	Condition of ground section, number and per cent			
		Sound	Decaying		Failed
			Slight	Medium	
1. 10.0–11.9	10 100.0	10 100.0	0 —	0 —	0 —
2. 8.0– 9.9	42 100.0	38 90.4	3 7.2	1 2.4	0 —
3. 6.0– 7.9	44 100.0	31 70.4	6 13.6	2 4.6	5 11.4
4. 4.0– 5.9	71 100.0	26 36.6	11 15.5	14 19.7	20 28.2
5. 2.0– 3.9	14 100.0	2 14.3	4 28.5	4 28.6	4 28.6
6. 0.0– 1.9	4 100.0	0 —	0 —	1 25.0	3 75.0
Totlas	185 100.0	107 57.8	24 13.0	22 11.9	32 17.3

* The distribution into retention lots was made on the basis of both weight and extraction data, depending on information applicable to the eleven groups in Table XIV.

creosotes in these groups, the differences are not at present real and tangible because they are masked by other factors, among which distribution of the preservative, and retention and penetration variables seem most important. The over-all indications are that in general one should insist on something more than 8, and probably more than 9, pounds of creosote per cubic foot, at the time of treatment, *in the outer inch* of the ground section of a southern pine pole. This, in simple terms, is in line with the conclusions reached about threshold retention requirements from the laboratory tests of creosoted ¾-inch cubes and from test plot results on ¾-inch stakes.

It is much harder to rate treated test posts in terms of per cent condition than it is to rate small specimens. Fig. 28 must be considered therefore as a generalization from the plot inspection data. It must be interpreted with the help of Fig. 27 and the retention by zones data in Tables XV and XVII. The reader who is at all familiar with pole line service records will recognize how much more detailed information there is for these test posts than there generally is for the ordinary pole line. Yet it is practically impossible to draw up precise, unequivocal statements about the results in certain of the post series. The conclusions must be broad.

The structure of the wood in a pole, the distribution of moisture content, the variation in density of the wood from the outside to the inside annual rings, and the distribution of creosote from the outside toward the inside all go to make up a resultant that is obviously complicated. Some of these factors have been reduced to a schematic figure (Fig. 29)

TABLE XVII — CREOSOTED SOUTHERN PINE TEST POSTS

Relation between average treatment retention in outer inch of sapwood and condition of posts at the 1950 inspection; Gulfport test plot.

Group	Average retention at treatment in outer 1 inch of sapwood					
	Sound		Decaying		Failed	
	n	lb/cu ft	n	lb/cu ft	n	lb/cu ft
5–10	64	10.14*	—	—	—	—
5		—	2	7.13	—	—
6		—	3	6.09	—	—
7		—	4	12.11	1	11.61
8		—	5	8.71	2	7.90
9		—	3	9.38	9	9.15
10		—	7	5.95	9	5.95
Totals	64	10.14	24	8.03	21	7.59

* All calculations are made on the basis of an 8-inch diameter at the point of sampling.

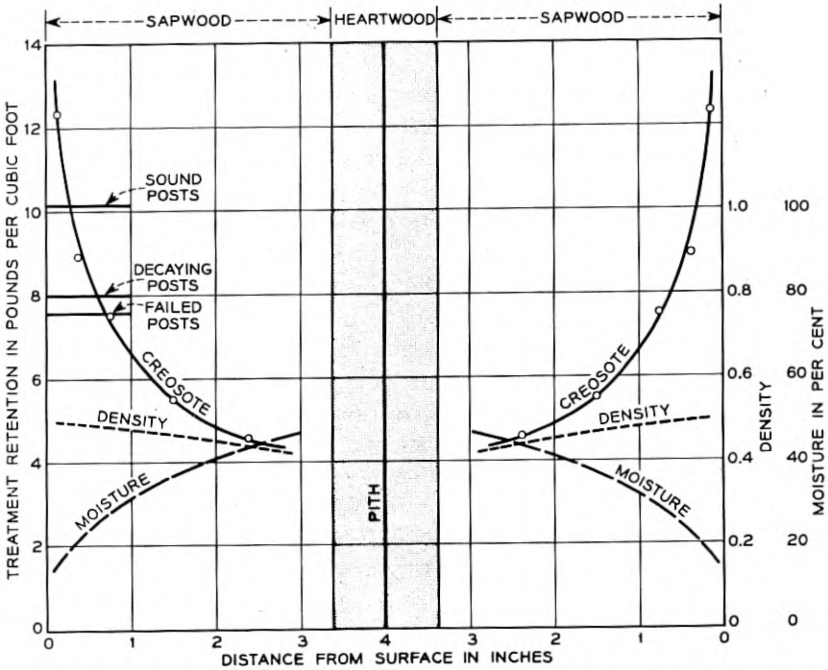


Fig. 29—Schematic diagram of a longitudinal section of a creosoted southern pine test post, showing average distribution of creosote in lb/cu ft; average density (oven-dry weight and volume as tested); average moisture content, oven-dry wood basis; and the retention levels in the outer inch of sound, decaying and failed specimens. See text.

representing a longitudinal section of an 8 inch diameter southern pine test post. The sound, decaying and failed retention levels calculated for the outer one inch (Table XVII) of such a diagrammatic post are also shown. When one considers all the variations in the wood itself and in the treatment, the site and environmental conditions where the pole is used, and the probability of the incidence of decay, the above general conclusions with respect to necessary retention seem reasonable and practicable. To refine and narrow the conclusions by repeating the post tests with the same creosotes — if one could get them — would certainly require the use of at least some posts with higher retentions and an obligatory assay of all of the treated specimens to be sure that the required retentions were actually in the wood in the right places. Apparently practical answers to questions about such required retentions can be answered very much more quickly by repeated series of soil-block tests in which many of the variables can be controlled.

EVALUATION BY POLE TEST LINES AND BY LINE EXPERIENCE; SERVICE TESTS

In 1932 a condensed report[84] of American Telephone and Telegraph Company experience with creosoted pole test lines, including brush, open-tank and pressure treatments, and covering northern (eastern) cedar, southern white cedar, chestnut and southern pine poles, ends with the following conclusions, among other generalizations:

"Because of the fact that many of the test specimens are still in service, conclusions can be reached only in the case of some of the less important problems whose solution has been sought. The possibility of extending the life of poles through preservative treatment is abundantly demonstrated, but the capabilities of the more effective processes of treatment (studies) can as yet only be estimated.

" . . . the indications are that in the cases not already affected, the beginning of decay attack will mainly be dependent upon changes in the quantity and the composition of the preservative retained in the individual poles."

This was only twenty years ago. Even in 1932 this service test report was more valuable as history than as a technical base for treatment specifications. The report reconfirmed the common knowledge that durable timbers like northern cedar and chestnut were better line units if properly butt-treated with creosote, and that adequate penetration and absorption of creosote were essential factors in the economy of pressure treatment of non-durable southern pine poles. However, by 1932

(a) The type of creosote used for treating the relatively large heart and small sapwood southern pine poles for the famous Washington-Norfolk and Montgomery-New Orleans lines was no longer available commercially;

(b) The type of virgin pine timber used was becoming scarcer and scarcer;

(c) The supply of commercial chestnut pole timber was just about completely exhausted;

(d) New and improved methods of butt treatment involving the use of machines for incising the ground section were being applied to northern cedar and to western cedar poles, and besides, the use of northern cedar was gradually shrinking to limited areas in telephone plant in the Northeastern and Lake States; and

(e) Because of vastly increased competition in the pole treating industry, and because the chestnut supply had failed, and because the excessive bleeding of creosote from the old style "12-pound" full-cell southern pine poles — like those in the cited lines — made them unacceptable either as replacements for chestnut or for new construction in large sections of the northern and central telephone plant areas, and for basic reasons of economy, new specifications for creosoted southern pine poles were being issued. In preparing these specifications the Laboratories reduced the retention requirements to 8 pounds of creosote per cubic foot and provided for treatment by an empty cell process with a view to providing a clean pole, and set definite quality limits on penetration to insure an economic service life. These moves were made after careful experiment, and after analysis of 8-pound treatment results. The experience in the pole test lines had an indirect effect rather than a determining effect on the proposed new penetration requirements.

The Laboratories by 1932 had been operating the Gulfport test plot for 7 years and the Chester test plot for 3 years. Research and development programs on laboratory and test plot evaluation procedures[123, 124, 129] were well under way. The aim of this broad program was to determine the necessary requirements for preservatives, for retention, for treatment and for penetration *before* the poles were placed in line; which is quite different from the philosophy of depending on service tests or records to reveal at some later date what was done wrong in the first place.

The reader should not get the idea that service tests can be dispensed with entirely. Material in service in the telephone plant is always under observation, casual or intensive, and the development of any obvious faults is corrected by the application of results of more and perhaps better laboratory experiments. In very many cases the faults are discovered in the laboratory or test plot before they are found in the field. However, one can never brush aside the insistent and determining effect of long and satisfactory field experience with treated wood.

For example, when the behavior of creosoted poles in line is good the service tests take on what seems to be the outstanding characteristic of their present function, namely that of a comforting confirmation of previous conclusions. The results of some of the Laboratories' analyses of the relation between penetration and decay in creosoted poles were published in July, 1936, in this Journal[29] and again in 1939.[3] These results confirmed the previous actions involved in specifying an empty cell treatment and a retention of 8 pounds of creosote per cubic foot. The poles were clean, they were being accepted generally for line construction, and the incidence of infection and probable early failure seemed to be about in line with anticipation.

Emphasis in the last two papers cited was on penetration, which is easily determined, relatively speaking, and not on retention. The relation of low retention to decay showed up in the Gulfport plot,[76] and it was explained by Laboratories' extraction and analysis of the creosote in the decaying or failed test stakes and test posts. One may say that such results confirmed suspicions; and they did, because creosoted poles in line — which were "related" to the decaying posts at Gulfport — were found on inspection to be behaving badly. It is believed that corrective measures in the way of supplemental ground line treatment were taken in time to give the poorly treated poles a reasonably satisfactory life.

In another striking set of circumstances, however, a number of cases of unprecedented premature failure of pine poles in line treated with a mixture of creosote and copper naphthenate petroleum revealed that the preservative solution had gone out of control and that in consequence the poles had not received the specified protective amounts of copper. Bell Telephone Laboratories' analyses of parts of the decaying poles showed that the decaying areas contained less than, and the sound areas more than, the Madison soil-block threshold of approximately 0.08 lb/cu ft of copper as metal.

These poles were immediately traced back to the supplier by their brand label. Present day pole treatment specifications in general require branding of each pole unit with symbols or code letters for the supplier's plant, the species of timber, the year of treatment, and the class and length of the pole. Each such pole becomes automatically a unit in a sort of universal service test in Bell System pole lines.

Service tests and service records are a significant part of the over-all process of evaluating wood preservatives, but insistence on service records as the most important criterion simply perpetuates the reputation of established preservatives and forever blocks or seriously impedes the development of new and promising materials. In the very nature of the case the results of service tests can be stated in the form of broad generalizations only. The truly technical approach must be made through better methods of measuring the effectiveness of preservative materials by accelerated field tests on a sufficiently large number of small stakes or by controlled laboratory experiments such as the soil-block tests.

DISCUSSION

There are a number of things in connection with the soil-block test procedure, its interpretation and the correlation of its results with the results of other evaluation methods that require further discussion. For example, questions have been raised vigorously on such matters as:

(a) The effect of variation in growth rate and density of the wood;

(b) The size and shape of the test block;

(c) The use of toluene or any other diluent with creosote to get low retention;

(d) The distribution of the preservative in the block;

(e) The practice of heat sterilization of the creosoted blocks;

(f) The whole philosophy of the weathering procedure; and

(g) The methods of control assay, such as weight, creosote extraction, analysis of extracted creosote, and lime fusion chloride determinations as applied to pentachlorophenol.

None of these questions can be simply brushed aside, in spite of the fact that some of the points raised have a little of the nature of quasi-technical road blocks that may temporarily slow the approach to *any* standard laboratory test for creosote. Experimental work now under way at Madison[41] and our own laboratories at Murray Hill will help answer some of the questions. The following discussion will at least explain the nature of the problems involved.

Density and Growth Rate

Data on the relation of block density and absorption at treatment have already been presented (Tables I, IV, V and VI). It is not yet evident that either density or growth rate has any material effect on the treatment retention thresholds for creosote when the tests are run on outdoor (or equally depleted) weathered blocks, and when the steps in the gradient retentions are properly spaced. When some present experiments are finished the matter may be resolved more conclusively.

Size and Shape of the Test Blocks

Some critics of the soil-block test have objected rather strenuously to the ¾-inch cube because the two transverse faces are so close to each other. The inference is that either (a) the preservative, and in this case creosote is usually meant, is lost too rapidly through the end grain of the wood or (b) that in the evaporation process, or in the course of the weathering procedures, the preservative may be concentrated on the transverse faces. The validity of these contentions is under investigation. However, the importance of block size and shape can be greatly exaggerated.

One need not assume, in order to plan for comparative tests of oil preservatives, that a block must or can be cut so that creosote will be lost from it at exactly the same rate or manner as creosote may be lost radially from the concentric annual rings of a post or pole. Block shape has varied almost as much as the test procedure employed. In a laboratory working plan for a study of the "Efficiency of Various Wood Preservatives," with the subtitle "The Efficiency of Various Wood Preservatives in Resisting the Attack of Fungi in Pure Cultures," signed by C. J. Humphrey, dated July 10, 1913, and approved by Howard F. Weiss, at that time director of the young U. S. Forest Products Laboratory, Humphrey proposed the use of eastern hemlock heartwood, cut into pieces measuring 1¼ x 1¼ x 2 inches, which were to be treated and quartered longitudinally "*to reduce the size sufficiently to allow their introduction into the* (*Erlenmeyer*) *flask*." (Writer's italics.)

Untreated wood blocks were to be used as culture media, and thirty days after inoculation the test pieces were to be put into the flasks and shaken up with the inoculated blocks. The test fungus was to be *Lentinus lepideus*. The culture period was to be 9 months, with culture blocks and mycelium "kept in a condition moist and warm enough for active growth" throughout the test period. Coincidently, Humphrey[61] and his colleagues were developing the broad foundation for the Petri dish toxicity test. The proposed wood block tests never reached a really satisfactory experimental level for treated wood, although they were employed for comparative natural durability tests.[57]

Breazzano[20, 21] in the same year experimented with blocks of beech wood cut from treated ties and measuring about 9 x 2 x 1 cm. As has been pointed out earlier in this paper he recommended blocks measuring 4 x 2 x 1 cm as a tentative standard, and later, in 1922, blocks 4 x 4 x 2 cm. were accepted as standard in Italy.[22] In the case of the latter the larger faces were to be transverse faces, cut across the grain.

Howe[56] reports tests of small sticks (blocks) of southern pine, measuring ⅜ x ⅜ x 6 inches, that were treated with salt preservatives and later inserted in "8-inch sterilized test tubes containing about 10 cc of standard malt agar." He also used sets of four small sticks in Petri dishes, placing them on 10 cc of nutrient agar medium that covered the bottom of the dishes. The fungus used was called *Fomes annosus*. To supplement these tests he mixed ground-up treated wood in different concentrations with agar media, and tested his mixes against this same *Fomes annosus* and a number of other wood-destroying fungi.

Howe and Curtin[32] were working on a broad plan aimed at correlating laboratory tests with test plot tests made, e.g., at Matawan, N. J., and with experience in line.

Snell[107] argued that one might obtain good growth by placing thin blocks of wood in tubes with agar, but that the growth of the test fungus on the agar and on the wood might be influenced by diffusion of the preservative into the agar medium. He proposed the use of thin plaques of wood, measuring ⅛ x 3 x 3 inches. These could be soaked to any required concentration of preservative and then tested by placing them over wet filter paper in Petri dishes. The test specimen would be supported just above the wet filter paper on sterilized wood strips. Inoculation was to be by the simple process of placing a small square of agar plus the growing test fungus directly on the upper surface of the test "block." Similarity to natural conditions and the control of moisture content of the wood were considered to be decided advantages for the method.

Cislak (see discussion of Snell and Shipley[108]) used similar plaques of wood measuring 4 x 4 inches and about 1/28-inch thick for experiments on evaporation and permanency of creosote.

Rhodes, Roche and Gillander[90] used blocks measuring ½ x ½ x 3 inches.

The European standard[33, 45] block measures 5 x 2.5 x 1.5 cm, with the long axis in the direction of the grain.

Schulze, Theden and Starfinger, in addition to the standard block, used "half" blocks, i.e., blocks measuring 5 x 2.5 x 0.75 cm. All factors considered, they do not regard the thinner block as an advantage, and they have held to the standard size.[54 (1)]

Lutz[77] in 1935 suggested the use of 2 x 2 x 5 cm. blocks, with the long sides dressed parallel to the fibers of the wood. He also used blocks measuring 1 x 1 x 5 cm.

Alliot[2] favored blocks measuring 5.0 x 1.0 x 0.5 cm. in the longitudinal, tangential and radial directions, respectively, for a French standard test.

In his recent tests[51, 52] Harrow has used 1½ x 1 5/16 x ⅞ inch blocks.

Sedziak[106] uses ¾-inch cubes cut from ¾-inch stakes *after treatment.*

The National Wood Manufacturers' Association[109] standard block size is 1.25 inches on the radial surface, 1.75 inches on the tangential surface and 0.25 inch thick, i.e., in the longitudinal direction of the grain.

Various size blocks have been used at Madison in soil-block tests of natural durability, but two sizes only have been employed commonly since 1944 in the above mentioned soil-block tests and agar-block tests. The soil-block is the ¾-inch cube, generally drilled with a ⅛-inch hole in the center of a tangential face. The agar-block was cut with two broad transverse surfaces measuring ¾ x 1½ inches, and with a distance along the grain of only ⅜ inch; and it was not drilled.[41] The Madison agar-block is basically the same sort of a block as the one described in the previous paragraph, and it resembles the Breazzano blocks[21] as far as maximum transverse surface exposure is concerned.

The list is incomplete, but it serves to illustrate one important point about laboratory evaluation tests, and that is the inevitable variation that creeps naturally into explorative research. In the writer's experience and opinion much of the variation in block size has been the result of a sort of forced adaptation, on the part of the investigator, to the size and shape of his laboratory glassware, coupled with certain practical problems of block procurement and manufacture. Obviously it is easy to use thin sticks in test tubes, thin sticks or plaques in Petri dishes, and relatively flat blocks in Kolle flasks. The ¾-inch cubes, which in essence, as has been stated before, are simply sections of the ¾-inch stake, handle easily in the soil-block cultures.

Criticism of the shape of the ¾-inch cube did not become pointed until after the publication of the first papers[95, 35] on the Laboratories' cooperative work with the Madison laboratory. It is argued — as mentioned before — that the evaporation of creosote from such blocks is unfairly rapid. However, the losses reported by Rhodes et al,[89] which will be discussed later, indicate that separation of the transverse faces will not prevent creosote evaporation. Assuming properly calculated gradient retentions, and the use of weathered blocks, it does not appear likely that the shape of the blocks — if kept constant within any given comparative test series — will affect the location of the treatment threshold retention.

Toluene as a Diluent for Creosote Treating Solutions

This subject is most controversial in this country. The use of acetone or chloroform has been widely accepted in Europe, but Schulze and Becker[104] warn that the use of any diluent may change the rate of evaporation of given creosote fractions, and may affect the rate of evaporation of whole creosote. However, there are points in the debate which can be stressed, namely:

(a) The treatment of test blocks to low and uniform retentions without a diluent is extremely difficult, if not practically impossible;

(b) The rate of loss of creosote by evaporation, and any change in the character of that loss that may result from toluene dilution might be evident in freshly treated blocks but not in weathered blocks; and

(c) The volatile fractions of undiluted creosote are lost fairly rapidly from small saplings[125] and from test blocks; and it is assumed that the use of toluene does not cause loss in the fraction above 355°C.

Discussion of Item (c) will be resumed in the section on weathering. In view of the perplexing character of this toluene dilution question steps are being taken to find out what happens. In the meantime the toluene diluent which has not been found to exert any measurable toxic effect in the soil-block tests at Madison[41] must be used to secure gradient retentions down to and below the threshold level.

The Distribution of the Preservative in the Block

The opinion has been expressed by a number of investigators[14, 22, 23, 100] that it is difficult to secure uniform distribution of the preservative in the test block, and that evaporation would cause concentration of certain preservatives on the block surface, where the toxic material might influence the behavior of the blocks in the test cultures. If such a concentration occurs the result would be to fix the threshold at lower over-all treatment retention than it would be if the preservative were not concentrated, for example, at the transverse faces. One is likely to agree that uniform distribution of some creosotes might be difficult without the use of a diluent, particularly in the low retention groups. Rhodes et al[89] were apparently satisfied that they got a fairly good distribution with a 16.7-pound retention. Experience at Madison and at Murray Hill indicates that the blocks are saturated with the toluene-creosote solution or with the toluene-penta-petroleum solution; and this confirms the ideas of Schulze, Theden and Starfinger[54 (1)] on the way oven-dried test blocks take up the treating solutions.

Data on the distribution of residual creosote in weathered blocks is presented in Table XXX, and discussed in the section on weathering. Preliminary assay of blocks treated with penta-petroleum, (a) just after treatment, (b) after weathering and (c) after testing show that the pentachlorophenol concentration is slightly *lower* at the transverse faces than it is in the middle section of the blocks. The same seems to hold true for the copper metal in blocks treated with copper naphthenate in toluene-petroleum.

It should be pointed out here that preservatives are rather fortunately distributed in treated wood in such a way that the outer fibers or annual rings normally contain much higher concentrations of the toxic material than is found in the inner fibers. (See Tables XV and XVII and Fig. 29).

Heat Sterilization of the Treated Blocks

In pure culture test experiments some form of sterilization must be used to avoid contamination by other organisms than the test fungi. Most frequently such sterilization, for the minimum necessary time, is accomplished by flashing the treated block through a hot flame or by steaming at 100°C and atmospheric pressure. Either of these procedures might cause a measurable loss of volatile creosote fractions from freshly treated blocks; but such losses appear to be negligible in the case of weathered blocks. At Bell Telephone Laboratories control data on these possible losses are being determined by extraction of control blocks after the sterilization phase. Such blocks are run through all the steps in the bioassay procedure up to planting in the soil-cultures. It will be recognized by anyone familiar with pressure treating methods that the 100°C temperature, usually held for 15 minutes only, represents a much gentler set of conditions than the after-treatment steaming for several hours at 240–259°F which is permitted in many specifications for creosoted poles.

