SUPPLEMENT

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Contents

QUESTIONS AND ANSWERS

Answers are occasionally omitted or reference is made to earlier Supplements in which questions of substantially the same form, together with the answers, have been published. Some answers contain more detail than would be expected from candidates under examination conditions.

For economic reasons, alternate issues of the Supplement are published in 32-page and 16-page sizes.

TELEPHONY AND TELEGRAPHY A, 1976

Students were expected to answer any 6 questions

Q 1 (*a*) *With the aid of simple sketches, explain how a telephone dial mechanism*

(*i*) *registers the dialled digit during the forward motion, and*
(*ii*) *transmits the registered digit during the return motion.*

(*b*) *Sketch simple circuits to show how the digits are sent to line from*

(*i*) *a telephone station, and*
(*ii*) *a Telex station.*

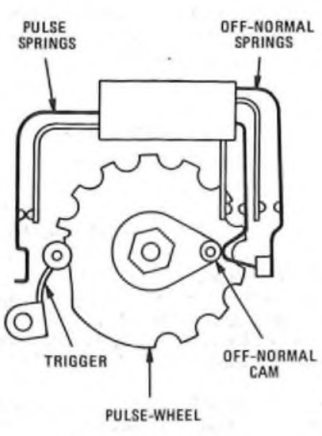
A 1 (*a*) (*i*) The basic mechanism of a trigger dial in its normal position is shown in sketch (*a*). The pulsing contacts are closed; the off-normal contacts are open. When the dial is rotated off-normal, the movement of the off-normal cam allows the off-normal contacts to close, while the leading edge of the first tooth on the pulse wheel pulls the trigger clear of the pulse springs against the tension of a restoring spring. This position, when the dial has been rotated a short distance, is illustrated in sketch (*b*).

Further rotation of the dial in the direction shown in sketch (*b*) causes the dial to register the dialled digit by means of the number of pulse-wheel teeth which pass under the trigger in its fully extended position. In the extended position, the trigger does not make contact with the pulse springs, and no digits can be transmitted.

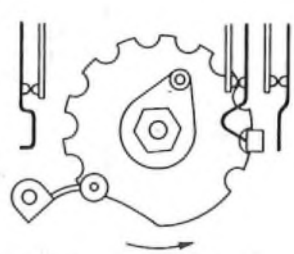
(*ii*) When the dial is released, it rotates in the opposite direction, as shown in sketch (*c*), under the influence of its restoring spring. During rotation in this direction, the leading face of the first pulse-wheel tooth (that is, the last tooth to have passed under the trigger) restores the trigger to its normal position under the pulse springs. The time taken for the trigger to restore ensures that the dial produces a minimum inter-digital pause of 240 ms, and also ensures that, before pulsing commences, the dial has accelerated to its full governed speed.

When the trigger reaches its normal position, further rotation of the pulse wheel causes the trigger to ride over each of the teeth. Now, however, as the trigger rides over each tooth, it lifts the moving spring of the pulsing contacts and so sends a disconnexion to line. As the valley between each tooth passes under the trigger, the pulsing contacts remake to send a loop to line. The relative widths of teeth and valleys are such as to produce a *make* : *break* loop-disconnect pulse ratio of 1 : 2.

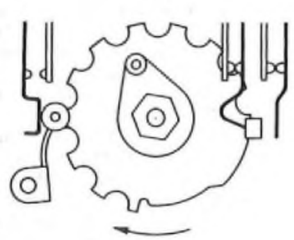
(*b*) (*i*) and (*ii*) The essential pulsing elements of a telephone circuit are shown in sketch (*d*); a simplified Telex pulsing circuit is shown in sketch (*e*).



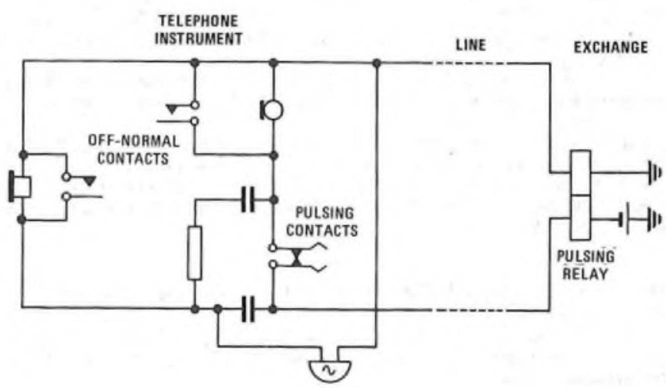
(*a*)