The Weathering of Creosote and Creosoted Wood

Creosote is a remarkably good wood preservative, and nothing in the following paragraphs is intended to detract from its reputation in that respect. Sometimes the creosote oozes or bleeds from the surfaces of treated units such as poles and crossarms, especially on hot, sunny days; and when such bleeding occurs the treated material is unsatisfactory for use in many parts of the telephone plant. In order to prevent bleeding difficulties and the consequent unhappy employee and public relations that result, the retention requirements have been held down to the commercial standard level of 8 pounds per cubic foot for southern pine poles, and the residue above 355°C limitations on the creosote have been kept in actual practice at 25 per cent or below except when war or post-war emergencies have interfered.

The creosotes vary in the proportion of readily volatile materials they contain, and these materials are lost from creosoted wood — largely by evaporation — under many different use and exposure conditions. The general facts about such losses have been reported over and over again since the turn of the century. The significance of such losses is still not broadly understood or appreciated. Their possible bearing on the weathering procedure to be used in the soil-block tests is of fundamental importance. Paraphrasing the quotation from Schmitz[99] cited in an earlier paragraph: It is really necessary to know how much creosote to inject into wood to allow for loss by volatility and to insure a residual of the preservative, remaining in sufficient amount, to protect the wood for an economical service period. Bell Laboratories has been deeply concerned with the question whether it is practicable under commercial conditions to specify enough retention at treatment to provide the necessary protective residual and at the same time require clean, satisfactory, treated poles and crossarms for Operating Company use.

General Considerations; Creosote Fractions

Before presenting definite evidence of creosote losses this seems as good a place as any to refer briefly to investigations that have been aimed at discovering what components, volatile or relatively stable, give creosotes their properties of toxicity and permanence. Three articles by Martin,[80] Rhodes[90] and Mayfield[81] are the latest American papers covering the general subject, the two former dealing with the technology of hydrocarbons and creosotes and creosote production and the latter reviewing the results of tests of whole creosotes and creosote fractions.

Teesdale's short report[113] in 1911 is one of the earlier records in this country of experiments aimed at determining the loss of creosote fractions from treated wood. He used a creosote with 49 per cent distilling below 250°C and a residue above 320°C of 28 per cent, which would have been about 20 per cent at 355°C. This oil was fractionated into 5 parts, I, to 205°C; II, 205–250°C; III, 250–295°C; IV, 295–320°C; and the residue above 320°C. These five fractions and a sample of creosote with a similar distillation range were used to treat air-seasoned pieces of *Pinus taeda*, mostly sapwood, cut 2 feet long from 5 to 6-inch diameter peeled posts. The retentions were about 18 pounds for the numbered fractions, 15 pounds for the residue, and 21 pounds for the whole creosote.

The treated pieces were open-piled in the laboratory for two months, with temperatures running from 60 to 80°F. At the end of that period the per cent losses of the five fractions and the creosote were respectively 34.7, 21.3, 15.9, 6.2, 4.0 and 5.4. The results were in line with expectations. The test period was short, and there were no outdoor weathering factors. Teesdale notes that the loss of the whole creosote was about the same as the losses in the two higher fractions, III and IV; and that a proportionately composited sample of the five fractions lost at the rate of fraction III, the total at the end of the two-month period being 15.8 per cent or about *three times* as great as for the whole creosote.

Loseby and Krogh[73] reported in 1944 on outdoor weathering tests of creosoted wood blocks; and they compared the weight losses in the blocks with evaporation losses from open Petri dishes. The creosote used was a relatively low residue oil produced at Pretoria. The residue above 355°C was 19.95 per cent, the specific gravity at 38/20°C was 1.088, and the amount distilling to 235°C was only 4 per cent. The test blocks were planed pieces of light weight *Pinus insignis* measuring 6 x 1½ x 1½ inches. The per cent moisture at the time of treatment was 10.3 per cent.

The creosote was fractionated into four parts; Fraction I, up to 270°C; Fraction II, 270–315°C; Fraction III, 315–355°C; and, Fraction IV, the residue above 355°C. The blocks were very heavily treated by a hot and cold soaking process with the straight creosote and each of the four fractions to the following average retentions, respectively: 52.2, 45.6, 50.3, 50.2 and 23.7 pounds per cubic foot. These high retentions place the experiments out of line with most of the others cited in this paper; but the South African tests are unique in that they supply evidence on the rate of creosote losses from such high retentions.

The relative order of losses of the materials, with the one having the highest losses first, was the same for the blocks that were hung on wire outdoors and for the open Petri dish samples, namely; Fraction I, II, whole creosote, Fraction III, and the residue. The latter actually showed no significant loss in either block or dish tests, and in the blocks there was a slight increase, possibly referable to oxidation, of a maximum of 2.2 per cent at the end of the three-year period. This gain was gone by the end of the 5½-year test. Fraction III was lost more rapidly from the *blocks* than from the *Petri dishes;* the reverse was markedly evident for Fractions I and II; whereas the pattern for the whole creosote was very similar after the first year. The rounded figures for losses from blocks treated with the whole creosote, at the end of 1, 2, 3 and 5½ years outdoor exposure, were 36, 42, 44 and 47 per cent; and the losses from the open dishes for the same periods were 34, 40, 43 and 48 per cent respectively. One may conclude from these tests and conditions that a loss level of about 50 per cent would have been reached in about six years in blocks treated to a reported 50-pound per cubic foot retention with an undiluted creosote.

Most of the laboratory toxicity tests on fractions have been run by the agar or the agar-block method.[54 (1), 104, 11, 101, 47] The results obtained by Schulze and Becker[104] are cited by Mayfield, who includes as his Fig. 1 a copy of the summary curves prepared by the Berlin investigators. The interested reader can profitably use the rather full excerpts of tabulated results of other investigators given in Mayfield's paper as an introduction to the difficulties of testing creosotes and of interpreting the results of such tests. Some of the main controversial points are brought out by Peters, Krieg and Pflug in 1937[83] who challenge the results of Petri dish agar toxicity tests with results of the German agar-block tests — one of the first such broad comparisons to be made. Broekhuizen[25] published his findings the same year in a comprehensive paper covering agar-block tests on creosotes and creosote relatives and creosote fractions. He discusses his results with different preservatives in relation to their toxic properties, their protective and preservative qualities, and the *permanence* of such qualities, and the bearing of these qualities on practical wood preserving procedures. His discussion of the importance of "weathering" tests and of his own results with such tests make up one of the best, if not the best consideration of this important phase of evaluation tests that had appeared up to that time.

Numerous references to weathering techniques will be found in van Groenou, Rischen and van den Berge;[118] and van Groenou's own paper[117] in 1940 is an excellent short review of previous work, giving his views of the pros and cons of different procedures with emphasis on the essential nature of some test to determine what changes are likely to take place in a preservative as a result of leaching or evaporation.

The following examples will be interesting as illustrations of the different techniques that have been employed in testing the toxicity, or the potential preservative value of creosotes. F. H. Rhodes and Gardner[91] in 1930 determined evaporation losses of creosote fractions from thin pads of dried ground Sitka spruce pulp. The whole creosote and the fractions were introduced in an ether solution. The impregnated pads were placed on top of agar cultures in Petri dishes. The test fungus was the usual *Fomes annosus*, more recently called simply Madison 517. They state that:

"It was found that under these conditions (the Petri dish covers loosely fitting — not water sealed) the more volatile preservatives vaporized from the test specimen, so that at the end of the month only a relatively small portion of the fungicide remained in the pulp."

They were using a domestic creosote with a specific gravity of 1.065 at 38/15.5°C and a residue above 355°C of 21 per cent. They tested fractions of the dead oil from which the tar acids and tar bases were removed, and fractions of the tar bases and the tar acids themselves. They determined percentage losses by evaporation for all lots by letting the treated pulp disks remain in covered Petri dishes for one month at 25°C. All of these reported losses for the dead oil occurred in fractions boiling below 316°C, for the tar acids in fractions with the same upper limit, and for the tar bases in the fractions boiling below 308°C. The losses varied inversely as the boiling range, the greater being in the low-boiling fractions, as was to be expected. The toxic limits for *Fomes annosus* that they determined showed a gradual increase as the boiling point increased, i.e., in agreement with other workers with agar tests they found the lower boiling fractions the most toxic. Their bibliography, with one or two exceptions, covers American articles only.

Rhodes and Erickson,[92] continuing the same general technique, but substituting mechanical pine pulp for spruce pulp, showed that much higher quantities of respective creosote fractions were required to kill *Fomes annosus* in the pulp cultures than in Petri dish agar cultures. They concluded from their experiments that "no one compound in coal-tar creosote is primarily responsible for its preservative power. . . . The fractions from water-gas tar oil are much less effective as preservatives than are those from coal-tar creosote oil. The chlorine derivatives of phenols and creosote and of naphthalene are more toxic to fungi than are the compounds from which they are obtained."

In 1933 Flerov and Popov[48] tested fractions of two creosotes by their soil-block method and compared the results with tests of the fractions emulsified in agar, following American Petri dish procedure in general. They found (English translation by Hildegard Kipp, Forest Products Laboratory):

"On wood the toxicity of the heavy fractions is considerably higher (compared to that of the lower fractions) and the most toxic fractions are those from 315 to 375°C. . . . In tests on wood, in addition to the preservative effect, the effect of the evaporation factor, which is of great importance with oily preservatives, is determined with (more or less) accuracy."

In 1951 Finholt[46] revives the use of emulsion of the creosote fractions in the old Petri dish or flask method, like history repeating itself. The reader can be excused if he senses a degree of confusion.

Creosote Losses

Baechler's 1949 paper[5] on the toxicity of oils before and after aging would be a fitting introduction to this section. The first task is to condense into simple statements or tables some of the available data on creosote losses from treated wood.

Curtin[32] cites Bond on the latter's experimental determination in 1910–11 of creosote losses from thoroughly air-seasoned red oak and maple railroad ties with approximately the same moisture content. Bond reports that:

1. Full cell treated red oak in 200 days between November 1910, and June 1911, lost *19.0 per cent;* and
2. Similar empty cell treated ties lost *52.7 per cent;* and
3. Full cell treated maple ties in 105 days from March to June 1911, lost *13.4 per cent;* and
4. Similar empty cell treated ties lost *23.0 per cent* of the creosote absorbed at treatment.

The losses were determined by weight. He shows that the losses were greater for the lighter treatments, and that in effect the losses were relatively greater for the empty cell than for the full cell treatments. His conclusions appear to hold good through all the subsequent cited data that admit of such comparisons.

Bateman's early, 1912, work[6] on oils extracted from two old piles that had been in service about thirty years indicated some considerable loss — in one case more than 35 per cent in the above water section — and relatively lower losses for water line and below water line sections. This confirmed a generally accepted common opinion. Losses below water line were apparently confined to the fraction distilling below 225°C in the case of the pile which he considers to have been treated with a pure coal-tar creosote. He calls *light* oils those distilling below 205°C. With the exception of the above water section of one pile all the samples of treated wood still contained about 17 pounds per cubic foot. He states:

"The creosote in the pile which was perfectly preserved contained originally at least 40 per cent of naphthalene fractions, a large portion of which remained in the wood. The creosote in the pile which was less perfectly preserved contained little or no naphthalene."

Service records on such oils resemble records from the Washington-Norfolk line and from the specimens examined by Alleman[1] in that the data have little if any bearing — other than historical — on the creosote use problems of today.

Schmitz et al[102, 103] report a loss of 25.9 per cent after five years service in track in red oak ties treated with a 60/40 creosote-coal tar solution, compared to a loss of 17.3 per cent after three years service. The losses, determined by extraction, varied inversely as the boiling range, as was expected. There was very little loss in the 315–355°C fraction, and no loss is indicated for the fraction boiling above 355°.

Bateman in 1922,[7] and again in 1936 (see Discussion and [108]) in connection with his explanation of the relation between the loss of the creosote fraction below 270°C and a formula for estimating the permanence or preservative life of creosote in treated wood, cites the earlier work of von Schrenk, Fulks and Kammerer, and Rhodes and Hosford on creosote losses from southern pine poles in the Washington-Norfolk and Montgomery-New Orelans lines of the American Telephone and Telegraph Company. Certain of the poles in question were installed in 1897 and removed in 1906 after about nine years service in line. The creosote used to treat these poles is reported to have had a specific gravity of 1.022–1.030 at 3°C above the melting point of the oil. The residue above 315°C was about 16 per cent; and the per cent naphthalene was "not less than 40 per cent." Using the pitch residue — the per cent boiling above 315°C — as a base for calculation, along with the change in residue determined from extracted creosote, von Schrenk, Fulks and Kammerer estimated the creosote losses, above and below ground line, for five poles from the Washington-Norfolk line, that are shown in Table XVIII. The poles were old growth longleaf pine, with a high heartwood volume. They were heavily treated — to about 16 pounds — by a full cell process. There is a wide variation in the results shown in the table

TABLE XVIII — CREOSOTE LOSSES, BASED ON RESIDUE INCREASE, FROM SOUTHERN PINE POLES

9 years exposure; Washington-Norfolk line, American Telephone and Telegraph Company—Data of von Schrenk et al.

Pole No.	Average loss, per cent	
	Top	Butt
1425	43.3	2.7
29	42.4	16.4
10749	50.9	20.8
2931	58.0	32.8
9700	70.8	60.3
Overall average	53.1	26.6

TABLE XIX — CREOSOTE LOSSES, BY EXTRACTION

Southern pine test posts, aerial sections; 1926 and 1927 series; Gulfport test plot—BTL (Waterman) data.

	Exposure period, months					
	8	17	22	31	32	46
	Average loss, per cent					
1926 Series						
8 lb empty cell	16		31		41	54
12 lb full cell	9		21		25	34
"Light" oil	17		32		41	55
"Mixed" oil	10		22		28	38
1927 Series						
"Light" oil						
8 lb empty cell		27		44		
12 lb full cell		29		36		
"Heavy" oil						
8 lb empty cell		16		36		
12 lb full cell		18		30		

Table XX — Creosote Losses from Southern Pine Posts

Analyses of original creosote and creosote extracted from outer 1 inch below ground line of posts in test 3, 6 and 7 years; Gulfport test plot; 1925 series; 12 lb. full-cell treatments — BTL (Waterman) data.

	Original creosote	After 3 years‡	After 6 years	After 7 years
Specific gravity	1.037 (38/15.5)	1.056 (60/60)	1.057 (60/60)	1.059 (60/60)
Distillation, water free basis, per cent, cumulative				
To 210°C	4.1	0.5	0.3	0.1
210–235	32.0*	4.0	3.0	3.4
235–270	44.8	22.6	22.1	26.3
270–300	57.2	38.3	39.0	43.0
300–315	73.8	45.9	47.2	50.7
315–355	88.9*	72.2	73.9	75.8
Residue	10.7†	26.4	25.2	24.1
Total	99.7	98.6	99.1	99.9
Sulph. residue, gm/100 ml	3.2	4.2	3.7	2.4
Tar acids, gm/100 ml	8.3	3.3	2.8	6.5
Estimated losses, per cent, based on residue increase		59.5	57.5	55.6

* To 360°C in original oil analysis.
† Above 360°C in original oil analysis.
‡ The values for 3, 6 and 7 years are averages of data for 2, 3 and 3 posts, respectively.

Table XXI — Creosote Losses from Southern Pine Posts

Analyses of original creosote and of creosote extracted from outer 1 inch zone below ground line of posts in test 4, 5 and 6 years; Gulfport test plot; 1926 series; 8 lb. empty-cell treatments — BTL (Waterman) data.

	Original creosote	After 4 yrs.	After 5 yrs.	After 6 yrs.
Specific gravity	1.044 (38/15.5)	1.080 (60/60)	1.084 (60/60)	1.131 (60/60)
Distillation, water free basis, per cent, cumulative				
to 210°C	1.3	0.3	0.9	—
210–235	34.2	1.7	2.9	—
235–270	60.1	15.7	21.1	—
270–300	74.1	38.4	39.7	6.6
300–315	79.9	49.2	50.0	15.2
315–355	96.3*	79.7	77.6	54.7
Residue	3.4†	19.5	21.6	44.1
Total	99.7	99.4	99.2	98.8
Sulph. res., gm/100 ml	1.7	0.5	1.1	—
Tar acids, gm/100 ml	8.2	2.4	2.7	3.9
Estimated losses, per cent, based on residue increase		69.2	72.2	86.4

* To 360°C in original oil analysis.
† Above 360°C in original oil analysis.

for the butt sections, but the top and butt figures, respectively, seem to bear some relation to each other.

R. E. Waterman in a Bell Telephone Laboratories' memorandum dated January 23, 1928, reported losses of creosote from poles removed from the Montgomery-New Orleans line and in a memorandum dated March 7, 1931, reported creosote losses, determined by periodic extractions, from the aerial sections of southern pine posts treated in 1926 and 1927. Companion posts are among the earliest lots reported on by Lumsden.[76] Part of Waterman's data are condensed in Table XIX. His figures confirm in general the conclusions reached by Bond,[32] namely, that the losses were greater for the light than for the heavy treatments, for the empty cell than for the full cell treatments, and in addition, for the lighter oil than for the heavier oil. Such conclusions are in line with what might be expected from the physical characteristics of the creosotes and general knowledge of the distribution and dispersion of the creosotes in the various treatments.

The losses shown in Table XIX are rounded figures that apply to the whole cross section of the pole-diameter posts. Of more significance are data on creosote losses from the *outer 1 inch* of the *below ground* section of companion posts in the Gulfport plot. Tables XX and XXI show distillation figures for the original creosotes and for the extracted oils, from full cell and from empty cell posts, after varying exposure periods up to about seven years. The oils were both low residue creosotes. The indicated percent losses are based on the increase in the residue above 355°C — of which more later. The losses are greater for the empty cell treatments than for the full cell treatments. The fact that so much of the loss occurred within the first four years is extremely important in evaluation philosophy.

Tables XXII and XXIII present data for whole cross sections of two posts that had begun to decay and that were removed for assay four years after installation at Gulfport. Table XXII shows the original analysis of the creosote and the average analysis of the extracted oil from the two posts. The indicated loss in the ground line decay area, figured from the residue increase, was *61.1 per cent.* Table XXIII shows the distribution of creosote at treatment by zones — from the outside toward the heartwood line — from extracted borings, and the distribution of creosote after removal from test, based on extraction of sectors cut from whole cross section disks. The indicated losses, figured from average over-all retention at treatment and after removal were 65.1 and 55.5 per cent, or an average of *60.3 per cent.* This figure can be considered to be in agreement with the 61.1 per cent figure cited above.

TABLE XXII — CREOSOTE LOSSES FROM SOUTHERN PINE POSTS

Analyses of original creosote and of creosote extracted from posts after 4 years in test; 1936 series; Gulfport test plot — BTL (Waterman) data.

	Original creosote	Average, extracted creosote*
Specific gravity	1.055 (38/15.5°C)	1.134 (60/60)
Distillation, water free, per cent, cumulative		
to 210°C	3.4	0.2
210–235	17.0	0.5
235–270	40.4	2.1
270–300	53.3	7.5
300–315	59.6	15.0
315–355	81.2	51.5
Residue	18.3	47.0
Total	99.5	98.5
Sulphonation, residue, gm/100 ml	4.7	3.7
Tar acids, gm/100 ml	7.9	12.5

* Average analyses of toluene extracted oils from disks cut adjacent to decay line; posts 273 and 280. The estimated average loss, based on residue increase, is 61.1 per cent.

TABLE XXIII — CREOSOTE LOSSES FROM SOUTHERN PINE TEST POSTS

By zones, by toluene extraction; 1936 series; Gulfport test plot — BTL (Waterman) data.

Post No.	Years in Test	Retention, by extraction lb/cu ft					
		Whole cross section	Zones, inches				Remainder of treated sapwood
			Outer ¼	Next ¼	Next ½	Next 1	
Original retentions, at treatment							
273	—	3.8*	6.9	3.5	3.8	3.9	3.0
280	—	4.7*	11.2	4.1	3.3	1.5	3.5
Retention after removal							
273	4	1.69†	1.51	1.20	1.34	1.76	2.50
280	4	1.64‡§	2.78	1.32	1.90	1.63	1.10

* Average analyses of boring samples.

† Average analyses of sectors cut from a disk taken 3 inches above maximum decay line.

‡ Average analyses of sectors cut from two disks taken 3 inches above and 3 inches below the maximum decay line. (See Table XXII for analyses of original and extracted creosote.)

§ Average loss, estimated from retentions at treatment and after removal:

For whole cross section .... 60.3 per cent

For outer 1 inch

Post 273 .... 70.3 per cent

Post 280 .... 64.0 per cent

Average .... 67.2 per cent

Calculated losses in the outer 1 inch of these two posts — assuming 8-inch diameter — are 64.0 and 70.3 per cent, respectively, or an average of 67.2 per cent; and this figure corresponds very closely with the four-year loss figure of 69.2 per cent calculated for the empty cell posts in Table XXI. The posts represented in Tables XXII and XXIII were obviously treated to retentions that were too low to be effective; but they illustrate what is likely to happen when too low retentions of highly volatile light creosotes are used in wood in contact with the ground.

Bateman[7] discusses a laboratory experiment to determine creosote losses, over a 70-day period from pieces of round post sections, 5 inches in diameter and 2 feet long; and he extends his comparison to data derived from experiments conducted for the San Francisco Marine Piling Commission. He reports the following treatment data for the three creosotes involved:

1. 18 lb/cu ft of a creosote with 92 per cent distilling below 275°C;
2. 10 lb/cu ft of a creosote with 40 per cent distilling below 275°C; and
3. 27.5 lb/cu ft of a creosote with 42 per cent distilling below 275°C.

His comparative loss data are condensed in Table XXIV.

Waterman and Williams[125] report creosote losses based on periodic extractions of comparable lots of specimens from treated round southern pine *saplings* exposed in the Gulfport test plot. Their data are condensed

TABLE XXIV — CREOSOTE LOSSES

By weight, from round southern pine post sections*, and from pile sections—Madison (Bateman) data.

Exposure period, days	Group†		
	18 lb/cu ft	10 lb/cu ft	27.5 lb/cu ft
	Average loss, per cent		
10	16.0	7.0	1.3
20	22.5	9.5	—
30	28.0	12.0	3.7
40	32.0	13.5	—
50	36.0	15.5	—
60	39.0	17.0	—
70	42.0	18.5	—
90	—	—	7.3
222	—	—	16.6
475	—	—	24.5
510	—	—	25.3
785	—	—	38.2

* Five inch diameter posts, 2 feet long.
† See text for group and creosote data.

TABLE XXV — CREOSOTE LOSSES FROM ROUND SOUTHERN PINE SAPLINGS

By extraction; Gulfport test plot—BTL (Waterman and Williams) data.