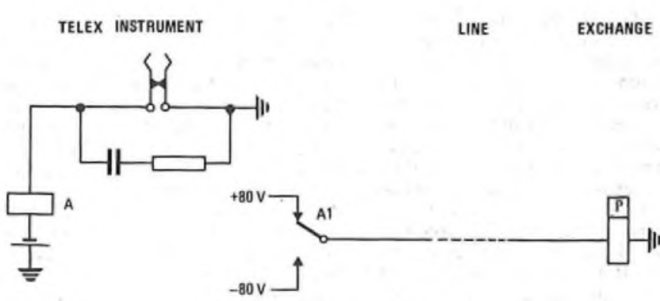
(*b*)



(*c*)



(*d*)



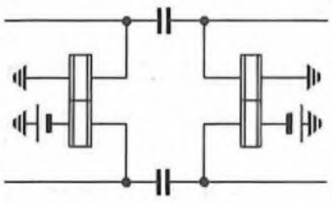
(*e*)

Q 2 (a) Draw circuits for transmission bridges of
(i) the capacitor type, and
(ii) the transformer type.

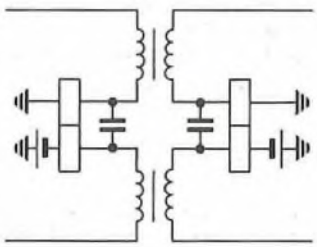
(b) Explain why each has a different effect on the speech frequencies transmitted across the transmission bridge.

(c) Which type of transmission bridge is normally used in a final selector, and what is its principal function?

A 2 (*a*) (*i*) and (*ii*) Capacitor and transformer-type transmission bridges are shown in sketches (*a*) and (*b*) respectively.

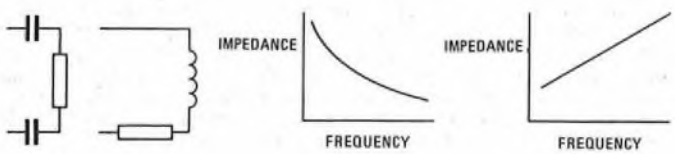


(*a*)



(*b*)

(*b*) Sketches (*c*) and (*d*) respectively show simplified equivalent circuits for the capacitor and transformer-type bridges at speech frequencies. It can be seen from sketch (*c*) that the only reactive component in the former is capacitance. The impedance of a capacitor is inversely proportional to the frequency, so that the impedance (and hence the loss) decreases as the frequency increases. A graph of the frequency characteristic is shown in sketch (*e*).



(*c*) (*d*) (*e*) (*f*)

For the transformer bridge, the simplified equivalent circuit of sketch (*d*) shows that the reactive component is inductance. For an inductor, the impedance (and hence the loss) is directly proportional to frequency. The frequency characteristic is shown in sketch (*f*), and can be seen to give an effect on the speech frequencies opposite to that of the capacitor bridge.

In practice, the effects of shunting relays, line constants and, in the transformer-bridge case, DC-blocking capacitors, have also to be taken into account.

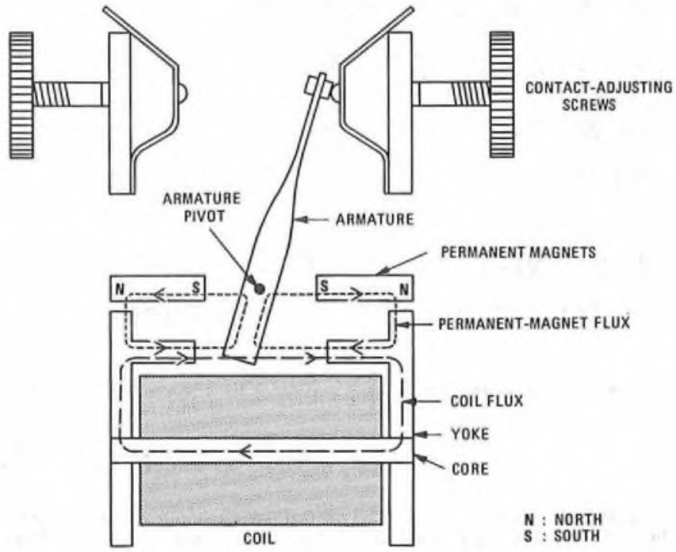
(*c*) A capacitor-type transmission bridge is normally provided in a final selector. The principal function of a transmission bridge is to provide separate feeds of DC transmitter-polarizing current to each side of the bridge, while allowing AC speech currents to pass across the bridge.