	Years in test					
	1		2		3	
	n	Av. loss, per cent	n	Av. loss, per cent	n	Av. loss, per cent
Empty-cell treatments; 3.5–15.1 lb/cu ft						
Above ground line	14	35.5	10	55.8	3	55.9
Below ground line	14	40.4	10	58.8	3	61.6
Full-cell treatments; 14.0–38 lb/cu ft						
Above ground line	9	22.7	7	44.2	4	59.8
Below ground line	9	13.9	7	39.2	4	42.4

TABLE XXVI — ANALYSES OF CREOSOTES USED IN WEATHERING WHEEL AND OUTDOOR EXPOSURE TESTS

Southern pine sapwood blocks—Koppers (Rhodes et al.) data.

	Creosote I	Creosote II
Specific gravity at 38/15.5°C	1.064	1.081
Distillation, per cent, cumulative		
0–210°C	3.8	1.8
210–235	21.2	13.7
235–270	41.4	30.2
270–315	59.8	46.3
315–355	80.9	64.9
Residue above 355°C	18.7	34.9
Water	0.6	0.6
Specific gravity of fractions		
235–315°	1.039	1.037
315–355°	1.106	1.104

TABLE XXVII — CREOSOTE LOSSES FROM SOUTHERN PINE SAPWOOD BLOCKS

Ether extraction; weathering wheel tests—Koppers (Rhodes et al.) data.

Exposure period, weeks	Average loss,* per cent	
	Creosote I	Creosote II
0	0.0	0.0
1	33.7	28.4
3	50.1	40.3
5	57.2	45.3
9	64.7	52.3

* Treatment retention 16.7 lb/cu ft.

in Table XXV. The loss figures are averages for all the creosotes used. It will be noted that there is a large variation in the treatment retention groups and that the losses were more rapid and definitely higher for the lower retention empty-cell specimens. Preliminary estimates from data on creosoted ¾-inch square stakes at Gulfport indicate losses of about the same order of magnitude, with the trend in the direction of relatively higher figures than those for the round saplings. This is in line with expectation because of the use of the toluene diluent and the practice of controlling the treatments to secure lower than threshold retentions in both full-cell and empty-cell treatments.

Creosote Losses from Treated Blocks

E. O. Rhodes and his colleagues have published two excellent papers[50, 89] that are most significant in a discussion of creosote losses from treated wood blocks. They used two creosotes, I and II, the analyses of which are shown in Table XXVI. Southern pine sapwood blocks measuring 0.5 x 0.5 x 3.0 inches, with the long axis in the direction of the grain, were treated and exposed on a weathering wheel in the laboratory, and also out of doors. The laboratory test specimens were treated to a retention of 16.7 pounds per cubic foot by soaking the blocks, heated to 105°C, in creosote at 100°C, in a sealed container. The creosote cooled during a five-hour soaking period to 40–50°C. The authors felt that the retention of about 16.5 pounds would facilitate extraction recovery of enough oil for analysis, and that "the treated portion of a tie or pole probably contains about this amount of oil." Bell Telephone Laboratories' experience has shown (Fig. 29) that the retention in the commercially treated 8-pound post averages only about 12 pounds per cubic foot in the outer ¼ inch of wood and that the retention drops off rapidly in the wood farther beneath the surface. The Rhodes blocks, therefore, must be regarded as heavily treated.

Losses of creosote were determined by ether extraction, and the change in character of the preservative was determined by distillation of the extracted oil. Corrections were made for resin extracted with the creosote. The losses of creosote from the test blocks on the weathering wheel are shown in condensed form in Table XXVII.

The authors report creosote loss from blocks treated to a 15.0-pound retention and exposed outdoors during the winter as 44.4 per cent; and similar blocks treated to a 16.5-pound retention lost 47.1 per cent in a nine-week exposure period during the summer. The results of the experiments were taken to mean that the losses were of about the same character in the outdoor winter and summer exposure tests, and that the blocks weathered on the wheel in the same way they did outdoors. Rhodes reaffirms his conclusions about the weathering wheel tests in 1936 in a discussion of the Snell and Shipley paper[108] in these words:

"Our consideration of this problem convinced us, and Snell and Shipley agree with the opinion, that natural weathering produced by heat and cold, rain and wind, involves not only evaporation but water leaching and mechanical losses of whole oil by water or by bleeding. . . . To simulate these conditions, we exposed blocks of wood treated with the test creosotes to variations in temperature, to moving water and to moving air. . . . In fact, we believe that our wood-block exposure tests include most, if not all, of the factors of natural weathering. . . ."

Now to go back a bit to Curtin's 1926[32] experiments, this time referring to his own weathering tests on creosoted wood. Table XXVIII is an interpretation of the results he obtained by exposing small blocks, cut from pressure treated 2 x 4 inch southern pine sapwood stakes, to natural out-of-door weathering. His procedure was extremely severe, but he was aiming at an extreme accelerated test for permanency. After treatment his 2 x 4 inch sample pieces were held in storage under cover for 2 months; and then they were cut up so that the test blocks were about 2 x ⅛ x ⅜ inches. These small pieces were exposed 15 feet above the ground on wooden trays for a four-month period from September, 1926, to January, 1927, and for a ten-month period from September, 1926, to July, 1927. Losses from the 2 x 4 pieces that must have occurred during their two-month storage period would increase the loss figures shown in the table. The losses are obviously greater for the lighter treatments, and for the lower residue oil.

Table XXVIII — Creosote Losses from Southern Pine Blocks*

Outdoor weathering tests—Based on data by Curtin.[32]

Creosote†	Trt. No.	Average retention lb/cu ft			Average loss, per cent	
		At treatment	After 4 mos.	After 10 mos.	After 4 mos.	After 10 mos.
1	1	23.18	19.69	17.13	15.1	26.1
1	2	17.13	11.69	10.88	31.8	36.5
1	3	19.38	6.50	5.44	37.4	47.6
2	1	26.19	18.75	15.00	28.4	42.7

* See text for description of blocks

† Creosote number 1 was a domestic oil, specific gravity 1.056 and residue above 355°C of about 26 per cent; number 2 was a British oil, specific gravity, 1.068 and residue above 355°C of about 19 per cent.

In their roof exposure, outdoor weathering tests Duncan and Richards[35, 39] have regularly found losses of creosote, by weight, in ¾-inch blocks treated with creosotes having residues above 355°C of 20–30 per cent, to run in the neighborhood of 45–50 per cent. All treatments were made by a full-cell vacuum process with toluene creosote solutions. The over-all exposure period consisted of three stages,[39] a three-week conditioning period in a constant humidity and temperature room at 30 per cent relative humidity and 80°F, a sixty-day outdoor exposure on a rack on the roof, and a three-week reconditioning to approximately constant weight in the same humidity room. The losses were figured from the original conditioned weight of the blocks, the creosote retention at treatment, and the calculated amount of creosote remaining as indicated by the weight of the weathered and reconditioned blocks. A condensed and simplified summary for two creosotes and two block shapes[39, 41] is shown in Table XXIX.

As might be expected under identical weathering conditions, the losses were higher in the case of the ⅜ x ¾ x 1½ inch blocks — with twice as much end grain exposed as the ¾″ drilled cubes — which presumably resulted in an accelerated longitudinal evaporation. In other words, a given percentage loss of creosote is arrived at sooner in the case of the block with greater transverse surface area. The loss increase amounts to about 7 per cent at the 8 to 10-pound retention level. This loss is about twice as great as it would be if calculated on the basis of the increased

Table XXIX — Creosote Losses from Longleaf Southern Pine Sapwood Blocks

Outdoor weathering tests—Madison (Duncan and Richards) data.

Treatment retention lb/cu ft	Creosote number		
	Coop. No. 2	5340	5340
	Creosote loss, per cent, by weight		
18*	44.0†	—†	—‡
16	45.0	—	—
14	45.9	—	—
12	47.0	—	—
10	48.0	42.5	50.0
8	49.5	44.3	51.2
6	51.3	46.5	54.6
4	54.3	49.8	60.0
2	60.0	54.5	75.0

* Retention by full-cell treatment under vacuum with toluene-creosote solutions.
† ¾-inch cubes.
‡ ⅜ x ¾ x 1-½ inches.

surface area only of the flat blocks. The losses are interpreted in general to mean that the exposure conditions were not as severe as those in the Rhodes[89] weathering wheel tests, and in line with his outdoor tests.

The results of recent determination at Bell Telephone Laboratories of creosote losses from ¾-inch cube blocks, by toluene extraction, are shown in Table XXX. The blocks were divided into thirds before extraction (Fig. 30) and the respective parts were further divided and then pooled in the extractor.

The blocks in each group of twenty were selected to represent the whole gradient of treatments with an average treatment retention for each group of 6.05 pounds per cubic foot. The average density of the blocks, oven-dry weight and volume basis, was the same for each group, namely 0.56. The oil used was cooperative creosote No. 11, a 50/50 blend of British vertical retort tar creosote and British coke oven tar creosote,[5, 12, 39] diluted as usual with toluene. The exposure period outdoors was sixty days at the Chester Field Station between January 4 and March 3, 1952, plus several days exposure on a bench in a steam heated laboratory both before and after the outdoor period. The average loss of 40.3 per cent is considered to be in line with losses for the same creosote in the Madison tests, considering the factor of the winter climate at Chester.

The distribution of the residual creosote in the test blocks as shown by the averages reported may be considered to be remarkably uniform. The differences in these averages are not regarded as statistically significant. Further discussion of losses from weathered blocks will be resumed in later paragraphs.

TABLE XXX — CREOSOTE LOSSES FROM LOBLOLLY-SHORTLEAF SOUTHERN PINE SAPWOOD BLOCKS, AND LB/CU FT REMAINING

By toluene extraction; outdoor winter weathering tests — BTL (Snoke and MacAllister) data.

Lot No.	n	Average retention at treatment, lb/cu ft	lb/cu ft remaining			
			Outer* third	Middle third	Outer third	Whole block
1	20	6.05	3.62	3.43	3.93	3.66
1	20	6.05	3.49	3.49	3.68	3.55

Average lb/cu ft of creosote remaining;
in outer thirds 3.68
in middle thirds 3.56
in whole blocks 3.61
Average loss of creosote; 2.44 lb/cu ft, or 40.3 per cent.

* ¾-inch cubes. See cutting diagram in Fig. 30.

Creosote Losses from Impregnated Filter Paper

There have been numerous criticisms of the use of creosote loss figures obtained by evaporation from open dishes in any consideration of creosote permanence.[108] The use of the losses reported in this section may be criticized in a similar manner, but they represent extreme acceleration and they seem to have a bearing on the interpretation of any weathering

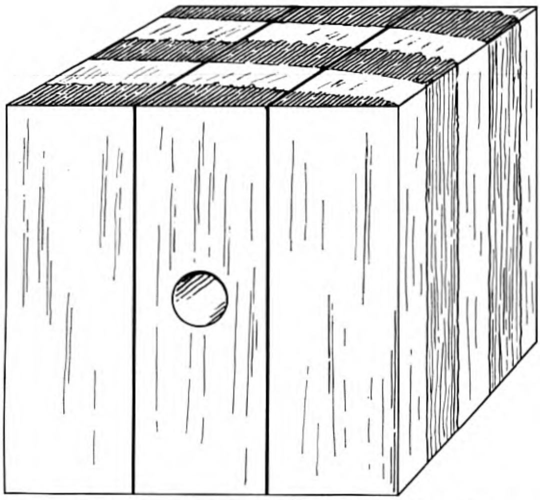


Fig. 30—Diagram of cutting plan for dividing a weathered creosoted block into three approximately equal parts for determination of residual creosote by toluene extraction.

tests in which evaporation plays the major part. Some years ago the late Heinrich T. Boving of Bell Telephone Laboratories, ran an extensive series of evaporation experiments on eight so-called Fulweiler creosotes. He used impregnated strips of filter paper, hung on quartz springs in protecting glass apparatus and exposed to a constant flow of air with no turbulence, and under constant temperature and humidity. His experiments were performed with great care. Repetitions gave excellent agreement. A condensation of his loss figures, rounded off to whole numbers, for seven- and fourteen-day exposure periods, are shown in Table XXXI. Let it be stated unequivocally at this point that it is recognized that there are important physical differences in the wood fiber combinations represented by filter paper, wood blocks, small round saplings, posts and poles; but the *quantitative* loss data seem to fall sooner or later into a similar pattern for all these test media, under the various conditions, and with the assumptions made. About the difference in the *qualitative* changes that take place during creosote losses there is much, but not enough, information.

An Interpretation of Creosote Losses

Frosch,[49] in describing certain physical characteristics of the Fulweiler oils, states that they may be considered as truly viscous solutions in

TABLE XXXI — CREOSOTE LOSSES, BY WEIGHT, FROM IMPREGNATED FILTER PAPER, AND CALCULATED INCREASE IN RESIDUE

BTL (Boving) data

Fulweiler creosote No.	Original residue above 335° C, per cent	Per cent loss		Calculated residue*	
		in 7 days	in 14 days	in 7 days	in 14 days
1	43.84	31	—	63.5	—
2	40.25	32	36	59.2	62.9
3	31.11	36	—	48.6	—
4	28.52	38	—	46.0	—
5	20.85	44	—	37.2	—
6	15.49	48	—	29.8	—
7	12.21	54	—	26.5	—
8	8.49	61	68	21.8	26.5

* Assuming that all loss occurs in fraction below 355°C.

TABLE XXXII — THEORETICAL CHANGES IN CREOSOTE

Loss of volatile fractions by evaporation; amount remaining of total fraction below 355°C.

1	2	Time period 1*					Time period 2				
		3	4	5	6	7	8	9	10	11	12
Fulweiler creosote No.	Assumed treatment retention lb/cu ft	Per cent loss	Residual oil lb/cu ft	Calculated per cent residue above 355°C in residual oil	More than 355° lb/cu ft	Less than 355° lb/cu ft	Per cent loss	Residual oil lb/cu ft	Calculated per cent residue above 355°C in residual oil	More than 355° lb/cu ft	Less than 355° lb/cu ft
2	8.00	32	5.44	59.2	3.22	2.22	36	5.12	62.9	3.22	1.90
8	8.00	61	3.12	21.8	0.68	2.44	68	2.56	26.5	0.68	1.88
2	10.00	32	6.80	59.2	4.03	2.77	36	6.40	62.9	4.03	2.37
8	10.00	61	3.90	21.8	0.85	3.05	68	3.20	26.5	0.85	2.35
2	12.00	32	8.16	59.2	4.83	3.33	36	7.68	62.9	4.83	2.85
8	12.00	61	4.68	21.8	1.02	3.66	68	3.84	26.5	1.02	2.82

* Time periods 1 and 2 represent exposures that would result in the per cent losses determined from Boving's 7 and 14 day tests, respectively. See text.

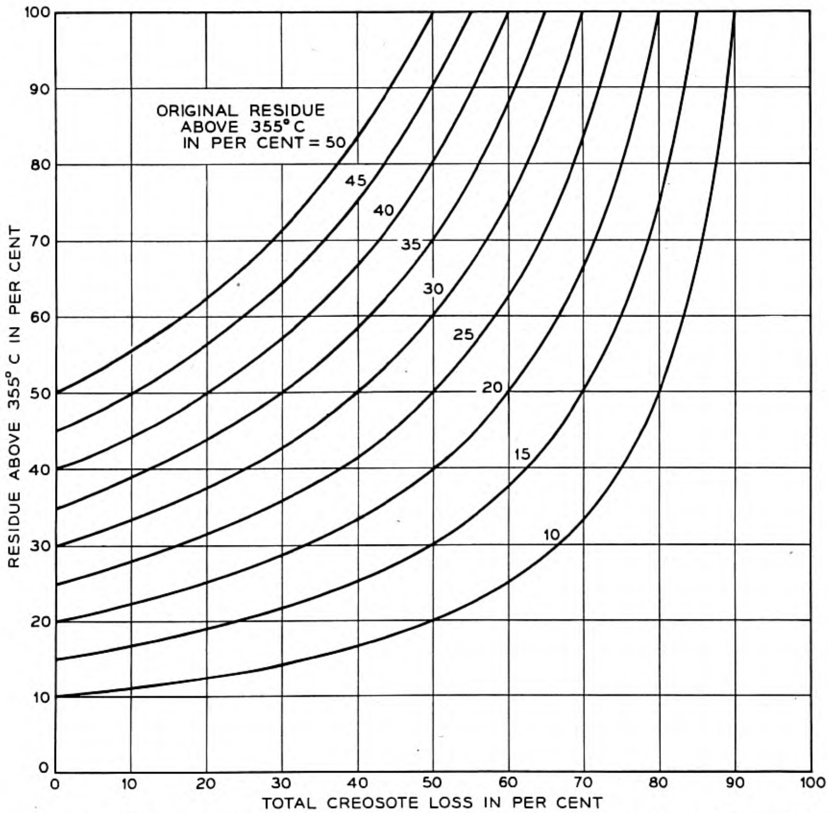


Fig. 31—Theoretical relation of the per cent increases in the residue above 355°C and the per cent losses of total creosote.

which the fraction below 355°C is the solvent and the fraction above 355°C is the solute; and that this condition does not hold for any other temperature point.

Let it be assumed that one may dodge the inferences of the results obtained by Hudson and Baechler[59] about the increase that may occur in the residue above 355°C as a result of oxidation. Here one would be in agreement with Schmitz et al.[102 p. 270] Let it be assumed further that all loss takes place in the fraction below 355°C, and that the residue above 355°C is inert, (cf. Loseby and Krogh[73]). The residue in the eight creosotes used by Boving would increase in their appropriate ratios to the figures shown for seven and fourteen days in Table XXXI. Losses for creosotes of different initial residues and consequent residue increases can then be represented by a family of curves for residue increase with quantitative loss such as those shown in Fig. 31. Now let it be assumed that these hypotheses can be applied to treated wood, with full realization that the application may err in the direction of oversimplification. Bateman and Cislak accept the general principles of the 355°C division point between volatile and nonvolatile creosote constituents in debating the theoretical aspects of creosote losses.[108] Rhodes[89] indicates some actual loss in the fraction above 355°C in his test blocks, possibly in part the result of oil displacement in the water phase of his tests, rather than any increase that might occur as a result of oxidation.[59] For the purpose at hand in this paper it is convenient to use the relations shown in the percent loss — residue increase curves in Fig. 31.

Table XXXII represents a theoretical approach to what would happen to the gross characteristics of the oils if, in some time period X, the losses from creosoted wood treated to 8-, 10- and 12-pound retentions became

Table XXXIII — Creosote Losses* from Southern Pine Sapwood Blocks†

Calculations of residual fractions below 355°C; weathering wheel tests.‡

1	2	3	4	5	6	7	8
		Creosote remaining		By extraction		By calculation§	
Exposure period, weeks	Creosote loss, per cent	per cent	lb/cu ft	Residue above 355°C, per cent	Residual fraction below 355°C lb/cu ft	Residue above 355°C, per cent	Residual fraction below 355°C lb/cu ft
Creosote I							
0	0.0	100.0	16.70‖	19.0	13.53	19.0	13.53
1	33.7	66.3	11.07	25.2	8.28	28.7	7.89
3	50.1	49.9	8.33	34.6	5.45	38.1	5.16
5	57.2	42.8	7.15	40.2	4.28	44.4	3.97
9	64.7	35.3	5.90	48.7	3.03	53.8	2.73
12‖	66.5	33.5	5.59	56.3	2.44	56.7	2.42
Creosote II							
0	0.0	100.0	16.70‖	33.6	11.09	33.6	11.09
1	28.4	71.6	11.96	43.7	6.73	46.9	6.34
3	40.3	59.7	9.97	51.4	4.85	56.3	4.36
5	45.3	54.7	9.14	57.3	3.90	61.4	3.53
9	52.3	47.7	7.97	60.8	3.12	70.4	2.36
12¶	55.0	45.0	7.50	65.0	2.53	74.7	1.89

* Losses based on ether extraction.
† Blocks ½ x ½ x 3 inches.
‡ See Bibliography, References 50 and 89.
§ Assuming that all loss occurs in the fraction below 355°C.
‖ Retention by soaking in undiluted creosote.
The 12-week figures were calculated from extrapolations of the loss curves.

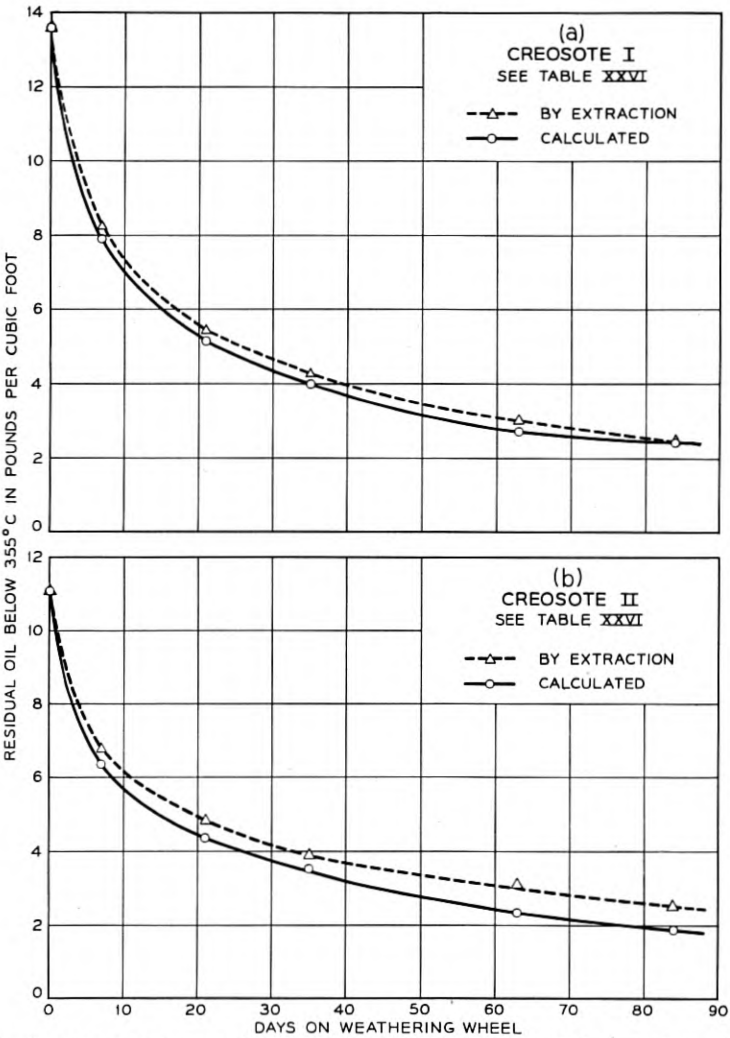


Fig. 32—Creosote losses; lb/cu ft of the fraction below 355°C remaining in the test blocks, by calculation and by ether extraction, based on weathering wheel test data. See text. (a) Creosote I, 19.0 per cent residue above 355°C. (b) Creosote II, 33.6 per cent residue above 355°C.

as great as those determined by the Boving evaporation experiments. The columns have been numbered to facilitate reference. Attention is directed to Columns 6 and 7, and to columns 11 and 12. It will be noted that at the end of Time Period 1 the calculated amounts of material boiling below 355°C in creosotes 2 and 8 are approaching the same magnitude in the respective 8-, 10- and 12-pound treatment groups; and by the end of Time Period 2 (columns 11 and 12), the calculated amounts in each respective treatment group are practically the same for these two creosotes that originally had residues above 355°C of 40.25 and 8.49 respectively. Actually, of course, the time periods would not be the same for the respective 8-, 10- and 12-pound treatments, but — on the basis of loss data already presented — would probably be relatively shorter for the lower retention group and relatively longer for the higher retention group.