Q 3 (a) With the aid of a sketch, explain the action of the magnetic circuit of a telegraph-type polarized relay.

(b) State and explain the reasons for 2 advantages that are gained by arranging the magnetic circuit so that the direction of magnetic flux in the armature does not change.

A 3 (*a*) The magnetic circuit of a telegraph-type polarized relay is shown in the sketch. It consists of a coil, a soft-iron core-and-yoke assembly, an armature and 2 permanent magnets. The permanent magnets produce fluxes as shown, and these are sufficient to hold the armature on either of the 2 contacts. The coil flux is arranged to assist one permanent-magnet flux while opposing, and thus weakening, the other. Thus, if the armature is resting on the MARK contact, the application of *marking* current to the coil has no effect other than to hold the armature more firmly in position. The application of *spacing* current, however, reverses the coil flux and alters the balance of the 3 fluxes in the armature. The force holding the armature on the MARK contact is reduced, the force of attraction to the SPACE contact is increased, and the armature moves across to the SPACE contact.

(*b*) The form of construction described provides a balanced movement having equal operate times in each direction. Two advantages gained by having a unidirectional flux are that



(*i*) a high speed of operation is ensured since there is a smaller inherent time lag than would be the case if the flux decayed to zero and built up in the opposite direction, and

(*ii*) hysteresis losses, and hence eddy currents, are low; this permits the use of a light-weight armature, which minimizes inertia and thus permits high-speed operation. The sensitivity is also improved since low hysteresis losses mean low power wastage, and the low armature inertia also aids the sensitivity.

Q 4 An AC signalling system for use on junctions between exchanges may be an in-band type or an out-of-band type.

(a) With the aid of a block diagram, give an outline description of the operation of one of the systems.

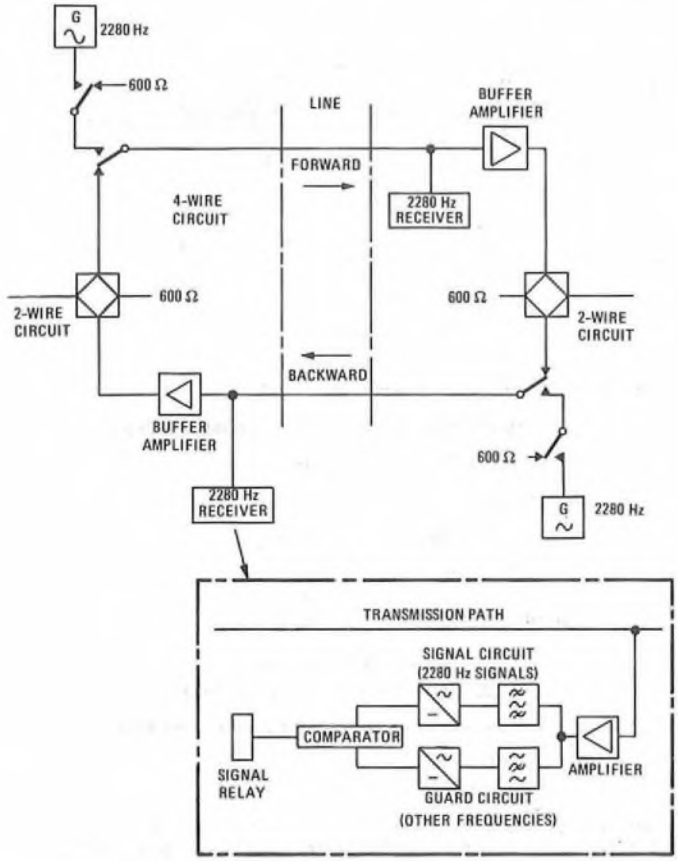
(b) State one advantage and one disadvantage of the out-of-band signalling system.

A 4 (*a*) A block diagram of an in-band signalling system using a single signalling frequency of 2280 Hz is shown in the sketch. In this 4-wire system, signalling is carried out within the speech band; thus, signalling is possible only prior to the completion of the subscriber-to-subscriber speech path, and during the clearing-down of the call. The sequence of signals shown in the table is transmitted.

Direction (⟶ Forward, ⟵ Backward)	Signal	Incidence
⟶ ⟶ ⟵	Seizure Dial pulses Called-subscriber answer	Before establishment of subscriber-to-subscriber speech path
⟵ ⟶ ⟵	Called-subscriber clear Calling-subscriber clear Release-guard	During clear-down sequence

All signals are transmitted as different-length pulses of 2280 Hz. The signalling tone is injected into the 4-wire side of the terminating unit, and is detected by the receiver at the distant end. On detection of signalling tone, the receiver biases the buffer amplifier to its cut-off point, thus preventing the signalling tone reaching the 2-wire circuit. Recognition of the appropriate tone codes by the associated relay-set establishes the necessary DC circuit conditions on the 2-wire side of the terminating unit.

The receiver contains a number of safeguards against accidental operation by voice frequencies. The sketch includes a block diagram of the receiver. To operate the signal relay, signalling tone (at 2280 Hz) must be present at some minimum level while, at the same time, all other frequencies must not exceed a maximum value. Thus, if a high level of 2280 Hz signal were present in some speech signals, the signal relay would not operate because the other frequencies present would inhibit its operation via the guard circuit and comparator.



(*b*) One advantage gained by the use of an out-of-band system is that signalling is not confined to the setting-up and clearing-down periods of a call. Since the signalling frequency is outside the commercial speech band, it can be separated from speech by filters, and signals can be passed while speech is in progress without either subscriber being aware of them. This facility can be used for *metering-over-junction* signals, for instance.

One disadvantage of an out-of-band system is that it increases the bandwidth required. Present frequency-division-multiplex systems commonly have a channel spacing of 4 kHz. Thus, a 0·9 kHz band is available in excess of the commercial speech band of 0·3–3·4 kHz. Out-of-band systems can exploit this gap between channels at little extra cost, but the additional bandwidth required may well be a disadvantage when considering, say, a new digital system for which channel bandwidths have yet to be fixed.

Q 5 (*a*) *Draw a trunking diagram for a 4000-line non-director Strowger exchange.*

(*b*) *For each selector involved in an own-exchange call, briefly describe the part it plays in the process of selecting the required line.*

(*c*) *State 3 important functions performed by the final selector after it has stepped to the required line.*

A 5 (*a*) See A8, Telephony and Telegraphy A, 1974, and A4, Telephony and Telegraphy A, 1975, *Supplement*, Vol. 68, p. 20, Apr. 1975, and Vol. 69, p. 34, July 1976.

(*b*) See A7, Telephony and Telegraphy A, 1973, *Supplement*, Vol. 67, p. 31, July 1974.

(*c*) After stepping to the required line, the final selector

(*i*) tests the selected line,
(*ii*) applies ringing current to the called line, if it is free,
(*iii*) applies supervisory tones to the calling line,
(*iv*) initiates the metering sequence when the called-subscriber answers, and
(*v*) provides transmission-bridge facilities.

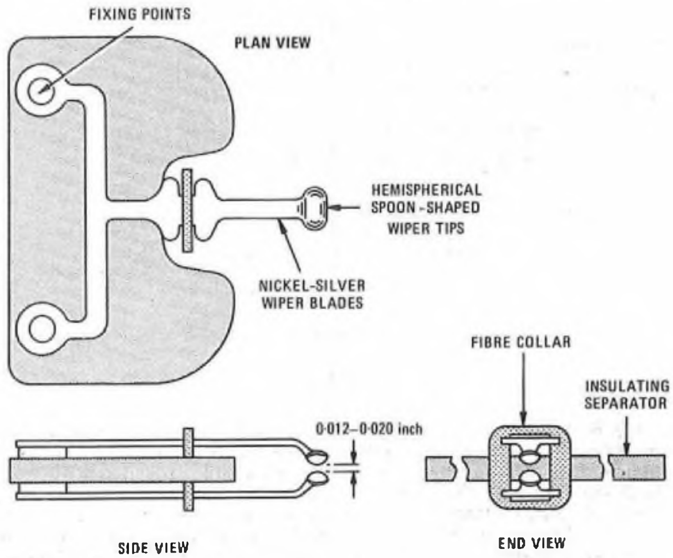
Q 6 (*a*) *With the aid of a sketch, describe the construction and adjustment of a wiper assembly for a 2-motion selector.*

(*b*) *Explain how the design of the wiper assists in maintaining a low-resistance connexion with the bank contacts.*

(*c*) *For what reason is a lubricant applied to the selector bank contacts, and what risks does this entail?*

A 6 (*a*) The sketch shows, in simplified form, the construction of a pair of wipers. They consist essentially of a pair of nickel-silver springs mounted on each side of an insulating separator. The ends of the springs are formed into spoon-shaped contact tips, and an insulating fibre collar is fitted over the blades, engaging in small lugs on the blades.