Table XXXIII illustrates two methods of arriving at estimates of the per cent and amount of creosote remaining in the respective fractions above and below 355°C, by extraction data and by calculation of the

Table XXXIV — Calculated Amount of Creosote Fraction below 355°C Remaining after Various Exposure Periods under Weathering Wheel Test Conditions

Exposure period, weeks	Treatment retentions—lb/cu ft		
	16.7	10.0	8.0
	Residual fraction below 355°C lb/cu ft		
Creosote I			
0	13.53	8.10	6.48
1	7.89	4.96	3.46
3	5.16	3.44	2.16
5	3.97	2.57	1.59
9	2.73	1.82	.99
12	2.42*	1.46	.84
Creosote II			
0	11.09	6.64	6.11
1	6.34	4.03	2.72
3	4.36	2.91	1.85
5	3.53	2.35	1.37
9	2.36	1.87	.81
12	1.89	1.58	.59

* See text for significance of horizontal lines, page 485.

theoretical residue increase related to the given loss data. The loss data are those of Rhodes et al (loc. cit.) (see Table XXVII) with extrapolation to a twelve-week period; column 2. The same holds for the residue data in column 5, from which the residual pounds per cubic foot below 355°C of Column 6 were calculated, by applying the respective percent residues to the pounds per cubic foot of total creosote remaining in the blocks, column 4. The figures in column 8 were calculated in the same way by using the theoretical percent residue figures in column 7. The data in columns 7 and 8 are represented by the curves in Fig. 32, for Creosotes I and II. The curves approach each other closely enough for the purposes of the present interpretation, particularly in the case of the 19 per cent residue oil.

Now, supposing that Rhodes et al had used two lower retentions, say 10 pounds and 8 pounds in their experiments in addition to the 16.7 pounds; and assuming that their losses would be of about the same magnitudes for the 16.7 and 10-pound treatments and slightly higher for the 8-pound treatment, one would come out with theoretical calculations like those shown in Table XXXIV. The significance of the magnitudes of these amounts below 355°C will be suggested in the following paragraphs.

The Gross Characteristics of the Residual Creosotes in Soil-Block Tests of Weathered Blocks

Madison test data for average losses of creosotes by weight, from treated blocks, during the periods of weathering and reconditioning are rounded off and plotted in Fig. 33. The usual procedure was to bring the weathered blocks into the laboratory and place them under constant temperature and humidity of 80°F and 30 per cent until the blocks came to approximate weight equilibrium. Examination of the figure indicates very definitely (1) heavier losses in the lower retentions respectively for all of the creosotes, and (2) lower losses for the higher residue oils than for the lower residue oils. These conclusions follow naturally from a consideration of the distillation ranges of the oils and the treatment conditions, confirming the inferences from other loss data already discussed.

These losses of creosote from 8- to 10-pound groups of weathered treated wood blocks seem to be in the neighborhood of 40–50 per cent by weight or by extraction. The aim of the weathering techniques now being developed at Bell Telephone Laboratories is to reach some definable end point in the protective life of the preservative that will reflect the end point or failure point of test plot specimens or poles in line.

Other things being equal, one can then consider and plan for treatment retentions sufficiently high to assure a protective residual of a given preservative for a long, economical service life of the treated plant units. This service life is not necessarily an indefinitely long life, nor is it the maximum physical life that might be obtained by using relatively larger quantities of preservatives than might be consistent with cleanliness requirements. Implicit is the idea of avoiding at all times heavy maintenance and replacement costs on account of decay, but particularly in the early life of the line. To accomplish these ends it is necessary to know how much preservative to use.

The following interpretation of the results of soil-block tests is confined to weathered block experiments, for the practical reasons previously cited. The interpretation is not offered as something entirely new; but it looks like a good working hypothesis, and it seems to help in explaining what may be happening in laboratory soil-block tests of creosote.

Table XXXV — Summary and Interpretation of Soil-Block Tests

Weathered, creosoted southern pine sapwood blocks; creosote losses; amounts and gross characteristics of residual oils at threshold retentions for *Lentinus lepideus*.

1	2	3	4	5	6	7	8	9	10	11
Item	Creosote No.*	Specific gravity 38/15.5°C	Residue above 355°C per cent	Threshold lb/cu ft	Per cent loss	Creosote loss lb/cu ft	Residual creosote lb/cu ft	Calculated residue above 355°C	Residual creosote	
									>355°C lb/cu ft	<355°C lb/cu ft
1	1	1.065	18.5	9.8	53.1	5.2	4.6	39.4	1.81	2.79
2	7	1.077	20.5	9.0	47.8	4.3	4.7	39.3	1.85	2.85
3	2	1.081	30.6	10.2	47.1	4.8	5.4	57.8	3.12	2.28
4	6	1.093	34.2	9.0	37.8	3.4	5.6	55.0	3.08	2.52
5	3	1.108	50.4	12.2	30.3	3.7	8.5	72.3	6.15	2.35
6	8	1.115	53.2	9.4	25.5	2.4	7.0	71.4	5.00	2.00
7	9a		21.2	5.7	40.4	2.3	3.4	35.6	1.21	2.19
8	9	1.001	20.0	5.8	43.1	2.5	3.3	35.1	1.16	2.14
9	10a		14.4	6.7	50.9	3.4	3.3	29.4	0.97	2.23
10	10	1.068	15.2	6.9	52.2	3.6	3.3	31.8	1.05	2.25
11	11	1.038	18.0	6.5	47.7	3.1	3.4	34.4	1.17	2.23
12	M1	1.107	41.9	8.0	33.8	2.7	5.3	63.3	3.35	1.95
13	M2	1.070	18.1	8.3	50.6	4.2	4.1	36.6	1.51	2.59
14	BTL 5340	1.088	20.9	7.5	46.6	3.5	4.0	39.1	1.57	2.43

* Creosotes 1, 2, 3, 6, 7, 8, 9, 10 and 11 are those in use in the Cooperative Creosote Tests (see Bibliography, References 12 and 39). Oils 9a and 10a are samples from the same lots as numbers 9 and 10. (See Bibliography, Reference 36.) For oils M1 and M2 see Bibliography, References 37 and 38. Creosote 5340 is shown in Table II.

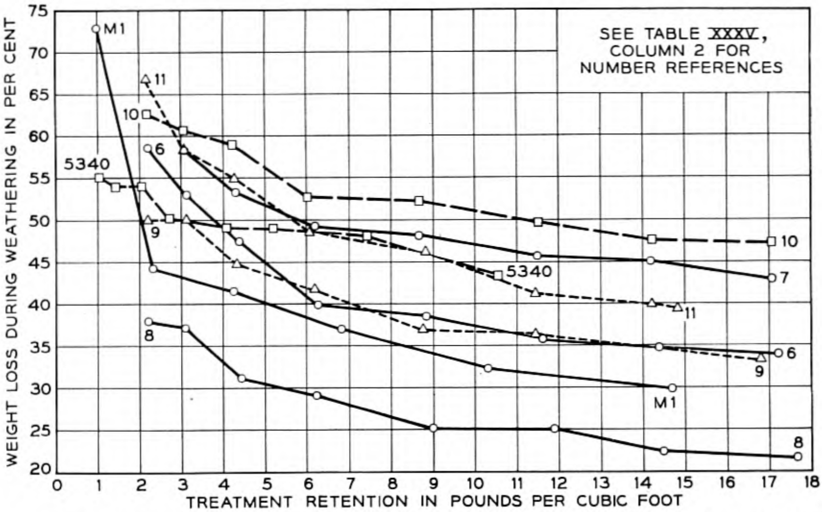


Fig. 33—Creosote losses in weathered, ¾-inch cube, southern pine sapwood test blocks; relation of per cent loss of total creosote to original residue above 355°C and retention at treatment, lb/cu ft. All treatments were made with toluene creosote solutions. The total elapsed time from treatment to final reconditioned weight was about 105 days, including 60 days outdoor exposure on the Laboratory roof at Madison, Wis. See Table XXXV for number references.

Table XXXV is a condensed set of data on fourteen lots of weathered, creosoted ¾-inch cube blocks. All of the tests were run at the Forest Products Laboratory at Madison, Wisconsin, in the Division of Forest Pathology, under the direct supervision of the same investigator, Dr. Catherine G. Duncan.[36, 37, 38, 39, 41] The test fungus was *Lentinus lepideus*, Mad. 534. Ten of the tests have been run in cooperation with Bell Telephone Laboratories, and four have been run more or less concurrently with other cooperators. The technique for handling the weathered blocks has been essentially the same, and it has been rigidly controlled, except for the vagaries of the weather itself, at all essential points.

The data for the creosotes (Cols. 3 and 4), for the thresholds (Col. 5), and for the amount of residual oil in the blocks at the time they were placed in test (Col. 8), and the per cent and amount lost (Cols. 6 and 7) are all taken from the published reports or from manuscripts either ready[36] or in preparation for publication.[41] The writer has calculated the residues in the residual oils, and the respective amounts remaining above and below 355°C (Cols. 9, 10 and 11) in pounds per cubic foot, on the assumption that all the loss occurred in the fractions boiling below 355°C. Particular attention is directed to the figures in Col. 11 — the calculated amounts remaining of the fractions boiling below 355°C. In terms of

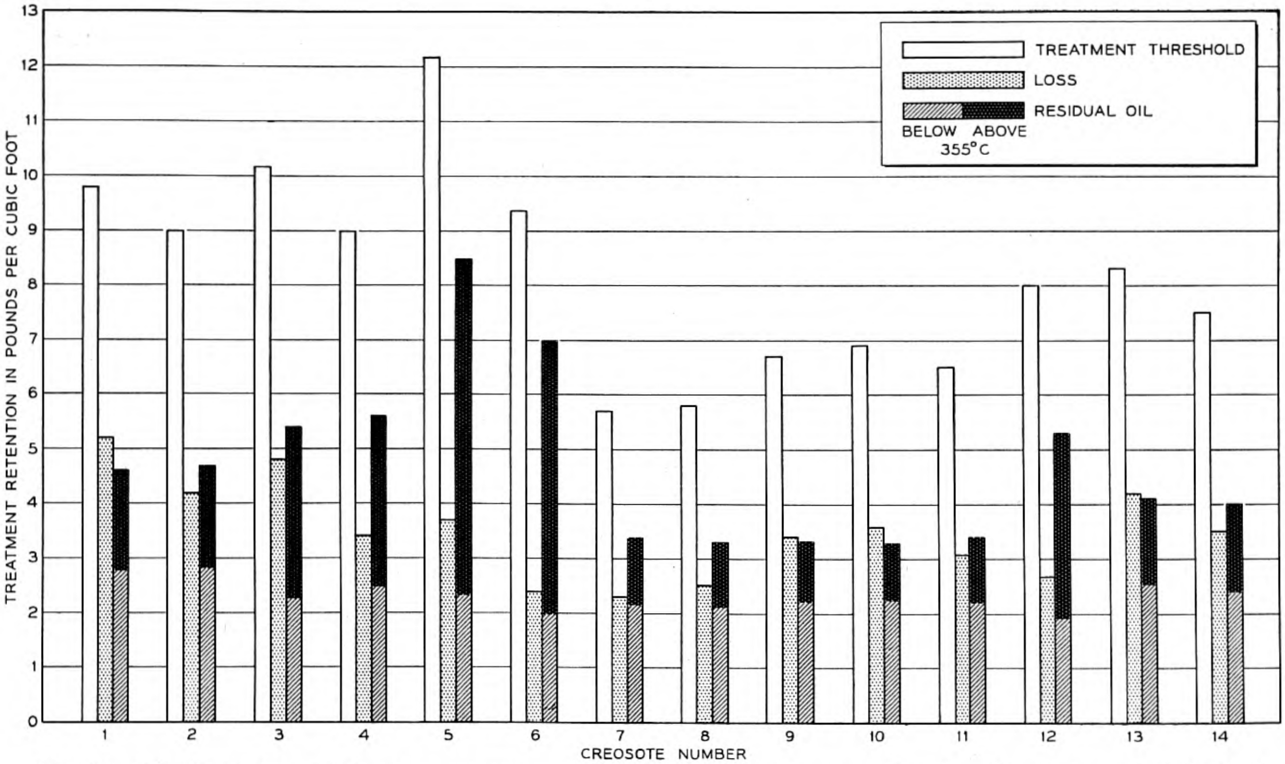


Fig. 34—Soil-block tests of fourteen creosotes against *Lentinus lepideus*; weathered southern pine sapwood blocks; comparison of treatment thresholds, losses during weathering and reconditioning, and gross characteristics of residual creosotes in terms of calculated amounts remaining of fractions above and below 355°C; based on Madison data. See Table XXXV.

pounds per cubic foot the over-all picture of the relative threshold amounts of creosote at the time of treatment, the oil lost, and the calculated proportional parts of the residual oil — after weathering — above and below 355°C, are shown graphically in Fig. 34. All of the data are in terms of pounds per cubic foot. If one bears in mind that the thresholds (Col. 5, Table XXXV) as given by the Madison investigators were located by a combination of visual observation of the blocks and extrapolation of straight lines through the weight loss data one may conclude that *in all of these tests the results were essentially the same for all the*

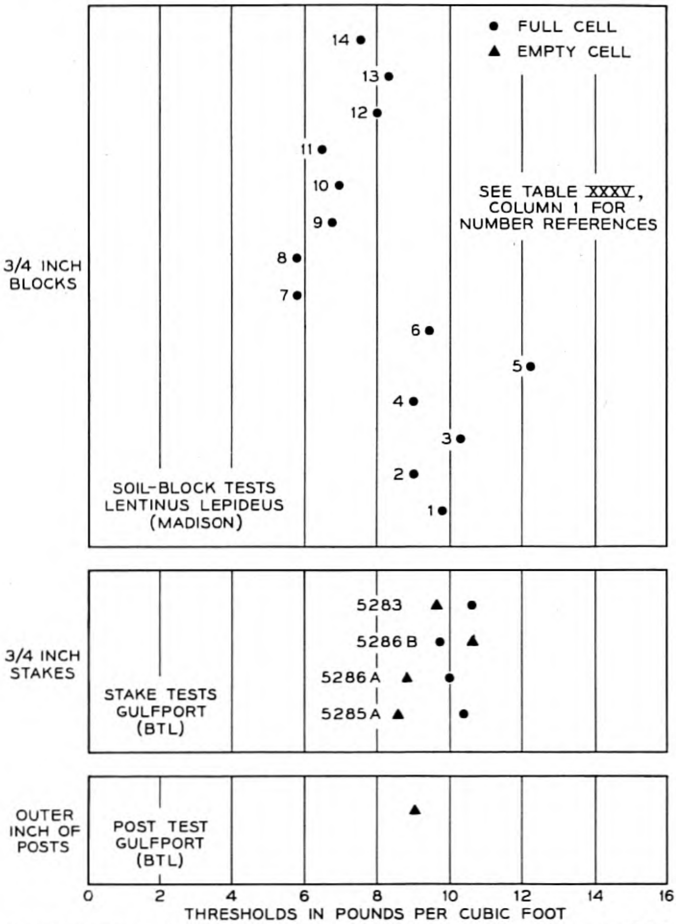


Fig. 35—Relative values of creosote thresholds by soil-block tests, ¾-inch stake tests, and post tests.

creosotes with respect to the indicated threshold amounts distilling below 355°C. Statistically, the figures in Table XXXV Col. 11, are not significantly different. Slight changes in estimating the thresholds would conceivably bring them all to approximately the same level. In any given set of experiments the level would also vary with the duration (and types) of a different weathering cycle.

It will be recognized that no attempt has been made to separate or define the gross components of the fraction remaining below 355°C. This calls for more study, and for the development of refined methods of extraction and assay by weight and by distillation. Also, no attempt has been made to interpret the value or significance of the residue above 355°C, either because of its possible retardation of the evaporation of lower fractions, or because of some potential mechanical blocking effect. The whole interpretation is based on the simple division of the creosotes into two parts, the part distilling below 355°C and the part distilling above 355°C. Refinement will depend upon better future experimental evidence.

As an illustration of the application of the hypotheses discussed in the preceding paragraphs, one may reexamine the data from the weathering wheel experiments.[89] Rhodes, Roche and Gillander used one creosote retention only in their weathering wheel experiments, namely 16.7 pounds per cubic foot. In commenting on their work C. S. Reeve[87 p. 78–79] noted Rhodes' emphasis on "the fact that toxicity without permanence is just as worthless as permanence without toxicity". Reeve and his colleagues carried out somewhat similar weathering experiments using slabs of wood about ⅛ inch thick, exposing the treated pieces to somewhat lower temperatures than those in the Rhodes' experiments and "following a procedure with circulated air, and heat, and water". The plan of the experiments called for the conduct of "weathering cycles with reduced increments of various oils in order to get down finally to a percent of impregnation right at the end of a weathering cycle which would actually yield a rotting specimen of *Lentinus lepideus*". The work had not progressed far enough to accomplish this end, but Reeve says

"The results . . . are in very close corroboration of what Mr. Rhodes has found. In other words, our loss curves with different oils running from relatively low residues to relatively high residues, have been almost parallel, I believe, with the loss curves which he has shown . . ."

Rhodes et al used essentially an agar-block method for testing their weathered blocks against *Lentinus lepideus*, Mad. 534, the same strain as that used in the Madison tests. The residual creosotes in the blocks treated with Oil I and Oil II (Table XXVI) are calculated to have been

5.90 and 7.97 pounds per cubic foot at the end of the nine-week weathering cycle (Table XXXIII); and by extraction these residuals contained 3.03 and 3.12 pounds per cubic foot of the fractions distilling below 355°C. On the basis of residue change alone these amounts are calculated at 2.73 and 2.36 pounds respectively. Rhodes states[89 p. 76] "In no case was a weathered specimen attacked by the fungus". Paraphrasing his next sentence, this proved that both Creosote I and Creosote II *at a treatment retention of 16.7 pounds per cubic foot* "were affording adequate protection at the end of nine weeks, equivalent to many years of actual service". Was there any reason to expect such specimens to decay?

It is easier to attempt an answer to that question now than it was in 1934. Duncan[41] shows a treatment threshold retention for "conditioned", i.e., unweathered, blocks of 1.6 pounds per cubic foot by agar-block tests, which is in close agreement with an average of 1.56 calculated from recent European tests reported by Schulze, Theden and Starfinger.[54(1)] Of more importance for the question at hand is Duncan's weathered agar-block creosote threshold, given as 5.0 pounds per cubic foot. The loss in creosote at this 5 pound level was 57 per cent, which left 2.2 pounds per cubic foot of residual creosote in the blocks. The residue is calculated to have risen, as a result of weathering losses of the lower fractions, from an original 20.9 per cent (Table II) to 48.5 per cent; and on the basis of this figure the 2.2 pounds of residual oil consisted of 1.07 pounds per cubic foot of the fraction above 355°C and 1.13 pounds of the fraction below 355°C.

The Rhodes' nine-week weathered blocks still contained roughly two to three times this amount below 355°C. One may conclude that the nine weeks weathered blocks should not have shown decay under the culture conditions; and certainly none of the more briefly weathered blocks should have shown evidence of attack. If original retentions of 10 and 8 pounds of creosote had been used in the experiments the test fungus might have attacked nine and twelve weeks weathered 8-pound blocks treated with either Creosote I or Creosote II (Table XXXIV).

Using the results of the Madison tests shown in Table XXXV and Fig. 34 as indices of what might have happened if Rhodes, Roche and Gillander had used 16.7, 10.0 and 8.0 pounds at treatment and if — instead of employing an approximate agar-block technique — they had run their evaluation tests by the soil-block method, one would have expected decay to show up as indicated by the horizontal lines in the data columns in Table XXXIV. In other words, in the 16.7 pound treatments *Lentinus lepideus* would probably have attacked the blocks if they had been weathered twelve weeks; and the 10.0 blocks would have been attacked at the end of five weeks weathering; while the 8-pound blocks would have been attacked after three weeks and about two weeks in the case of Creosote I and II, respectively. Under all these assumptions the treatment threshold for creosote in these block tests would probably have been set at above 8 pounds, and possibly around 9 pounds or more per cubic foot, in 1934; which would agree very well with the evidence obtained from soil-block tests, ¾-inch stake tests, and test posts that has been presented in this paper.

Older records show very substantial quantities of the fraction below 355°C remaining in well treated wood after long service. Alleman's 1907 paper[1] is a classic. Writing of the increasing use of creosoted wood he states: "Recent reports . . . have clearly shown that, while proper treatment gives remarkably good results, much of this timber was not properly treated and has not lasted as it should". All of which in his opinion ". . . makes it imperative that we should know, as completely as possible, just what constitutes efficient creosote treatment. The different sorts of oils are believed to have different preservative values when injected into timber, but there is, unfortunately, a lack of uniformity in opinion".

Alleman chose, as the best method for finding some of the answers, an extraction of oils from treated timber that had given good service. As solvents he used absolute alcohol and subsequently anhydrous benzene. He fractionated the extracted creosotes with a view to determining the character of the oils, deciding to make his cuts so as to collect the distillate as follows: I to 170°C; II 170–205°C; III 205–245°C, which he regarded as the naphthalene fraction; IV 245–270°C; V 270–320°C; VI 320–420°C; and VII the residue above 420°C.

The wood from which he extracted the creosotes consisted of ties, mostly British, piles from England and the United States, paving blocks, and a section of creosoted wood duct removed in perfect condition after fourteen years service in Bell Telephone plant in Philadelphia. The English piles had been in service forty-three years, the other old samples all averaged a little over twenty years. Alleman's extractions showed — after all these years in use — that there were on the average over 9 pounds per cubic foot of oil remaining in the ties and English piles; nearly 9 pounds remaining in the conduit; and about 16 pounds remaining in the American piles and the paving blocks.

The writer has calculated the residue above 355°C in these extracted oils to have varied from about 23 to about 42 per cent; and the pounds per cubic foot of oil distilling below 355°C remaining in the treated wood ran from a low of about 5.5 pounds to about 11.0 pounds. The wood had apparently remained sound. Alleman cites the difficulties of arriving at precise judgments but concludes "that *10 pounds* of creosote per cubic foot is ample for railroad ties, and that piles require from 10 to 20 pounds" according to location.

Alleman's discussion of the relative amounts of *light* and *heavy* oil that might be desirable are not applicable to present day oils and commercial conditions. His perplexities remain — in almost identical form or in modernized version — and his extraction results are a long way from those reported by Lumsden,[76] and those cited elsewhere in this paper. Breazzano was beginning the application of biological tests in Italy[20] with a view to correlating chemical and fungicidal characteristics of preservatives. This type of endeavor was later to be pressed vigorously by Bateman,[8] whose work has already been mentioned.

If the interpretation offered is supported by additional experiments already under way the Madison data in Table XXXV and Figs. 34 and 35 will be recognized as representing one of the most consistent series of laboratory tests for the evaluation of creosotes that has ever been run.

The Evaluation of Greensalt

The satisfactory performance of posts and poles treated with greensalt[74, 79] has been reported in Lumsden's paper.[76] So far the very satisfactory results at the Gulfport test plot are accurate indices of what has been found by examination of poles in line. The only data on greensalt treated ¾-inch stake tests reported in this paper are shown in Table XII and Figs. 19 and 23. Summaries of additional stake data are now in preparation for publication.

The indications are that the threshold for greensalt under Gulfport conditions is 1.42–2.1 pounds per cubic foot for ¾-inch stakes. The average life estimates for the two treatment groups — 0.57 and 1.17 pounds of dry salt per cubic foot — (Table XIII) compare most favorably with the estimates for the four creosotes in the same table. The number of greensalt specimens is large enough to warrant the conclusion that the lines used for estimating the threshold in Fig. 23, in their trend to fall off to the right soon after the sixth year of exposure, indicate that particular period as a critical one for comparisons and interpretations. Commercially treated southern pine poles meeting the standard specification requirements for retention — 1 pound of dry salt per cubic foot — have about 2.0 pounds of dry salt in the outer inch. The agreement between the stake and post tests seems good.

In soil-block tests[70, 95] *Poria incrassata, Poria monticola* and *Lenzites trabea* have been resistant to greensalt K, whereas *Lentinus lepideus* is very susceptible. The writer interprets the results of these three types of evaluation procedure, by soil-block, by ¾-inch stakes, and by posts or poles to mean that:

1. The conditions for decay, as far as *Lenzites trabea* are concerned, are much more favorable in the soil-block culture bottle than they are in the above ground part of a pole under normal outdoor exposure conditions;
2. The incidence of attack or infection at the ground line by *Lenzites trabea, Poria incrassata* and *Poria monticola* at Gulfport is relatively rare; and that
3. The success of greensalt K in southern pine poles may be attributed in large part to the susceptibility of the ubiquitous *Lentinus lepideus* to the combination of salts in the greensalt preservative.

Incidentally one may cite the following example of confirming results in tests of another salt preservative. Harrow's[51] experiments with soil-block tests resulted in locating a threshold for zinc chloride on unweathered blocks at 0.28 pounds per cubic foot (at treatment) for *Lenzites trabea*, Madison 617 and 0.53 for *Poria vaporaria*. Richards and Addoms[95] found approximately 0.25 for Madison 617, and approximately 0.50 for *Poria monticola*, Madison 698. These two Porias are possibly the same species. The similarity in the thresholds appears to be the definite result of following the same technique, rather than a haphazard coincidence.

The Evaluation of Pentachlorophenol

The highly toxic properties of pentachlorophenol have been established by exhaustive Petri dish agar toximetric tests.[53] A 5 per cent solution of penta in a light petroleum solvent is the preservative of reference in the recommended standard test for evaluating oil-soluble wood preservatives of the National Wood Manufacturers Association.[109] This test, as pointed out previously, is an agar-block test. Duncan reports[41] that threshold determinations based on soil-block tests of a 5 per cent solution of penta in petroleum (cf. Tables V and VI) against *Lenzites trabea*, a critical fungus for this preservative, have not varied more than ± 0.2 pounds from 4.8 pounds per cubic foot in 7 series of weathered block experiments over a five-year period. Recent Bell Telephone Laboratories' soil-block tests confirm this result, with the same fungus and the same petroleum carrier.

These results are confirmed at the Laboratories' Gulfport test plot (unpublished data). In two separate series — 1937 and 1938 — ¾-inch stakes were treated at retentions slightly below the threshold cited by Duncan[41] with 5 per cent solutions of pentachlorophenol in light petroleum (gas oil) and with two coal tar creosotes. The performance after six and seven years was approximately the same for both the penta solutions and the creosotes. However, completely favorable results on test posts have been reported for Gulfport[76] and Saucier, Miss., test plots. Early tests on 2 x 4 inch stakes are now being critically examined, and more tests are in progress.[18] Penta treated posts are installed in the Saucier plot, where they are under periodic observation and comparison along with posts treated with the cooperative creosotes.[12] All of these experiments will greatly facilitate correlation of the results of different test methods.

As far as pole line tests are concerned one can only echo the report[76] that up to this time not one of the tens of thousands of poles in line that were treated with either straight penta-petroleum or with mixtures of penta-petroleum and creosote have been reported as failing because of decay.

Swedish Creosote Evaluation Tests

Rennerfelt and Starkenberg[88] report that of fourteen stakes measuring 1.5 x 1.5 x 100 cm, that were cut from the middle and inner sapwood of creosoted Scotch pine poles, none are sound after ten years in the test plot (May, 1950). The stakes were rated as 3 with slight decay, 8 with medium decay, and 3 with severe decay. Apparently these results cannot be correlated with definite treatment retentions.

On the other hand, the same authors state that stakes measuring 2 x 5 x 50 cm, treated to an average retention of 5.55 pounds per cubic foot of creosote (undiluted) have all decayed in a 4.5-year exposure period in greenhouse decay chamber tests. Additional experiments have been started, presumably with stakes at higher retentions, "in order to determine whether it is possible to correlate results from such decay chamber experiments with the results obtained in field and service tests."

In another series of experiments Rennerfelt and Starkenberg find after seven years (as of May, 1950) that creosoted stakes measuring 2 x 5 x 50 cm are showing different degrees of resistance to wood-destroying fungi in their four different test plots. The difference in behavior in different test plots — which is more or less to be expected — holds true for salt as well as creosote treatments. Creosote and Bolidens (zinc-chromium-arsenic) are the better performing preservatives, with creosote in the lead. However, there have been a total of three failures in the creosoted stakes treated with average retentions of 3.6 (two stakes) and 5.6 (one stake) pounds of creosote. Stakes treated to an 8.6 pound retention are showing slight to medium decay, and in two plots slight decay has been found on a total of three stakes treated to a 12.1-pound retention. The stakes are all treated without the addition of any diluent, such as toluene, to the creosote. The reader, bearing in mind the differences in the site conditions, can make interesting comparisons between the small stake test results obtained in Sweden and in the Gulfport test plot.

Of further interest, however, is the fact that in the Swedish tests, round posts treated with average retentions of 5.37 and 5.80 pounds per cubic foot are all rated as sound after seven years exposure. One can assume that at treatment the outer annual rings of such posts contained 8 pounds or more of creosote and this amount has been sufficient to protect the posts in the Swedish climate. Rennerfelt has stated personally to the writer that one would have to proceed with caution in Sweden in the direction of increasing the creosote retention for poles, on account of public reaction against bleeding. His test results — the only ones of their kind available from Europe to the writer's knowledge — seem to be in line with Bell Telephone Laboratories findings. They would be more interesting if he had used soil-block tests for correlation.

Shortening the Bioassay Test

Besides speeding up the weathering period by the use of a weathering wheel, or by the method of alternating water and controlled heat cycles now being developed at Bell Telephone Laboratories, there are two other avenues of approach to shortening the bioassay test. One is the use of thin wood veneer test units in place of wood blocks, in the methods of impregnation and exposure to fungus action proposed by Breazzano[23, 72] and Hopkins and Coldwell.[55] Breazzano claims a maximum of accuracy because of uniform distribution in thin pieces of wood, 0.6–0.7 mm, of the preservative to be tested and because the fungus attack and passage *through* the thin strip gives a quick visual indication of the necessary protective threshold. He also claims advantages for his Italian method because it is not necessary to use any culture medium at all — he exposed his wood strips over water only (cf. Waterman et al[126]) — and because no tedious weighing techniques and record making are required. His arguments are intriguing, but his method seems to be quite out of question for testing toxicity-permanence relations of volatile preservatives like creosote. Evaporation losses would be very rapid, close to those reported by Cislak in discussion following the presentation of the Snell and Shipley paper,[108] and would approach those obtained by Boving in his impregnated filter paper experiments (see Table XXXI). Furthermore, the results of tests such as his, which involve the principle of inoculation by placing "fungus on wood," instead of "wood on fungus" as in the soil-block test, require a lot of translation to interpret their significance in practical wood preservation. Rabanus[86] brought this matter out into the debate very clearly 20 years ago. Liese et al[72] answer Breazzano's objections to the block test.

There have been no further reports on the procedures followed by Hopkins and Coldwell. Their methods are subject to some of the same criticisms that have been mentioned in the preceding paragraphs, particularly if one were to consider such techniques in a search for a way to speed up the culture tests. However, whether one agrees with them or not, their introduction of the idea of applying strength tests leads directly to a discussion of strength losses, as against weight losses, as criteria for establishing preservative thresholds.

Toughness or Impact Tests for Determining Preservative Effectiveness

Trendelenburg[115] published in 1940 his scheme for testing the strength of treated blocks that had been exposed to fungus attack. He was aiming at a technique that would shorten the time period of fungus tests on wood, but he was also looking for some other criterion than weight loss as a measure of fungus attack. He used the relative impact strength values of matched sound and decaying test specimens as indices of the degree of decay. Boards were carefully quarter-sawed first into pieces of double specimen width plus saw kerf tangentially, and of double specimen length. From these blanks four test specimens were cut that measured 8.5 x 8.5 x 120.0 mm. The pairs were considered to be matched laterally and vertically since every effort was made to cut them from the same annual rings. One piece from each pair was exposed to fungus attack and the other served as a control. The fungus cultures were made in Kolle flasks on malt agar. Specimens were placed radial side down directly on the growing fungus surface. In the pendulum testing machine the impact load was always applied to the upper radial face, so that the lower more or less infected radial face represented the tension side of the specimen in the breaking test.

Trendelenburg showed that the per cent change in strength caused by the test fungus in the early stages of decay was much more pronounced than the concomitant changes in weight or density. He called attention to the fact that the German Standard[33] for testing wood preservatives contains a stipulation that a weight loss of less than 5 per cent shall not be considered significant unless there is visual evidence of actual wood destruction by the test fungus. He presents data to show that a loss of 50 per cent of the relative original impact strength in spruce and fir occurred after about fifteen days in test, and that the weight loss for the test pieces was *about 3 per cent only*. Impact strength seems to be affected much more than the bending or compression strength. He was confident that his method would not only shorten the time of the bioassay test but also give more reliable and more significant results than those based on weight loss alone.

After Trendelenburg's death his ideas have been further tested and developed by von Pechmann and Schaile.[119] In addition to trying out the suitability of the strength test procedure, they have explored the changes in the wood structure with the microscope, and, as decay progressed, they have determined the gross relation between weight loss and solubility in sodium hydroxide.

They present as an example comparable data for the German Standard agar-block test and a test run by Trendelenburg's method, using pine wood (presumably *Pinus sylvestris*) and the test fungus *Coniophora cerebella*, against a proprietary preservative. The absorption of the preservative was essentially the same in each test; but not enough preservative was used to permit determination of the threshold. The main results were that in the impact strength test procedure in fifty days the strength reduction was 66.4 per cent and the weight loss 12.7 per cent; whereas in the standard agar-block procedure in four months the weight loss was only 2.0 per cent. Von Pechmann and Schaile feel that it is possible to save 2½ to 3 months time by using the strength test technique, and that with proper attention to detail the results will be more definite and just as reliable as those obtained in the longer period required for the standard agar-block tests.

The precise Trendelenburg technique has not been tried out in this country in any comparative tests on wood preservatives but toughness test data on small specimens of wood and veneer, sound, fungus stained and decayed, treated and untreated, have been accumulating at the Forest Products Laboratory at Madison, Wis. The fact that strength loss begins earlier and may increase more rapidly than weight loss or than change in specific gravity in natural infections in the heartwood was shown by the writer[27] and his colleagues shortly after the end of World War I. Confirming data were secured in later tests.[28, 98] Scheffer[97] showed the same results, on a more definite basis, by growing *Polystictus versicolor* on red gum sapwood in large test tubes and testing matched specimens for various strength properties as decay progressed.

The Trendelenburg technique has possibilities, but it will be some time before one can say whether it is practicable to take full advantage of it in developing supplemental bioassay tests. The one outstanding difficulty in the way of extensive use of strength tests on small specimens lies in the procurement, in the very variable southern pine, for example, of a requisite quantity of straight grained quarter sawed wood for the manufacture of the *matched blocks*. Small scale check tests are practicable, according to the writer's experience. The cost of personal supervision and manufacture of any large number of specimens would appear in advance to be exorbitant. Still, strength loss as a result of attack in treated wood is important, and Trendelenburg's ideas may win more proponents, if only as a supplemental procedure, after the soil-block technique has become more firmly established and appreciated.

OTHER ACCELERATED BIOASSAY TESTS

There are a number of items that must be mentioned before bringing this long paper toward its conclusion. There are, for example, other types of outdoor exposure tests on wood and of laboratory block tests than those cited. Two types only will be used as illustrations of the efforts that are being made to evaluate preservatives by other than the traditional service test, namely: Verrall's[120, 121] *above ground* outdoor testing procedure, and the experiments of Tippo et al[114] with large block tests devised to determine effective concentrations of preservatives for prevention or control of decay in wooden ships.

Since the soil-block test is essentially a laboratory simulation of controlled ground line conditions, there is a need for some other type of test that will approximate the above ground conditions to which treated wood may be exposed. Verrall treats pieces of dressed nominal 2 x 4 inch southern pine sapwood and exposes them to the varying wet and dry, hot and cold weather conditions at Saucier, Miss. One of two pieces has a 45° end cut. This end cut is toe-nailed to the side of the other piece, which is then nailed upright on a supporting treated or untreated rail support, with the V up. This permits maximum hazard as far as catching water is concerned. His results are furnishing valuable information about pentachlorophenol, copper naphthenate and organic mercury compounds, for example. His techniques are applicable to other preservative problems, and other investigators are using his scheme in Canada for general studies,[105, 106] and at Ann Arbor, Mich., for testing the amount of preservative needed in the upper, or above ground section, of thin sapwood poles.

The work of Tippo and his associates represents one phase of an extensive set of experiments in which large (6 x 5 x ⅞ inch) and small (3 x 5 x ⅞ inch) specimens are made up to simulate a butt-block assembly, and exposed to the attack of certain critical fungi by adding to the block assembly another inoculum block (3 x 5 x ⅞ inch) that has been thoroughly infected. The assembled units are kept in a warm and practically saturated atmosphere until the reaction of the fungus to the different preservatives can be determined. This work is being expanded in view of the importance of minimizing decay in wooden ships.

The point to be made here is that Verrall's tests and Tippo's tests should be evaluated carefully before the service test program is broadened extensively.

OTHER OBSERVATIONS

Some of the results of Suolahti's interesting studies[111] on the influence of wood at a distance on the intensity and direction of growth of fungus filaments (mycelium) have been confirmed by preliminary experiments at Bell Laboratories. Small sterilized southern pine sapwood blocks enclosed in either test tube or Petri dish cultures exert a positive pull on the filaments that is effective over a distance of several centimeters. The growth of the mycelium is more luxuriant, and the filaments are definitely drawn in the direction of the wood. Without attempting any interpretation of the significance of this phenomenon one may be permitted to point out that such studies strongly support the very great desirability — if not the necessity — of using wood in any studies directed toward evaluation of wood preservatives.

In view of the nearly forty years of prior work both here and in Europe, in which it was definitely established that certain higher fungi were the principal causes of decay in wood, it is hard to see why Weiss[128] spent so much time and such careful work on trying to test wood preservatives by using bacteria as his bioassay agents. Following the presentation of his paper before the Society of Chemical Industries in 1911 some of Weiss's critics pointed out that his methods were unrealistic as far as oil preservatives were concerned, one of his commentators suggesting that the proper approach to the problem of preservative evaluation was *to test treated and untreated wood,* (unsterilized and sterilized) under conditions favorable for fungus growth.

Tamura in 1931[112] used an assembly of two pieces of treated wood molding between which he inserted a properly sized piece of untreated wood, the whole being exposed over the surface of an agar culture of the test fungus. He did not attempt to add a block of infected wood to his setup as Trippo did; but his procedure illustrates an attempt of some twenty years ago to test the protective action of preservatives in the laboratory. His statement that sterilization might drive off a significant amount of volatile preservative from freshly treated blocks, but that the sterilizing process would probably have very little effect on the preservative residual in weathered blocks anticipated similar views expressed in the present paper.

For testing initial toxicities one can still use the Petri dish or agar flask method; but it is about as unrealistic as Weiss's procedure as far as tests of toxicity and permanence of wood preservatives are concerned. The results can be presented for their academic interest, and the investigator can keep safely aloof from the perilous practical problems of wood preservation unless he attempts to translate his data into terms of *permanence and preservative value.* Then his practical colleagues as well as his technical friends point out to him, truly with a vengeance, the error and unrealistic character of his efforts.

It may be, as Rabanus[86] has suggested, that closely similar results can be obtained by the agar and by the agar-block method in the case of certain definite toxic chemicals, particularly water soluble ones. If so, the Petri dish or agar-flask method could be used with such preservatives, and the results of the tests could be applied in practice, *after* such agreement between agar and block tests was firmly proven and established.

It is completely unrealistic to attempt to arrive at significant values for the volatile fractions of creosote by confining them within a tightly closed culture dish. If such materials are really transient their toxic function can operate only during the early life of the treated wood. In a whole creosote, for example, the relatively higher toxicity of the volatile low boiling fractions supplies an important initial power to the preservative, which power is evidently diminished as the volatiles leave the treated wood. The degree of change in toxicity, as measured by the agar method, in new and aged or weathered creosotes, as shown by the works of Snell and Shipley,[108] Schmitz et al,[102, 103] Baechler[5] and others is distinctly realistic as an index of how an oil may be altered by time and weather — at least with respect to a measure of its toxic properties. Such changes have great practical significance when minimum quantities of a preservative are employed, either for reasons of economy or in order to insure cleanliness in the treated product. The trouble is that the results of the Petri dish or agar-flask method do not indicate directly *how much* preservative to use with a view to providing the necessary residual and effective preservative.

The agar-block tests are somewhat better in this respect. While additional proof is necessary the results of such tests may be good indices of retention requirements for certain water-borne preservatives. The comparison of agar-block and soil-block tests made by Warner and Krause[122] is incomplete, and therefore somewhat unsatisfactory, particularly in view of the title of their article. They do not follow through, and they repeat some of the inferential objections raised by others as to the effect of soil differences and methods of interpretation. The comparison by Finholt et al[46, 47] is also inconclusive chiefly because of the nature and design of their experiments. So also, but to a lesser degree, are the comparisons that one may draw from the first two papers on the soil-block and agar-block tests from Madison,[95, 35] which after all, were in the nature of preliminary or reconnaisance studies, introducing the first trials of an outdoor weathering technique, and developing the necessary steps in the broader and more comprehensive plans followed later.

Duncan[41] has now brought out a full scale comprehensive study of the agar-block and soil-block techniques, which, however, deals with oil type preservatives only. She shows definitely — as was indicated in the earlier Madison work — that the test fungi are more aggressive under the more natural and more realistic environment of the soil-block cultures. Definite evidence of the very important better control of moisture content in the soil-block tests is presented. The soil-block thresholds for the different preservatives are generally higher in the soil-block tests, although the *order* of effectiveness is essentially the same.

Sedziak's[106] recent paper comparing results of his tests on buried soil-blocks and results of tests by a soil-block technique approximate to that used at Madison and at Bell Laboratories is not convincing with respect to the implied superiority of the buried block method. His paper covers work begun after the early soil-block was started at Bell Telephone Laboratories, but before the extensive experiments at Madison were initiated. Satisfactory comparison of the work of the Madison and Ottawa laboratories is difficult because Sedziak has used a different set of test organisms, including the European *Coniophora cerebella* for example, and *Lenzites sepiaria*, which is apparently not a satisfactory discriminating organism for creosote and pentachlorophenol. He has omitted the very critical *Lentinus lepideus* that has been employed at Madison since 1944. While he interprets threshold retentions for penta and for copper naphthenate that are close to those obtained at Madison, the steps in his gradient retentions leave one wishing that the Madison and Ottawa plans could have been more closely harmonized before the most recent Ottawa work was started, at least to the point of some tests with the same procedures and same test fungi. However, nothing in Sedziak's results negates the general conclusions reached at Madison and at Bell Laboratories about the value of the present soil-block technique for the testing of creosote and other oil type preservatives.

CONCLUSIONS

1. In the course of this paper evidence and interpretations have been presented to show that the soil-block technique incorporating a weathering procedure is a practical, rapid method of bioassay and that the results obtained from this method are in general agreement with accelerated stake and long time pole-diameter post tests on the same or similar preservatives. For example, it is shown that a creosote retention at treatment of 9–10 pounds per cubic foot is necessary to insure a satisfactory degree of preservative permanence in test blocks, in ¾-inch stakes and in the outer 1 inch of test posts. There is no reason to believe that this minimum limit does not apply to the outer 1 inch of poles in line.

2. The good reputation of well creosoted material is reaffirmed by these findings. Moreover, they show why failures have occurred and indicate what should be done to forestall such failures.

3. Since the results of the block tests are essentially the same as the results of the much longer stake and post tests, the block test data can be used at once as a basis for the establishment of the necessary amounts of the respective preservatives distribution in the wood where they will do the most good. The possibilities of bleeding increase as the retention is increased, so the bioassay technique becomes an essential tool for closer appraisal of effective wood preserving power.