Wiper blades are adjusted by setting each blade outwards from the root before the collar is placed over them. When the collar is fitted, it holds the wipers together to form a tension-balanced pair, so that, for small deflexions, the gap between the contact tips is constant.



(*b*) The design features which assist in maintaining a low-resistance contact with the bank are listed below.

(*i*) The blade material (nickel-silver) maintains its tension for long periods.

(*ii*) The shape of the wiper tip gives a small contact area, and thus the contact pressure (force per unit area) is high.

(*iii*) The tension-balanced-pair nature of the wipers ensures that the tips exert equal pressure even if there is an overall deflexion in one direction.

(*iv*) The damping effect of the collar reduces contact bounce.

(*c*) Lubricant is applied to selector bank contacts to reduce friction between the banks and the wiper tips. The risk which this entails is that dust particles become trapped in the oil, and this mixture of oil and dust forms an abrasive paste which can lead to very rapid wear of wipers and, more seriously, banks.

Q 7 (*a*) *With the aid of a sketch of the circuit elements concerned, describe the circuit operation of a 100-outlet group selector during the action of hunting for and seizing a free outlet.*

(*b*) *Explain how mechanically-operated spring sets stop the hunting action when all outlets are busy.*

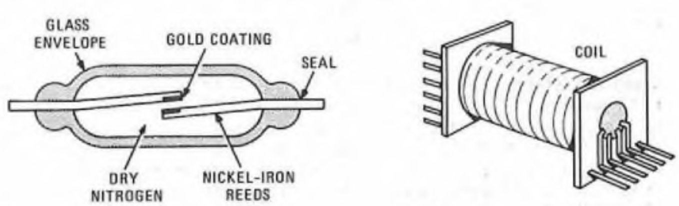
Q 8 *With the aid of sketches,*

(*a*) *describe the construction of a dry-reed insert,*

(*b*) *explain how groups of reed inserts are assembled to form a cross-point relay for a switching matrix, and*

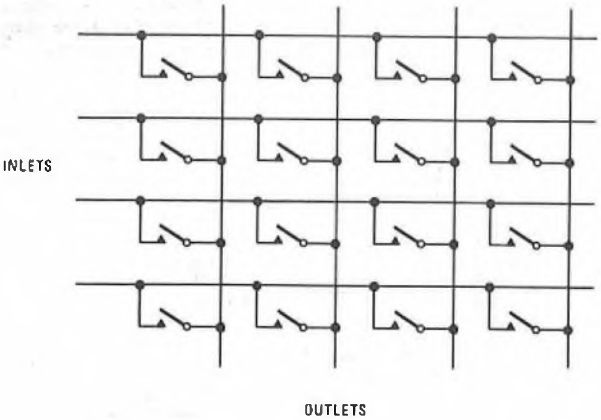
(*c*) *explain how a number of such relays can be arranged and wired to form a switching matrix.*

A 8 (*a*) The construction of a dry-reed insert is shown in sketch (*a*). The insert consists of 2 nickel-iron reeds sealed in a glass envelope containing an inert atmosphere of dry nitrogen. The contact areas of the reeds are coated with gold to reduce contact resistance.



(*a*) (*b*)

(*b*) A coil-and-former assembly for a crosspoint reed relay is shown in sketch (*b*). It consists of a plastic former with a coil wound on it. Four reed inserts are placed through the centre of the former, and connexions to the reeds are brought out to tags on the former cheeks, as are those to the coil. In use, the entire assembly is mounted in a metal can which provides screening and mechanical protection.



(*c*)

(*c*) A typical matrix arrangement for reed-relay switching is shown in sketch (*c*). An inlet is connected to an outlet by operating the appropriate reed. (In practice, several wires are associated with each inlet and outlet, so that one reed must be operated for each wire.) To facilitate construction of the matrix, the tags to which the reeds are connected are arranged to be in directions perpendicular to each other at opposite ends of the relay (sketch (*b*)). Thus, a matrix of reed relays can conveniently be wired by connecting horizontal commons (inlets) to one end of the reeds, and vertical commons (outlets) to the other.