4. It is important now to recognize that the soil-block results with creosote, for example, reveal the fact that the results of the European agar-block tests are — in all cases — too low to represent indices of actual requirements in treated wood. Therefore, the results for creosote tabulated by Schulze, Theden and Starfinger[54 (1)] have to be corrected upward by some multiplying factor; perhaps of the order of three or four, before they can be correlated with the results on blocks, stakes and posts presented in this paper. For the true scientific solution of the problems of these different techniques, perhaps an international task force may be required.

5. The interpretations presented in this paper indicate that the use of

the Laboratories' controlled weathering procedure will provide a means for determining truly effective threshold retentions for oil-type and salt-type preservatives for comparable service requirements.

6. These threshold determinations are supplying data that would have been most valuable in planning retention gradients in small stake and pole diameter post tests in test plots; and, in general, the bioassay tests explain and confirm the unequivocal results of experience.

7. Through the soil-block test a ready method is available for use in the quality control of a wood treater's current product at plants where large supplies of preservatives are received from one or more constant sources and are stored in bulk. No present method of bioassay control is sufficiently rapid to be effective or practicable on mixed samples taken at treating plants receiving preservatives at frequent intervals in small lots from varied sources.

8. The soil-block development may soon make it possible to reach approximately the ideal in which the long-time service tests of treated material in line will confirm the Laboratories' rapid test results with respect to preservative requirements. Coupled with results-type requirements (wherein the end product — not the steps of manufacture and treatment — are defined) viz., (a) retention *in the wood*, (b) penetration, and (c) cleanliness, there will then be even better assurance than at the present of the quality performance always expected under Bell System specifications.

ACKNOWLEDGEMENTS

It is hoped that the preceding pages will be accepted by the reader for what they are — condensed results of teamwork over the years — into which have been woven some individual ideas, opinions and interpretations. The writer is responsible for the literature review and for the selection of the items discussed. In one way or another various members of the Timber Products Group of Bell Telephone Laboratories have contributed to the collection and analysis of the supporting data. Thanks are due especially for help on the soil-block section to J. Leutritz, Jr., L. R. Snoke and Ruth Ann MacAllister; on the ¾-inch stake section to J. Leutritz, Jr.; on the text post and pole line sections to G. Q. Lumsden and A. H. Hearn; on the review and editing of the manuscript to R. J. Nossaman, F. F. Farnsworth and G. Q. Lumsden; on the zealous checking and assembly of test, tables and figures to Jean E. Perry; and on the correlation of the Madison cooperative test data to Catherine G. Duncan. Dorothy Storm's untiring efforts as a secretarial task force are deeply

appreciated. Without such help from these people the paper could not have been prepared in its present substantial form and substance.

BIBLIOGRAPHY

1. Alleman, G., Quantity and Character of Creosote in Well-Preserved Timbers. Circ. 98, Forest Service, U. S. Dept. Agric., May 9, 1907.
2. Alliot, H., Mèthode d'Essais des Produits Anticryptogamiques. Inst. Nat. du Bois Bull. Techn. 1, 1945.
3. Amadon, C. H., Recent Observations on the Relation between Penetration, Infection and Decay in Creosoted Southern Pine Poles in Line. Am. Wood Preservers' Assoc., Proc., **35**, pp. 187–197, 1939.
4. Baechler, R. H., Relations Between the Chemical Constitution and Toxicity of Aliphatic Compounds. Am. Wood Preservers' Assoc., Proc., **43**, pp. 94–111, 1947.
5. Baechler, R. H., The Toxicity of Preservative Oils Before and After Artificial Aging. Am. Wood Preservers' Assoc., Proc., **45**, pp. 90–95, 1949.
6. Bateman, E., Quantity and Quality of Creosote Found in Two Treated Piles After Long Service. Circ. 199, Forest Service, Forest Products Laboratory Series, U. S. Dept. Agric., May 22, 1912.
7. Bateman, E., Coal-Tar and Water-Gas Tar Creosotes: Their Properties and Methods of Testing. U. S. Dept. Agric., Bul. No. 1036, pp. 57–65, Oct. 20, 1922.
8. Bateman, E., A Theory on the Mechanism of the Protection of Wood by Preservatives. Am. Wood Preservers' Assoc., Proc., 1920, p. 251; 1921, p. 506; 1922, p. 70; 1923, p. 136; 1924, p. 33; 1925, p. 22; 1927, p. 41.
9. Bavendam, W., Die pilzwidrige Wirkung der im Holzschutz benutzten Chemikalen. Mitteilungen d. Reichsinst. f. Forst und Holzwirtschaft. No. 7, 20 pp. Aug., 1948.
10. Baxter, D. V., *Pathology in Forest Practice.* 2nd Ed. XI + 601 pp. John Wiley and Sons, New York, 1952.
11. Bertleff, V., Prüfung der Fungiziden Eigenschaften Ostrauer Steinkohlenteer Imprägnieröle und ihrer Bestandteile. Chem. Zeitung, **63**, p. 438, 1939.
12. Bescher, R. H., and others, Cooperative Creosote Tests. Am. Wood Preservers' Assoc. ,Proc., **46**, pp. 68–79, 1950.
13. Bienfait, J. L., and T. Hof, Buitenproeven met geconserveerde palen. I ste mededeling. Circ. 8 Ser. III, Conservering en Veredeling No. 3, Central Instituut voor Materiaal Onderzoek. Delft, Holland, Dec., 1948.
14. Bland, D. E., A Study of Toxicity of Australian Vertical Retort Creosote Oils to *Lentinus Lepideus* Fr., *Polystictus Versicolor* (L) Fr., and Madison 517. Australian Council Sci. and Ind. Jl. **15**, pp. 135–146, 1942.
15. Blew, J. O., Comparison of Wood Preservatives in Stake Tests. Am. Wood Preservers' Assoc., Proc., **44**, pp. 88–119, 1948.
16. Blew, J. O., C. A. Richards, and R. H. Baechler, Evaluating Wood Preservatives. For. Prod. Res. Society, Proc., **5**, pp. 230–238, (Wood-block tests, pp. 235–236), 1951.
17. Blew, J. O., Comparison of Wood Preservatives in Mississippi Post Study (1952 Progress Report). Report No. R1757, Forest Products Laboratory, U. S. Forest Service, Madison, Wis.
18. Blew, J. O., Comparison of Wood Preservatives in Stake Tests (1952 Progress Report). Report No. D1761, Forest Products Laboratory, U. S. Forest Service, Madison, Wis.
19. Boyce, J. S., *Forest Pathology.* 550 p. McGraw-Hill Book Company, Inc., New York, 1948.
20. Breazzano, A., Metodo Biologico di Controllo dei Sistemi di Preservazione, dei Legnami Adottato dall' Instituto Sperimentale delle FF. SS. (English translation: Biological Method of Testing Systems of Wood Preservation Adopted by the State Railway Experiment Institute.) Tech. Rev. Italian Railways, **4**, No. 5, pp. 3–8, Nov., 1913. Forest Products Laboratory, Madison, Wis.

21. Breazzano, A., Metodi Normali di Prova sulla Putrescibilitá dei Legnani. (English translation: Standard Methods of Testing the Putrescibility of Wood.) Extract from Report of the Ninth Meeting of the Italian Association for the Study of Building Materials; Turin; April, 1922. Forest Products Laboratory, Madison, Wis.
22. Breazzano, A., Osservazioni sul Metodo dei Blocchetti di Legno in Usa nell Analisi Tossimetrica delle Sostanze Conservatrici del Legno. Revista Technica Delle Ferrovie Italiane. **47,** No. 6, June 15, 1935. (Observations on the Wood Block Method in the Toximetric Analysis of Wood Preservatives. Excerpt from Technical Review of Italian Railways. English translation; Forest Products Laboratory.)
23. Breazzano, A., Methods for Determining the Fungicidal Power of Wood Preservatives. International Assn. for Testing Materials, Proc., London Congress, Group C—Organic Materials, Sub. group 3, pp. 484–486, April, 1937.
24. British Standard Method of Test for the Toxicity of Wood Preservatives to Fungi. British Standards Institution. British Standard No. 838, 17 pp., April, 1939.
25. Broekhuizen, S., Onderzoekingen over de Conserverende Waarde van een Aantal Houtconserveermiddelen. Rapp. Comm. Gebruikw. inh. hout. Deel II, Bijlage II, pp. 89–122, The Hague, 1937.
26. Cartwright, K. St. G., and W. P. K. Findlay, *Decay of Timber and its Prevention*. VI plus 294 p. London: His Majesty's Stationery Office, 1946. Reprinted 1948.
27. Colley, R. H., The Effect of Incipient Decay on the Mechanical Properties of Airplane Timber. (Abstract) Phytopathology, **11,** p. 45, 1921.
28. Colley, R. H., T. R. C. Wilson, and R. F. Luxford, The Effect of *Polyporus Schweinitzii* and *Trametes Pini* on the Shock-resistance, Compression Parallel to Grain Strength, and Specific Gravity of Sitka Spruce. Forest Products Laboratory, Project L-243-J1, Typewritten Report, 29 p., 32 figs., 4 plates. July 3, 1925.
29. Colley, R. H., and C. H. Amadon, Relation between Penetration and Decay in Creosoted Southern Pine Poles. Bell Sys. Tech. Jl., **15,** pp. 363–379, July, 1936.
30. Colley, R. H., Some Observations on the Selection and Use of Modern Wood Preservatives. Reports, Twenty-fourth Session, Communications Section, Association of American Railroads, October, 1947, pp. 17–25.
31. Colley, R. H., Wood preservation and Timber Economy. Forest Products Institute of Canada. Papers presented at the First Annual Convention, Ottawa, Oct. 30–31, 1950.
32. Curtin, L. P., B. L. Kline, and W. Thordarson, V—Weathering Tests on Treated Wood. Ind. and Eng. Chem. **10,** No. 12, pp. 1340–1343, Dec., 1927.
33. DIN (Deutsche Normen) DVM 2176, Blatt 1. Prüfung von Holzschutzmitteln. Mykologische Kurzprüfung (Klötzchen Verfahren). Berlin, Aug. 1939. (New Edition DIN 52176).
34. DIN DVM 52176, Blatt 2. Prüfung von Holzschutzmitteln. Bestimmung der Auslaugbarkeit. Berlin, May 1941. (Reprinted Oct., 1948).
35. Duncan, C. G., and C. A. Richards, Methods of Evaluating Wood-Preservatives: Weathered Impregnated Wood Blocks. Am. Wood Preservers' Assoc., Proc., **44,** pp. 259–264, 1948.
36. Duncan, C. G., A comparison of Two English Creosotes Produced from Coke-oven Coal Tar and Vertical-retort Coal Tar. Mss. Office Report, Division of Forest Pathology, Bur. Pl. Ind., Forest Prod. Lab., Madison, Wis., Jan. 21, 1949.
37. Duncan, C. G., and C. A. Richards, Evaluating Wood Preservatives by Soil-block Tests: 1. Effect of Carrier on Pentachlorophenol Solutions; 2. Comparison of a Coal Tar Creosote, a Petroleum Containing Pentachlorophenol or Copper Naphthenate and Mixtures of Them. Amer. Wood Preservers Assoc., Proc., **46,** pp. 131–145, 1950.
38. Duncan, C. G., and C. A. Richards, Evaluating Wood Preservatives by Soil-Block Tests: 3. The Effect of Mixing a Coal Tar Creosote and a Penta-

chlorophenol Solution with a Petroleum; a Creosote with a Coke Oven Tar or Pentachlorophenol Solution. Amer. Wood Preservers' Assoc., Proc., **47**, pp. 264–274, 1951.

39. Duncan, C. G., and C. A. Richards, Evaluating Wood Preservatives by Soil-Block Tests: 4. Creosotes. Amer. Wood Preservers' Assoc., Proc., **47**, pp. 275–287, 1951.
40. Duncan, C. G., Evaluating Wood Preservatives by Soil-Block Tests: 5. Lignite-Tar and Oil-Tar Creosotes. Amer. Wood Preservers' Assoc., Proc., **48**, 1952.
41. Duncan, C. G., Soil-Block and Agar-Block Techniques for Evaluation of Oil-Type Wood Preservatives: Creosote, Copper Naphthenate and Pentachlorophenol. Division of Forest Pathology, Bur. Pl. Ind., Forest Products Lab, Madison, Wis., Special Release No. 37, Jan., 1953.
42. Eden, Johan, och Erik Rennerfelt, Undersökninar enligt klotsmetoden av nägra fraimpregneringsmedel. (Studies on Wood Preservatives, According to the Block Method) Meddelanden från Statens Skogsforskningsinstitut, Bd. **35**, No. 10, 1946.
43. Eden, Johan, och Erik Rennerfelt, Fält och rötkammarförsök avsedda att utröna skyddsverkan hos olika fräimpregneringsmedel. (Field and Decay-chamber Experiments to Ascertain the Protective Effect of Various Wood Preservatives.) Meddelanden från Statens Skogsforskningsinstitut. Bd. **38**, No. 4, 1949.
44. Falck, R., Die wichtigsten reinen Holzschutzmittel, die Methoden ihre Wertzahlen, Eingenschaften und Anwendung. Hausschwammforschungen, **8**, pp. 18–20, 1927.
45. Findlay, W. P. K., A Standard Laboratory Test for Wood Preservatives. British Wood Preserving Assoc., Jl., **5**, pp. 89–93, 1935.
46. Finholt, R. W., Improved Toximetric Agar-dish Test for Evaluation of Wood Preservatives. Anal. Chem., **23**, No. 7, pp. 1039–1039, July, 1951.
47. Finholt, R. W., M. Weeks, and C. Hathaway, New Theory on Wood Preservation. Ind. and Eng. Chem. **44**, No. 1, pp. 101–105, Jan., 1952.
48. Flerov, B. C., und C. A. Popov., Methode zur Untersuchung der Wirkung von antiseptische Mitteln auf holzzerstörende Pilze. Angew. Bot. **15**, pp. 386–406, 1933.
49. Frosch, C. J., V—The Correlation of Distillation Range with Viscosity of Creosote. Physics, **6**, pp. 165–170, May, 1935.
50. Gillander, H. E., C. G. King, E. O. Rhodes, and J. N. Roche, The Weathering of Creosote. Ind. and Eng. Chem., **26**, No. 2, pp. 175–183, Feb., 1934.
51. Harrow, K. M., Toxicity of Water-Soluble Wood-Preservatives to Wood-Destroying Fungi. New Zealand Jl., Sec. B., **31**, No. 5, pp. 14–19, Mar., 1950.
52. Harrow, K. M., Note on the Soil Moisture Content Used with the Leutritz Technique for Testing Toxicity of Wood Preservatives Against Fungi. New Zealand Jl. Sci. and Tech., **4**, pp. 39–40, Jan., 1951.
53. Hatfield, I., Information on Pentachlorophenol as a Wood Preserving Chemical. Am. Wood Preservers' Assoc., Proc., **40**, pp. 47–65, 1944.
54. Holzschutzmittel Prüfung und Forschung III. Wissenschaftliche Abhandlungen der Deutschen Materialprüfungsanstalten: Berlin-Dahlem. II Folge, Heft 7, 132 p. Springer-Verlag, Berlin, Göttingen, Heidelberg, 1950.
    1. Schulze, B., G. Theden u K. Starfinger, Ergebnisse einer vergleichenden Prüfung der Pilzwidrigen Wirksamkeit von Holzschutzmitteln. pp. 1–40.
    2. Becker, G., Ergebnisse einer vergleichenden Prüfung der insektötenden Wirkung von Holzschutzmitteln. II Teil. pp. 40–62.
    3. Becker, G., Prüfung der "Tropeneignung" von Holzschutzmitteln gegen Termiten. pp. 62–76.
    4. Becker, G. u B. Schulze, Laboratoriumsprüfung von Holzschutzmitteln gegen Meerwasserschädlinge. pp. 76–83.
55. Hopkins, C. Y., and B. B. Coldwell, Surface Coatings for Rotproofing Wood. Canadian Chemistry and Process Industries. N. R. C. No. 1256, Dec., 1944.
56. Howe, P. J., Weathering and Field Tests on Treated Wood. Amer. Wood Preservers' Assoc., Proc., 1928, p. 192.

57. Hubert, E. E., A study of Laboratory Methods Used in Testing the Relative Resistance of Wood to Decay. Univ. of Idaho Bulletin, **34,** No. 15, July, 1929.
58. Hubert, E, E., *An Outline of Forest Pathology.* 543 pp. John Wiley and Sons, New York, 1931.
59. Hudson, M. S., and R. H. Baechler, The Oxidation of Creosote—Its Significance in Timber Treating Operations. Amer. Wood Preservers' Assoc., Proc., **36,** pp. 74–112, 1940.
60. Hudson, M. S., Poria radiculosa, a Creosote Tolerant Organism. For. Prod. Res. Soc., Jl., **2,** No. 2, pp. 73–74, June, 1952.
61. Humphrey, C. J., and R. M. Flemming, The Toxicity to Fungi of Various Oils and Salts, Particularly Those Used in Wood Preservation. Bul. No. 227, U. S. Dept. Agric., Aug. 23, 1915.
62. Hunt, G. M., and G. A. Garratt, *Wood Preservation.* VIII, 457 p. McGraw-Hill, New York, 1938. (In course of revision.)
63. Hunt, G. M., and T. E. Snyder, An International Termite Exposure Test—Twenty-First Progress Report. Amer. Wood Preservers' Assoc., Proc., **48,** 1952.
64. Jacquiot, C., Controle de l'Efficacité des Fongicides Utilisés pour l'Imprègnation des Bois. Étude Critique de la Technique Standard Anglaise et de la Norme Allemande DIN DVM 2176. Principes pour l'Établissement d'une Norme Française. Extr. d'Ann. l'Ecole Eaux Forêts, **8,** pp. 185–206, 1942.
65. Kaufert, F. H., A Survey of Laboratory Methods Used in the Evaluation of Wood Preservatives. Report of Committee P-6, Appendix A. Amer. Wood Preservers' Assoc., Proc., **45,** pp. 55–59, 1949.
66. Leutritz, J., Jr., The Toxic Action of Various Compounds on The Fungus *Lentinus Lepideus* Fr.). Unpublished Thesis, Columbia University, Nov., 1933.
67. Leutritz, J., Jr., Laboratory Tests of Wood Preservatives. Bell Lab. Record, **16,** No. 9, pp. 324–328, May, 1938.
68. Leutritz, J., Jr., Acceleration of Toximetric Tests of Wood Preservatives by the Use of Soil as a Medium. Phytopathology, **39,** No. 10, pp. 901–903, Oct., 1939.
69. Leutritz, J., Jr., Outdoor Tests of Wood Preservatives. Bell Lab. Record, **22,** No. 4, pp. 179–182, Dec., 1943.
70. Leutritz, J., Jr., A Wood-Soil Contact Culture Technique for Laboratory Study of Wood-Destroying Fungi, Wood Decay and Wood Preservation. Bell Sys. Tech. Jl., **25,** No. 1, pp. 102–135, Jan., 1946.
71. Liese, Nowak, Peters, Rabanus, Krieg, and Pflug., Toximetrische Bestimmung von Holzkonservierungsmitteln. Angew. Botanik, pp. 484–488, Nov.–Dec., 1935.
72. Liese, J. et al., Toximetrische Bestimmung von Holzkonservierungsmitteln. Angew. Chemie, 48, Beihefte 11, 1935.
73. Loseby, P. J. A., and P. M. D. Krog, The Persistence and Termite Resistance of Creosote and Its Constituent Fractions. Jour. South African Forestry Assoc., Jl., No. 11, pp. 26–32, June, 1944.
74. Lumsden, G. Q., and A. H. Hearn, Greensalt Treatment of Poles. Amer. Wood Preservers' Assoc., Proc., **38,** pp. 349–361, 1942.
75. Lumsden, G. Q., Proving Grounds for Telephone Poles. Bell Lab. Record, **22,** pp. 12–14, Sept., 1943.
76. Lumsden, G. Q., A Quarter Century of Evaluation of Wood Preservatives in Poles and Posts at the Gulfport Test Plot. Amer. Wood Preservers' Assoc., Proc., **48,** 1952.
77. Lutz, M. L., Méthodes Permittant de Déterminer la Résistivité des Bois Bruts ou Immunisés Soumis a l'Attaque par les Champignons Lignicoles. Ann. l'Ecole Nat. Eaux Forêts 5, pp. 317–327, 1935.
78. Mahlke-Troschel-Liese, Holzkonservierung (Wood Preservation), 3rd Ed. XII + 571 p. Springerverlag, Berlin/Göttingen/Heidelberg, 1950.
79. McMahon, W., C. M. Hill, and F. C. Koch, Greensalt—A New Preservative for Wood. Amer. Wood Preservers' Assoc., Proc., **38,** pp. 334–348, 1942.
80. Martin, S. W., Characterization of Creosote Oils. Amer. Wood Preservers' Assoc., Proc., **45,** pp. 100–130, 1949.

81. Mayfield, P. B., The Toxic Elements of High Temperature Coal Tar Creosote. Amer. Wood Preservers' Assoc., Proc., **47**, pp. 62–85, 1951.
82. Narayanamurti, D., V. Ranganathan, Ragbir Singh, T. R. Chandrasekhar, and A. Banerjee, Studies on Coal Tar Creosote as a Wood Preservative, Part II. Indian Forest Bulletin, No. 144, 1948, 7 + 43 pp., Jl. of India Press, Calcutta, 1950.
83. Peters, F., W. Krieg and H. Pflug, Toximetrische Prüfung von Steinkohlenteeröl. Chem. Zeit., **61**, pp. 275–285, 1937. (English Edition, Pub. Int. Adv. Off. Wood Pres. The Hague. 1937.)
84. Preservative Treatment of Poles. (Condensed from report by American Telephone and Telegraph Company of Aug. 3, 1931.) Am. Wood Preservers' Assoc., Proc., p. 237, 1932.
85. Preservatives Committee; Report of Committee P-6. Am. Wood Preservers' Assoc., Proc., **48**, 1952.
86. Rabanus, Ad., Die Toximetrische Prüfung von Holzkonservierungsmitteln. Angew. Bot. 13, p. 352–371, 1931. (Partial translation in English. Am. Wood Preservers' Assoc., Proc., pp. 34–43, 1933.)
87. Reeve, C. S., Comment on Creosote-Permanence Toxicity Relationships. Am. Wood Preservers' Assoc., Proc., p. 78–79, 1934.
88. Rennerfelt, Erik, och Bo Starkenberg., Träskyddskomittens fält- och rötkammarförsök. (Field and Decay-Chamber Experiments with Wood Preservatives.) Meddelanden från Statens Skogsforskningsinstitut, Bd., **40**, No. 4, 1951.
89. Rhodes, E. O., J. N. Roche, and H. E. Gillander, Creosote Permanence-Toxicity Relationships. Am. Wood Preservers' Assoc., Proc., pp. 65–78, 1934.
90. Rhodes, E. O., History of Changes in Chemical Composition of Creosote. Am. Wood Preservers' Assoc., Proc., **47**, pp. 40–61, 1951.
91. Rhodes, F. H., and F. T. Gardner, Comparative Efficiencies of the Components of Creosote Oil as Preservatives for Timber. Ind. and Eng. Chem., **22**, No. 2, pp. 167–171, Feb., 1930.
92. Rhodes, F. H., and I. Erickson, Efficiencies of Tar Oil Components as Preservative for Timber. Ind. and Eng. Chem. **25**, pp. 989–991, Sept., 1933.
93. Richards, A. P., Cooperative Creosote Program; Preliminary Progress Report on Marine Exposure Panels. Am. Wood Preservers' Assoc., Proc., **48**, 1952.
94. Richards, C. A., Methods of Testing Relative Toxicity of Wood Preservatives. Am. Wood Preservers' Assoc., Proc., **19**, pp. 127–135, 1923.
95. Richards, C. A., and R. M. Addoms, Laboratory Methods for Evaluating Wood Preservatives: Preliminary Comparison of Agar and Soil Culture Techniques Using Impregnated Wood Blocks. Am. Wood Preservers' Assoc., Proc., **43**, pp. 41–56, 1947.
96. Richards, C. A., Laboratory Decay Resistance Tests—Soil-block Method. (In Cooperative Creosote Tests by R. H. Bescher et al.) Am. Wood Preservers' Assoc., Proc., **46**, pp. 71–76, 1950.
97. Scheffer, T. C., Progressive Effects of Polyporus Versicolor on the Physical and Chemical Properties of Red Gum Sapwood. U. S. Dept. Agric., Tech. Bul., No. 527, Sept., 1936.
98. Scheffer, T. C., T. R. C. Wilson, R. F. Luxford, and Carl Hartley, The Effect of Certain Heart Rot Fungi on the Specifid Gravity and Strength of Sitka Spruce and Douglas-Fir. U. S. Dept. Agric., Tech. Bul., No. 779, 24 pp., May, 1941.
99. Schmitz, H., Laboratory Methods of Testing the Toxicity of Wood Preservaservatives. Ind. and Eng. Chem., Anal. Ed., **1**, No. 7, pp. 76–79, April, 1929.
100. Schmitz, H., and Others, A Suggested Toximetric Method for Wood Preservatives. Ind. and Eng. Chem., Anal. Ed., **2**, p. 361, 1930.
101. Schmitz, H., and S. J. Buckman, Toxic Action of Coal-Tar Creosote. Ind. and Eng. Chem., **24**, No. 7, pp. 772–777, 1932.
102. Schmitz, H., W. J. Buckman and H. von Schrenk, Studies of the Biological Environment in Treated Wood in Relation to Service Life. Changes in the Character and Amount of 60/40 Creosote-Coal Tar Solution and Coal Tar

and Decay Resistance of the Wood of Red Oak Crossties after Five Years Service. Am. Wood Preservers' Assoc., Proc., **37,** pp. 248–297, 1941.
103. Schmitz, H., H. von Schrenk, and A. L. Kammerer, Studies of the Biological Environment in Treated Wood in Relation to Service Life, III. Am. Wood Preservers' Assoc., Proc., **41,** pp. 153–179, 1945.
104. Schulze, B., und G. Becker, Untersuchungen über die pilzwidrige und insektentötende Wirkung von Fraktionen und Einzelstoffen des Steinkohlenteeröls. Holzforschung., **2,** No. 4, pp. 95–127, 1948.
105. Sedziak, H. P., The Wood-Block Soil Method of Accelerated Testing of Wood Preservatives. Report of Committee P-6, Appendix B. Am. Wood Preservers' Assoc., Proc., **46,** pp. 55–58, 1950.
106. Sedziak, H. P., The Evaluation of Two Modern Wood Preservatives. For. Prod. Res. Soc., Proc., 1952.
107. Snell, W. H., The Use of Wood Discs as a Substrate in Toxicity Tests of Wood Preservatives. Am. Wood Preservers' Assoc., Proc., **25,** pp. 126–129, 1929.
108. Snell, W. H., and L. B. Shipley, Creosotes—Their Toxicity, Permanence and Permanence of Toxicity. Am. Wood Preservers' Assoc., Proc., **32,** pp. 32–114, 1936.
109. Standard of N. W. M. A., Method for Testing the Preservative Property of Oil-soluble Wood-Preservatives by Using Wood Specimens Uniformly Impregnated. Nat. Wood Mfg. Assoc., M-1-51, April 27, 1951.
110. STAS 650-49, Incercarea Toxicitătii Substantelor de Impregnat Contra Ciupercilor. Comisiunea de Standardizare (Rumania) April 1, 1950.
111. Suolahti, Osmo, Über Eine das Wachstum von Fäulnispilzen Beschleunigende Chemischen Fernwirkung von Holz. Statens Tekniska Forskningsanstalt. 95 p. Helsinki, Finland, 1951.
112. Tamura, T., New Methods of Test on the Toxicity and Preservative Value of Wood Preservatives. Phytopathologische Zeitschrift, **3,** No. 4, pp. 421–437, 1931.
113. Teesdale, C. H., Volatilization of Various Fractions of Creosote After Their Injection into Wood. Circ. 188, Forest Service, Forest Products Laboratory Series, U. S. Dept. Agric., Oct. 17, 1911.
114. Tippo, O., J. M. Walter, S. J. Smucker, and W. Spackman, Jr., The Effectiveness of Certain Wood Preservatives in Preventing the Spread of Decay in Wooden Ships. Lloydia, **10,** pp. 175–208, Sept., 1947.
115. Trendelenburg, R., Über die Abkurzung der Zeitdauer von Pilzversuchen an Holz mit Hilfe der Schlagbiegeprüfung. Holz als Roh- und Werkstoff, **3.** No. 12, s. 397–407, Dec., 1940.
116. van den Berge, J. Beoordeeling van de Waarde van Fungicide Stoffen voor Houtconserveering. 183 p. N. V. Technische Boekhandel, J. Waltman, Delft, Holland, 1934.
117. van Groenou, H. Broese, Weatheringsproeven met Houtconserveermiddelen, (Weathering Tests with Wood Preservatives). Materiaalenkennis, **7,** No. 10, pp. 63–65, Oct., 1940.
118. van Groenou, H. Broese, H. W. L. Rischen and J. van den Berge, Wood Preservation During the Last 50 Years. XII + 318 p. A. W. Sijthoff, Leiden, Holland, 1951.
119. von Pechmann, H., u. O. Schaile, Über die Änderung der dynamischen Festigkeit und der Chemischen Zusammensetzung des Holzes durch den Angriff Holzzerstörender Pilze. Forstwissenschaftliches Zentralblatt, **69,** No. 8, s. 441–465, 1950.
120. Verrall, A. F., Progress Report on Tests of Soak and Brush Preservative Treatments for Use on Wood off the Ground. Southern Lumberman, Aug. 1, 1946.
121. Verrall, A. F., Decay Protection for Exterior Woodwork. Southern Lumberman, June 15, 1949.
122. Warner, R. W., and R. L. Krause, Agar-Block and Soil-Block Methods for Testing Wood Preservatives. Ind. and Eng. Chem., **43,** No. 5, pp. 1102–1107, May, 1951.
123. Waterman, R. E., Forecasting the Behavior of Wood Preservatives. Bell Lab. Record, **11,** No. 3, pp. 67–72, Nov., 1932.

124. Waterman, R. E., F. C. Koch, and W. McMahon, Chemical Studies of Wood Preservation. III. Analysis of Preserved Timber. Ind. and Eng. Chem., Anal. Ed., **6,** No. 6, pp. 409–413, Nov. 15, 1934.
125. Waterman, R. E. and R. R. Williams, Chemical Studies of Wood Preservation. IV. Small Sapling Method of Evaluating Wood Preservatives. Ind. and Eng. Chem., Anal. Ed., **6,** No. 6, pp. 413–418, Nov. 15, 1934.
126. Waterman, R. E., J. Leutritz, and C. M. Hill, A Laboratory Evaluation of Wood Preservatives. Bell System Tech. Jl., **16,** pp. 194–211, April, 1937.
127. Waterman, R. E., J. Leutritz and C. M. Hill, Chemical Studies of Wood Preservation: The Wood-block Method of Toxicity Assay. Ind. and Eng. Chem., Anal. Ed., **10,** No. 6, pp. 306–314, June, 1938.
128. Weiss, J. M., The Antiseptic Effect of Creosote Oil and Other Oils Used for Preserving Timber. Soc. Chem. Ind. Jl., **30,** No. 23, Dec. 15, 1911.
129. Williams, R. R., Chemical Studies of Wood Preservation. I. The Problem and Plan of Attack. Ind. and Eng. Chem., Anal. Ed., **6,** No. 5, pp. 308–310, Sept. 15, 1934.

Abstracts of Bell System Technical Papers* Not Published in this Journal

Principles and Applications of Converters for High-Frequency Measurements. D. A. Alsberg[1]. *I.R.E., Proc.*, **40**, pp. 1195–1203, Oct., 1952. (Monograph 2030).

The heterodyne method permits measurements over wide frequency bands with the standards operating at a fixed frequency. The accuracy of such measurements depends upon the performance of heterodyne conversion transducers or converters. Design principles are derived to maximize linearity and dynamic range and minimize zero corrections. These principles have been applied to point-to-point and sweep measurements of delay, phase, transmission, and impedance.

Ferroelectric Storage Elements for Digital Computers and Switching Systems. J. R. Anderson[1]. *Elec. Engg.*, **71**, pp. 916–922, Oct., 1952. (Monograph 2014).

These ferroelectric storage devices, although still comparatively new, show great promise. They can store up to 2,500 bits of information per square inch in a surface only a few thousandths of an inch thick with pulses less than a microsecond long.

Transistors in Switching Circuits. A. E. Anderson[1]. *I.R.E., Proc.*, **40**, pp. 1541–1558, Nov., 1952. Corrections to Figs. 17, 18 and 19 giving synopses published in December issue, pp. 1732 and 1733.

The general transistor properties of small size and weight, low power and voltage, and potential long life suggest extensive application of transistors to pulse- or switching-type systems of computer or computer-like nature.

It is possible to devise simple regenerative circuits which perform the normally employed functions of waveform generation, level restoration, delay, storage (registry or memory), and counting. The discussion is limited to point-contact type transistors in which the alpha or current gain is in excess of unity and to a particular feedback configuration.

Such circuits, which are of the so-called trigger type, are postulated to involve negative resistance. On this basis an analysis, which approximates the negative-resistance characteristic by three intersecting broken lines, is developed. Con-

* Certain of these papers are available as Bell System Monographs and may be obtained on request to the Publication Department, Bell Telephone Laboratories, Inc., 463 West Street, New York 14, N. Y. For papers available in this form, the monograph number is given in parentheses following the date of publication, and this number should be given in all requests.

[1] Bell Telephone Laboratories.

clusions which are useful to circuit and device design are reached. The analysis is deemed sufficiently accurate for first-order equilibrium calculations.

Transistors having properties specifically intended for pulse service in the circuits described have been developed. Their properties, limitations, and parameter characterizations are discussed at some length.

Mobility of Electrons in Germanium. P. P. Debye[1] and Esther M. Conwell[1]. Letter to the Editor. *Phys. Rev.*, **87,** pp. 1131–1132, Sept. 15, 1952.

The Telephone Industry in National Defense. C. A. Armstrong[2]. *Telephony*, **143,** pp. 44–46, 114, Oct. 25, 1952.

Infrared Absorption in High Purity Germanium. H. B. Briggs[1]. Letter to the Editor. *Jl. Opt. Soc. Am.*, **42,** pp. 686–687, Sept., 1952.

New Infrared Absorption Bands in p-Type Germanium. H. B. Briggs[1] and R. C. Fletcher[1]. Letter to the Editor. *Phys. Rev.*, **87,** pp. 1130–1131, Sept. 15, 1952.

Automatic Switching for Nation-Wide Telephone Service. A. B. Clark[1] and H. S. Osborne[2]. *A.I.E.E., Trans. Commun. and Electronics Sect.*, **2,** pp. 245–248, Sept., 1952. (Monograph 2015).

Western Electric's Service with Standards. K. B. Clarke[3]. *Standardization,* **23,** pp. 332–338, Oct., 1952.

Properties of Silicon and Germanium. Esther M. Conwell[1]. *I.R.E., Proc.*, **40,** pp. 1327–1337, Nov., 1952.

This article provides the latest experimental information on those fundamental properties of germanium and silicon which are of device interest, currently or potentially. Electrical properties, especially carrier density and mobility, have been treated in greatest detail. Descriptive material has been provided to the extent necessary to give physical background.

Effects of Space-Charge Layer Widening in Junction Transistors. J. M. Early[1]. *I.R.E., Proc.*, **40,** pp. 1401–1406, Nov., 1952.

Some effects of the dependence of collector barrier (space-charge layer) thickness on collector voltage are analyzed. Transistor base thickness is shown to decrease as collector voltage is increased, resulting in an increase of the current-gain factor (α) and a decrease in the emitter potential required to maintain any

[1] Bell Telephone Laboratories.
[2] American Telephone and Telegraph Company.
[3] Western Electric Company.

fixed emitter current. These effects are shown to lead to two new elements in the theoretical small-signal equivalent circuit. One, the collector conductance (g_c'), is proportional to emitter current and varies inversely with collector voltage. This term is the dominant component of collector conductance in high-quality junction transistors. The other element, the voltage feedback factor (μ_{ec}), is independent of emitter current, but varies inversely with collector voltage. The latter element is shown to modify the elements of the conventional equivalent tee network.

Four-Terminal p-n-p-n *Transistors*. J. J. EBERS[1]. *I.R.E., Proc.*, **40,** pp. 1361–1364, Nov., 1952.

The equivalent circuit of a *p-n-p-n* transistor is obtained. It is demonstrated that a *p-n-p* transistor and an *n-p-n* transistor can be connected so that the combination has the same equivalent circuit as the *p-n-p-n* structure. A simplified circuit is obtained which can be used when the *p-n-p-n* transistor is connected as a hook-collector transistor. A method of adjusting the current gain of *p-n-p-n* transistors by external means is given as well as experimental results.

Dynamics of Transistor Negative-Resistance Circuits. B. G. FARLEY[1]. *I.R.E., Proc.*, **40,** pp. 1497–1508, Nov., 1952.

A general method is presented for calculating approximately the behavior of many nonlinear circuits by dividing the region of operation into subregions, within each of which the circuit may be considered linear to a good approximation. The method is applied to a high-speed transistor switching circuit as an illustrative example.

Regenerative Amplifier for Digital Computer Applications. J. H. FELKER[1]. *I.R.E., Proc.*, **40,** pp. 1584–1596, Nov., 1952.

A description of the negative-resistance properties of the point-contact transistor is presented as an introduction to the description of a regenerative amplifier. The choice of circuit parameters for the amplifier is discussed and a sample design presented. The illustrative amplifier regenerates digital information at a megacycle rate and develops pulses with rise times of less than 0.05 μsec. It operates from supply voltages of $+6$ and -8 volts, with a battery drain of less than 0.05 watt. A complete set of computer building blocks has been designed around the amplifier. Their use is illustrated in two computer applications.

Evidence for Domain Structure in Anti-ferromagnetic CoO From Elasticity Measurements. M. E. FINE[1]. Letter to the Editor. *Phys. Rev.*, **87,** p. 1143, Sept. 15, 1952.

Optical Position Encoder and Digit Register. H. G. FOLLINGSTAD[1], J. N. SHIVE[1] and R. E. YAEGER[1]. *I.R.E., Proc.*, **40,** pp. 1573–1583, Nov., 1952.

The usefulness of transistors in systems has been given a feasibility proof through the construction and operation of a six-digit position encoder and serial-

[1] Bell Telephone Laboratories.

output digit register. This system performs the functions of photoelectric encoding, pulse regeneration, digit storage, reflected-to-natural binary translation, and digit shifting by means of circuits using transistors and other semi-conductor devices. The model occupies a volume of about ¼ cubic foot, weighs seven pounds, and consumes 16 watts of power.

Comparison of Recording Processes. J. G. FRAYNE[4], *I.R.E., Trans.*, PGA-7, pp. 5–8, May, 1952. *S.M.P.T.E., Jl.*, **59**, pp. 313–318, Oct., 1952.

The three common forms of sound recording may be classed as mechanical (disk), photographic and magnetic. All three methods are in common use today and each is employed in a field for which it appears to be peculiarly fitted. The purpose of this article is to examine briefly the factors which determine the fidelity of each method. By fidelity we mean how true the tonal range can be reproduced, the amount and nature of harmonic distortion present, the signal-to-noise ratio possible with each method, and the amount of wow or flutter that may be expected under average conditions of reproduction for each recording process.

Transistor Shift Register and Serial Adder. J. R. HARRIS[1]. *I.R.E., Proc.*, **40**, pp. 1597–1602, Nov., 1952.

A small set of basic functions, such as binary memory and elementary binary logic, can be remarkably versatile; such functions are important in switching and computing. This paper describes a piece of computing equipment which can store a pair of binary numbers and add them, producing the sum a digit at a time. The equipment is built from a basic set of functional blocks, all of which are designed around transistors. This set of building blocks consists of a binary cell, a pulse amplifier, a pulse amplifier with delay, and logic circuits. The binary cell is a flip-flop; amplifiers are monostable circuits, and logic is performed in diode gates. Some interesting special features arise from the use of transistors. These features are discussed and the designs are evaluated.

Charge Transfer and the Mobility of Rare Gas Ions. J. A. HORNBECK[1]. *Jl. Phys. Chem.*, **56**, pp. 829–831, Oct., 1952.

Ion-atom collisions in the rare gases between an atomic ion and a parent gas atom, such as Ne^+ and Ne, involve quantum mechanical symmetry effects which though rigorously inseparable have been listed as (a) a force of resonance attraction, (b) a force of resonance repulsion, and (c) charge exchange. Drift velocity measurements at high fields show that this complicated interaction may be represented to a good approximation by the hard sphere model of kinetic theory in which the collision cross section is several times the viscosity cross section of the atoms themselves.

Broad Band Matching with a Directional Coupler. W. C. JAKES[1]. *I.R.E., Proc.*, **40**, pp. 1216–1218, Oct., 1952. (Monograph 2033).

---

[1] Bell Telephone Laboratories.
[4] Westrex Corporation.

This paper presents the results of a theoretical and experimental study of a waveguide matching technique which allows a directional coupler to be located any distance away from the discontinuity causing the original mismatch and a broad-band match to still be obtained.

Design curves are included which give the required coupling coefficient of the directional coupler and the power loss for a given initial mismatch and desired vswr reduction. Experimental confirmation of the theory is also presented.

New General-Purpose Relay for Telephone Switching Systems. A. C. KELLER[1]. *Elec. Engg.*, **71,** pp. 1007–1012, Nov., 1952. Monograph 2034).

This new general-purpose electromagnetic relay, called the AF type is a wire spring relay. With variations providing slow release or marginal characteristics, it is known as the AG and AJ relay, respectively. It provides improved performance at lower cost.

Spherical Model of a Ferromagnet. H. W. LEWIS[1] and G. H. WANNIER[1]. Letter to the Editor. *Phys. Rev.*, **88,** pp. 682–683, Nov. 1, 1952.

In order to interpret the properties of the tetragonal crystal $ND_4D_2PO_4$ (deuterated ADP) a thermo-dynamic treatment has been developed which relates the observed crystal structure change and the dielectric constant change at the transition temperature to the appearance of spontaneous polarization. For an antiferroelectric crystal, the average spontaneous polarization is zero, being oppositely directed for adjacent layers, but the square of the spontaneous polarization is large. This results in quadratic strain components which cause a change in the crystal structure below the transition temperature. It is shown that the change observed is consistent with an antiferroelectric arrangement with one of the α axes being the antiferroelectric axis. The dielectric constants in all three directions suffer a large drop below the transition temperature.

Piezoelectric, Dielectric, and Elastic Properties of $ND_4D_2PO_4$ (Deuterated ADP). W. P. MASON[1] and B. T. MATTHIAS[1]. *Phys. Rev.*, **88,** pp. 477–479, Nov. 1, 1952. (Monograph 2036).

Transistors in Our Civilian Economy. J. W. MCRAE[1]. *I.R.E., Proc.*, **40,** pp. 1285–1286, Nov., 1952.

I. R. E. Editor's Note: At relatively long intervals there appear on the technical and industrial horizons devices of such broad scope and major significance that they profoundly affect the fields of their use. One of these epochal developments is the transistor, which bids fair to take its place beside the electron tube as one of the foundation stones of future communications and electronics.

It is accordingly timely and suitable that certain of the probable future industrial uses and effects of the transistor should be here analyzed in a guest editorial by an engineer especially qualified for this task, and who is a Fellow and Director of the Institute, and a Vice President of Bell Telephone Laboratories.

[1] Bell Telephone Laboratories.

Domain Properties in $BaTiO_3$. W. J. MERZ[1]. Letter to the Editor. *Phys. Rev.*, **88,** pp. 421–422, Oct. 15, 1952.

Notes on Methods of Transmitting the Circular Electric Wave Around Bends. E. S. MILLER[1]. *I.R.E., Proc.*, **40,** pp. 1104–1113, Sept., 1952. (Monograph 2037).

The tendency for energy to be converted out of the circular electric wave in bent round pipe may be avoided by one of three general approaches: (1) by removing the degeneracy between TE_{01} and TM_{11}, (2) by converting to a normal mode of the bent guide at both ends of the bend, and (3) by utilizing dissipation in the unwanted modes to prevent power transfer to them. All three approaches are discussed. Normal attenuation in round pipe should be effective in moderating straightness requirements. Elliptical guide and special waveguide structures may be used to negotiate intentional bends; bending radii in the range one to 1,000 feet appear acceptable at 50,000 mc for waveguides ⅜-inch to 2 inches in diameter, respectively.

Multi-Element Directional Couplers. S. E. MILLER[1] and W. W. MUMFORD[1]. *I.R.E., Proc.*, **40,** pp. 1071–1078, Sept., 1952. (Monograph 2038).

It is shown that the backward wave in a directional coupler is related to the shape of the function describing the coupling between transmission lines by the Fourier transform. This facilitates the design of directional couplers for arbitrary directivities over any prescribed frequency band. Tightly coupled directional couplers are analyzed in simple terms, and it is shown that any desired loss ratio, including complete power transfer between lines, may be achieved. The theories are verified using waveguide models operating at 4,000, 24,000 and 48,000 mc, and it is indicated that the work is applicable to many types of electrical and acoustic transmission lines.

Transistor Noise in Circuit Applications. H. C. MONTGOMERY[1]. *I.R.E., Proc.*, **40,** pp. 1461–1471, Nov., 1952.

Linear circuit problems involving multiple noise sources can be handled by familiar methods with the aid of certain noise spectrum functions, which are described. Several theorems of general interest dealing with noise spectra and noise correlation are derived. The noise behavior of transistors can be described by giving the spectrum functions for simple but arbitrary configurations of equivalent noise generators. From these, the noise figure can be calculated for any desired external circuit.

Transistor Noise in Circuit Applications. H. C. MONTGOMERY[1]. *I.R.E., Proc.*, **40,** pp. 1314–1326, Nov., 1952.

The invention of the transistor provided a simple, apparently rugged device that could amplify — an ability which the vacuum tube had long monopolized. As with most new electron devices, however, a number of extremely practical limitations had to be overcome before the transistor could be regarded as a

[1] Bell Telephone Laboratories.

practical circuit element. In particular, the reproducibility of units was poor — units intended to be alike were not interchangeable in circuits; the reliability was poor — in an uncomfortably large fraction of units made, the characteristics changed suddenly and inexplicably; and the "designability" was poor — it was difficult to make devices to the wide range of desirable characteristics needed in modern communications functions. This paper describes the progress that has been made in reducing these limitations and extending the range of performance and usefulness of transistors in communications systems. The conclusion is drawn that for some system functions, particularly those requiring extreme miniaturization in space and power as well as reliability with respect to life and ruggedness, transistors promise important advantages.

In Search of the Missing 6 *Db.* W. A. MUNSON[1] and F. M. WIENER[1]. *Acoustical Soc. Am., Jl.,* **24,** pp. 498–501, Sept., 1952. (Monograph 2019).

The unexplained difference in sound pressure in the ear canal which appears to exist when equally loud low frequency tones are presented alternately from an earphone and from a loudspeaker has bedeviled acousticians for many years and, unfortunately, still continues to do so. There are presented here the results of some of the measurements carried out at the Bell Telephone Laboratories which show the magnitude of the effect and various attempts at explaining it. While no satisfactory explanation has been found, it is hoped that publication of these results will stimulate interest in the problem.

Nation-Wide Numbering Plan. W. H. NUNN[2]. *A.I.E.E., Trans., Commun. and Electronics Sect.,* **2,** pp. 257–260, Sept., 1952. *Elec. Engg.,* **71,** pp. 884–888, Oct., 1952. (Monograph 2015).

At the present time a great variation in the types of telephone numbers exists. This is because of the number of telephones in communities of different sizes. With the advent of local dialing and now nation-wide dialing, a uniform numbering system has become necessary.

How to Detect the Type of an Assignable Cause. P. S. OLMSTEAD[1]. *Ind. Quality Control,* **9,** pp. 32–34, **36,** Nov., 1952.

Silicon p-n *Junction Alloy Diodes.* G. L. PEARSON[1] and B. SAWYER[1]. *I.R.E., Proc.,* **40,** pp. 1348–1351, Nov., 1952.

A new type of p-n junction silicon diode has been prepared by alloying acceptor or donor impurities with n- or p-type silicon. The unique features of this diode are: (a) reverse currents as low as 10^{-10} amperes, (b) rectification ratios as high as 10^8 at 1 volt, (c) a Zener characteristic in which $d(\log I)\ d(\log V)$ may be as high as 1,500 over several decades of current, (d) a stable Zener voltage

[1] Bell Telephone Laboratories.
[2] American Telephone and Telegraph Company.

which may be fixed in the production process at values between 3 and 1,000 volts and (e) ability to operate at ambient temperatures as high as 300°C.

Hard Rubber. H. Peters[1]. *Ind. and Eng. Chem.*, **44,** pp. 2344–2345, Oct., 1952.

As judged by the literature, the general trend during the past year on the subject of hard rubber has been toward de-emphasis of fundamental research and more emphasis on use. Plastics, through substitution, continue to make gains in the field of hard rubber. A renewed interest is again shown in the use of latex ebonite for industrial applications. The patent situation appears to be unusually active and the interest in synthetic hard rubbers continues to increase.

Application of Information Theory to Research in Experimental Phonetics. G. E. Peterson[1]. *Jl. Speech and Hearing Disorders*, **17,** pp. 175–188, June, 1952.

Principles of Zone-Melting. W. G. Pfann[1]. *Jl. of Metals*, **4,** pp. 747–753, July, 1952. *A.I.M.E. Trans.*, **194,** pp. 747–753, 1952. (Monograph 2000).

In zone-melting, a small molten zone of zones traverse a long charge of alloy or impure metal. Consequences of this manner of freezing are examined with respect to solute distribution in the ingot, with particular reference to purification and to prevention of segregation. Results are expressed in terms of the number, size, and direction of travel of the zones, the initial solute distribution, and the distribution coefficient.

Nonsynchronous Time Division with Holding and with Random Sampling. J. R. Pierce[1] and A. L. Hopper[1]. *I.R.E., Proc.*, **40,** pp. 1079–1088, Sept., 1952. (Monograph 2041).

There is a general type of system in which an indefinitely large number of transmitters can have access to any of an indefinitely large number of receivers over a medium of limited bandwidth. In these systems, signal-to-noise ratio goes down as more transmitters are used simultaneously. This paper describes a particular system which sends samples by means of coded pulse groups sent at random times. The signal-to-noise ratio is good in the absence of interference and the effect of interference is minimized by holding the previous sample if a sample is lost. An experimental system worked satisfactorily and gave close to the predicted signal-to-noise ratio. Such a system might be used to provide communication and automatic switching in rural telephony, or for other applications.

Fundamental Plans for Toll Telephone Plant. J. J. Pilliod[2]. *A.I.E.E., Trans., Commun. & Electronics Sect.*, **2,** pp. 248–256, Sept., 1952. (Monograph 2015).

---

[1] Bell Telephone Laboratories.

[2] American Telephone and Telegraph Company.

Organization of the Engineering Profession. D. A. QUARLES[5]. *Elec. Engg.*, **71,** pp. 963, 964, Nov., 1952.

Since the organization of the American Society of Civil Engineers 100 years ago, professional engineering has assumed a major role in American life. The goal now to be attained is the closer organization of the entire engineering profession.

We begin a New Institute Year. D. A. QUARLES[5]. *Elec. Eng.*, **71,** pp. 867–868, Oct., 1952.

In an address before the recent Pacific General Meeting in Phoenix, Mr. Quarles, President of the Institute, evaluates the evolution in A.I.E.E. organization and policy as the Institute enters a new administrative year.

Mean Free Paths of Electrons in Evaporated Metal Films. F. W. REYNOLDS[1] and G. R. STILWELL[1]. Letter to the Editor. *Phys. Rev.*, **88,** pp. 418–419, Oct. 15, 1952.

Single-Sideband System for Overseas Telephony. N. F. SCHLAAK[1]. *Electronics*, **25,** pp. 146–149, Nov., 1952.

Single-sideband transmitter furnishes four voice channels for overseas telephone service. Pushbutton tuning permits rapid frequency shifts and load-control circuit minimizes interchannel crosstalk and out-of-band radiation. Copper-oxide and germanium varistors replace modulator tubes.

Automatic Toll Switching Systems. F. F. SHIPLEY[1]. *A.I.E.E., Trans., Commun. and Electronics Sect.*, **2,** pp. 261–269, pp. 889–897, Oct., 1952. (Monograph 2015).

The new system was designed to implement the nation-wide switching plan which integrates the telephone switching network of the entire nation into a single unit. Requiring a high order of mechanical intelligence, this system is one of the most comprehensive ever devised.

Properties of M-1740 p-n *Junction Photocells.* J. N. SHIVE[1]. *I.R.E., Proc.*, **40,** pp. 1410–1413, Nov., 1952.

The *p-n* junction photocell has a sensitivity of 30 ma per lumen for light of 2,400 degrees K color temperature, corresponding to a quantum yield approximately unity in the spectral range from visible to the long wave cutoff at 1.8 microns. Dark currents of a few microamperes are observed at room temperature, with a temperature coefficient of about +10 per cent per degree C. Both dark and light currents exhibit saturation in the range from 1 to 90 volts applied. The frequency response is flat into the 100-kc region. Short-circuit noise currents are observed around 20 $\mu\mu$a in a 1-cps band at 1,000 cps. The photocell element is encapsulated in a plastic housing $\frac{1}{4} \times \frac{3}{16} \times \frac{3}{8}$ inch in dimensions.

[1] Bell Telephone Laboratories.
[5] Sandia Corporation.

Interpretation of e/m Values for Electrons in Crystals. W. Shockley[1]. Letter to the Editor. *Phys. Rev.*, **88,** p. 953, Nov. 15, 1952.

Transistor Electronics: Imperfections, Unipolar and Analog Transistors. W. Shockley[1]. *I.R.E., Proc.*, **40,** pp. 1289–1313, Nov., 1952.

The electronic mechanisms that are of chief interest in transistor electronics are discussed from the point of view of solid-state physics. The important concepts of holes, electrons, donors, acceptors, and deathnium (recombination center for holes and electrons) are treated from a unified viewpoint as imperfections in a nearly perfect crystal. The behavior of an excess electron as a negative particle moving with random thermal motion and drifting in an electric field is described in detail. A hole is similar to an electron in all regards save sign of charge. Some fundamental experiments have been performed with transistor techniques and exhibit clearly the behavior of holes and electrons. The interactions of holes, electrons, donors, acceptors, and deathnium give rise to the properties of *p-n* junctions, *p-n* junction transistors, and Zener diodes. Point-contact transistors are not understood as well from a fundamental viewpoint. A new class of *unipolar* transistors is discussed. Of these, the *analog* transistor is described in terms of analogy to a vacuum tube.

Unipolar "Field-Effect" Transistor. W. Shockley[1]. *I.R.E., Proc.*, **40,** pp. 1365–1376, Nov., 1952.

The theory for a new form of transistor is presented. This transistor is of the "field-effect" type in which the conductivity of a layer of semiconductor is modulated by a transverse electric field.

Since the amplifying action involves currents carried predominantly by one kind of carrier, the name "unipolar" is proposed to distinguish these transistors from point-contact and junction types, which are "bipolar" in this sense.

Regarded as an analog for vacuum-tube triode, the unipolar field-effect transsistor may have a mμ of 10 or more, high output resistance, and a frequency response higher than bipolar transistors of comparable dimensions.

Control of Frequency Response and Stability of Point-Contact Transistors. B. N. Slade[1]. *I.R.E., Proc.*, **40,** pp. 1382–1384, Nov., 1952.

The frequency response and stability of point-contact transistors are determined to a large degree by control of the point-contact spacing and germanium resistivity. Stability is particularly important in amplifiers in which the impedances of the emitter and collector circuits are very small in the frequency range in which the transistor is designed to operate. Satisfactory stability has been obtained with developmental transistors having a frequency cutoff (3-db drop in the current amplification factor, alpha) ranging from 10 to 30 mc. These transistors operate under approximately the same dc bias conditions used with lower-frequency transistors, and have an average power gain of approximately 20 db. By means of the methods outlined, transistors which oscillate at frequencies as high as 300 mc have been made.

[1] Bell Telephone Laboratories.

Junction Transistor. M. SPARKS[1]. *Sci. Am.*, **187,** pp. 29–32, July, 1952.
It is one of two forms of the remarkable device that amplifies electricity by the flow of electrons in a crystal. An account of its underlying principles and present state of development.

Telephone Answering Services. L. R. STANG[6]. *Telephony*, **143,** pp. 53–55, 123–124, Oct. 25, 1952.

Traffic Engineering Design of Dial Telephone Exchanges. J. A. STEWART[1]. *Telephony*, **143,** pp. 13–16, 41–42, Oct. 18, 1952.

Low-Drain Transistor Audio Oscillator. D. E. THOMAS[1]. *I.R.E., Proc.*, **40,** pp. 1385–1395, Nov., 1952.
A nine-element transistor audio oscillator is described. This oscillator operates with relatively low drain from a single 6-volt battery. The oscillator gives reliable performance with an output uniform to approximately ± 1 db with substantially all type 1768 point-contact transistors and without any circuit element adjustment required for variation in transistor parameters from unit to unit or with transistor ambient temperature.

Transistor Amplifier — Cutoff Frequency. D. E. THOMAS[1]. *I.R.E., Proc.*, **40,** pp. 1481–1483, Nov., 1952.
The effect of positive feedback through the internal base resistance of a transistor on circuit cutoff frequency is considered.

Transistor Reversible Binary Counter. R. L. TRENT[1]. *I.R.E., Proc.*, **40,** pp. 1562–1572, Nov., 1952.
The feasibility of performing a fairly complex switching function using a few elementary transistor circuits is illustrated and experimentally verified. The specific function discussed is reversible vinary counting. The mechanism used to achieve reversibility and the circuitry within each building block is described. Operating margins and suggestions for design improvements for systems application are given.

Effect of Electrode Spacing on the Equivalent Base Resistance of Point-Contact Transistors. L. B. VALDES[1]. *I.R.E., Proc.*, **40,** pp. 1429–1434, Nov., 1952.
A theoretical expression for the equivalent base resistance r_b of point-contact transistors is derived here. This expression is shown to check experimental values reasonably well if the severity of some assumptions made for purposes of analysis is considered. Electrode spacing, germanium-slice thickness, and resistivity of the semiconductor are shown to be the properties that affect r_b primarily.

[1] Bell Telephone Laboratories.
[6] Illinois Bell Telephone Company.

Measurement of Minority Carrier Lifetime in Germanium. L. B. VALDES[1]. *I.R.E., Proc.*, **40,** pp. 1420–1423, Nov., 1952.

A method for measuring the lifetime of minority carriers in germanium is described. Basically, it consists of liberating the carriers optically on a flat face of a crystal and measuring the concentration of minority carriers as a function of distance from the point of liberation. The mathematical model is analyzed and experimental results are presented here.

Drift Velocities of Ions in Krypton and Xenon. R. N. VARNEY[1]. *Phys. Rev.*, **88,** pp. 362–364, Oct. 15, 1952. (Monograph 2028).

Drift velocities and mobilities of ions of Kr and Xe in their respective parent gases have been measured over a wide range of values of E/p_0, the ratio of electric field strength to normalized gas pressure. Two ions appear in each gas identified as Kr^+ and Kr_2^+ in Kr and Xe^+ and Xe_2 in Xe. The relation that drift velocity varies as $(E/p_0)^{\frac{1}{2}}$ at high E/p_0 has been found to hold for the atomic ions and has been used to determine the equivalent hard sphere cross sections at high fields. The cross sections are 157×10^{-16} cm^2 for Kr and 192×10^{-16} cm^2 for Xe. The Langevin theory of mobilities gives excellent agreement with experimental results extrapolated to zero field strength provided that, in the theory, the hard sphere cross section is taken as large for the atomic ions and very small for the molecular ions. The range of the polarization forces is such as to render them insignificant in atomic ion collisons and of primary importance in molecular ion collisions.

Junction Transistor Tetrode for High-Frequency Use. R. L. WALLACE[1], L. G. SCHIMPF[1] and E. DICKTEN[1]. *I.R.E., Proc.*, **40,** pp. 1395–1400, Nov., 1952.

If a fourth electrode is added to a conventional junction transistor and biased in a suitable way, the base resistance of the transistor is reduced by a very substantial factor. This reduction in r_b permits the transistor to be used at frequencies ten times or more higher than would otherwise be possible. Tetrodes of this sort have been used in sine-wave oscillators up to a frequency of 130 mc and have produced substantial gain as tuned amplifiers at frequencies of 50 mc and higher.

Nature of Solids. G. H. WANNIER[1]. *Sci. Am.*, **187,** p. 39 Dec., 1952.
The theory that explains their various properties is a comparatively recent development of physics. From it practical benefits already begin to flow.

Magnetic Double Refraction at Microwave Frequencies. M. T. WEISS[1] and A. G. FOX[1]. Letter to the Editor, *Phys. Rev.*, **88,** pp. 146–147, Oct. 1, 1952.

Stress Relaxation in Plastics and Insulating Materials. E. E. WRIGHT[1]. *A.S.T.M. Bull.*, **184,** pp. 47–49, Sept., 1952. (Monograph 2024).

[1] Bell Telephone Laboratories.

Organic materials are in ever-increasing use for mechanical and electrical devices where satisfactory performance is required over long periods of time and under a wide assortment of atmospheric influences. Although the dimensional stability of plastics and electrical insulating materials under no-load conditions has been established reasonably well, there is an important gap in existing knowledge with regard to a material's ability to maintain adequate counter-stresses under compressive loading.

A.S.T.M. Method of Test D 621 – 51[2] uses a constant load system and measures the material's resistance to gross deformation. However, this fails to simulate the usual application where the material is subjected to constant deflection (such as fastening devices, inserts, etc.) and is required to maintain adequate counter-stresses for the foreseeable life of the part. Therefore, it has been necessary to integrate data from Method D 621 with long practical experience in order to extrapolate between two dissimilar systems.

This paper describes a constant-deflection procedure for direct measurement of stress relaxation or change thereof, thereby permitting evaluation in terms of a material's ability to maintain a tight assembly under conditions simulating actual use. Apparatus for carrying out the test is described and typical data illustrating its usefulness are included.

Contributors to this Issue

Charles Clos, C.E., New York University, 1927; New York Telephone Company, plant extension engineering, valuation and depreciation matters, intercompany settlements and tandem and toll fundamental plans, 1927–47. Pratt Institute, Evening School, Mathematics Instructor, 1946–49. Bell Telephone Laboratories, studies on development planning for local and toll switching systems and research in switching probability, 1947–. Member of A.I.E.E., New York Electrical Society, Mathematical Association of America, A.A.A.S., American Statistical Association, Iota Alpha, and Tau Beta Pi.

R. H. Colley, A.B., Dartmouth College, 1909; A.M., Harvard University, 1912; Ph.D., George Washington University, 1918; Austin Teaching Fellow in Botany, Harvard University, 1910–12; Instructor in Botany, Dartmouth College, 1909–10 and 1912–16; Pathologist, Division of Forest Pathology, Bureau of Plant Industry, U. S. Department of Agriculture, 1916–28. Bell Telephone Laboratories, 1928–52. Dr. Colley was chairman of Committee 05 — Wood Poles, of the American Standards Association for nearly twenty years. He was president of the American Wood-Preservers' Association 1943–44. During his years with the Laboratories he worked particularly on development and research problems connected with material and preservative treatment specifications for poles and other timber products used in outside plant. His more recent activities were directed toward improvement of laboratory techniques for evaluating wood preservatives, and toward the development of a coordinated plan for fundamental research on oil preservatives. He was Timber Products Engineer for the Laboratories from 1940 to 1950, and Timber Products Consultant from 1950 to 1952. His article in this issue of the Journal was prepared before his retirement on May 31, 1952.

Karl K. Darrow, B.S., University of Chicago, 1911. He studied at the Universities of Paris and Berlin in 1911 and 1912, specializing in physics and mathematics; Ph.D., University of Chicago, 1917. He then joined the staff of Bell Telephone Laboratories, at that time known as the Engineering Department of Western Electric Company. Here his work has included the study, correlation, and representation of scientific information for his colleagues, keeping them informed of current advances made by workers in fields related to their own activities. As a corollary to his work, Dr. Darrow appears from time to time before scientific and lay audiences to lecture on current topics in physics and the related sciences. He has taken an active interest in education, teaching physics during summer and other sessions at Stanford, Chicago, and Columbia Universities and at Smith College. From 1944 to 1946, he served as consultant to the Metallurgical Laboratory in Chicago. Dr. Darrow is the author of *Introduction to Contemporary Physics* (1939), *Electrical Phenomena in Gases* (1932), *Renaissance of Physics* (1936), *Atomic Energy* (1948). He is a member of the American Physical Society, which he has served as secretary since 1941, the Physical Society of London, Société Francaise de Physique, the American Philosophical Society, of which he was a counsellor for four years, and the International Union of Pure and Applied Physics, of which he was vice president from 1947 to 1951. In 1949 he received an honorary doctorate from the Université de Lyon.

George R. Frost, Bell Telephone Laboratories, 1943–. Mr. Frost has had over twenty-three years of service with the Bell System. Since 1943 he has been an instructor at the Laboratories' School for War Training and in the Communications Development Training Program, until recently when he became a member of the Publication Department's group responsible for displays. From 1941–43 he taught communications at Fenn College, and from 1946–47 mathematics at Pratt Institute.

William Keister, B.S. in E.E., Alabama Polytechnic Institute, 1930; Bell Telephone Laboratories, 1930–32 and 1936–. As a member of the switching department, he is currently preparing text material and teaching general switching circuit theory and telephone switching systems to members of the technical staff. Co-author of *The Design of Switching Circuits*, with S. H. Washburn and A. E. Ritchie (Van Nostrand, 1951). Member of A.I.E.E., Eta Kappa Nu, Tau Beta Pi, and Phi Kappa Phi.

Alistair E. Ritchie, A.B., Dartmouth College, 1935; M.A., Dartmouth College, 1937. Bell Telephone Laboratories, 1937–. As a member of the switching development group, Mr. Ritchie tested panel and crossbar circuits and made noise studies on panel and crossbar systems until

1942. He then became an instructor in the Laboratories' School for War Training. From 1945–51, he taught switching circuit design. Since 1951, in the switching engineering group, he has been working out new techniques for measuring telephone traffic in central offices. Co-author of *The Design of Switching Circuits* with W. Keister and S. H. Washburn (Van Nostrand, 1951). Member of the A.I.E.E.

H. A. Stone, Jr., B.S., Yale University, 1933. Bell Telephone Laboratories, 1936–. A member of the Transmission Development Department, Mr. Stone is in charge of a group engaged in the development of inductors and loading coils and cases. He previously assisted in the development of inductors and networks for use in military radio and telephone projects, and the design of radar pulse generators. Member of the A.I.E.E.

Roger I. Wilkinson, B.S. in E.E., 1924, Professional Engineer (honorary), 1950, Iowa State College; Northwestern Bell Telephone Company, 1920–21; American Telephone and Telegraph Company, 1924–34; Bell Telephone Laboratories, 1934–43 and 1946–. U. S. War Department, Washington and South Pacific, 1943–45. Mr. Wilkinson has been engaged in applications of the mathematical theory of probability to telephone problems. Medal for Merit, 1946. Member of A.S.E.E.; A.S.A.; Institute of Mathematical Statistics; American Society for Quality Control; Fellow, Operations Research Society of America; Associate Member of A.I.E.E.; and Member of Eta Kappa Nu; Tau Beta Pi; Phi Kappa Phi; and Pi Mu Epsilon.