Q 9 *(a) Sketch graphs to show typical variations in the originated telephone traffic at a local exchange*

(i) during a period of 24 h, and
(ii) within the busy hour.

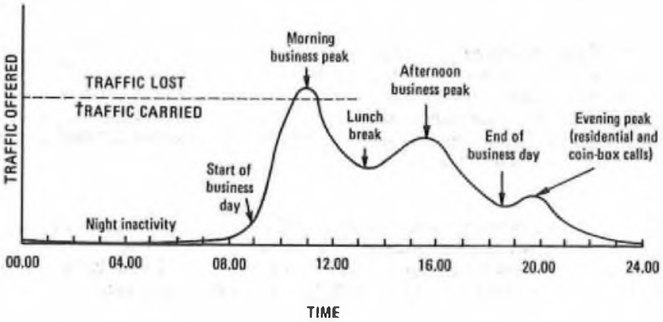
Indicate the reasons for the variations, and mark on your graph the traffic that is lost.

(b) A group of N circuits is offered 5 erlangs of traffic and gives a grade of service of 0·01. If the average duration of a call is 3 min,

(i) how much traffic is carried during the busy hour, and
(ii) how many calls are lost during the busy hour?

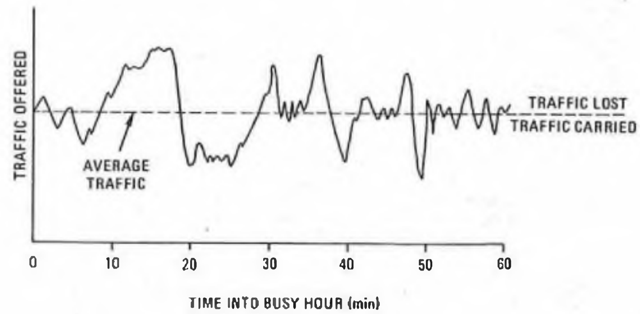
(c) Briefly explain why N must be significantly greater than 5.

A 9 (*a*) (*i*) A graph showing typical traffic variations over a 24 h period is given in sketch (*a*). Reasons for the variations are marked on the graph.



(*a*)

(*ii*) Sketch (*b*) shows a graph of typical traffic variations over the busy hour. The variations are due not, as in sketch (*a*), to specific events, but merely to the random nature of traffic. It should be noted that not only does the busy-hour traffic randomly fluctuate in the manner shown; each hour of sketch (*a*), if plotted to the same scale as sketch (*b*), would show a similar effect, although the average level, about which the random variations occur, would vary with the time of day.



(*b*)

(*b*) (*i*) The traffic lost during the busy hour

$$= \text{traffic offered} \times \text{busy-hour grade of service},$$
$$= 5 \times 0{\cdot}01 = 0{\cdot}05 \text{ erlang}.$$

The traffic carried

$$= \text{traffic offered} - \text{traffic lost},$$
$$= 5 - 0{\cdot}05 = \underline{4{\cdot}95 \text{ erlangs}}.$$

(*ii*) The number of calls

$$= \frac{\text{traffic (in erlangs)}}{\text{average duration of a call (in hours)}}.$$

Therefore, the number of calls lost during the busy hour

$$= \frac{0{\cdot}05 \times 60}{3} = \underline{1}.$$

(*c*) A traffic level of 5 erlangs is equivalent to 5 circuits continuously engaged for 1 h. Thus, at first sight, it would appear that only 5 circuits are necessary. This would be true if perfectly smooth traffic were offered (and setting-up and clearing-down times were neglected) but, in practice, traffic is offered either completely at random or, at best, only partially smoothed. Thus, there is a random fluctuation above and below the average level of 5 simultaneous calls. With random (pure-chance) traffic, it can be shown statistically that the probabilities of 0, 1, 2, 3, 4, 5, 6, 7, 8, 9, 10 etc. calls occurring at any instant follows a Poisson distribution; from this mathematical function, the number of circuits needed to give a specified grade of service can be estimated. It is found that, especially for small volumes of traffic where the random peaks are high compared with the average level, the number of circuits required is considerably greater than the traffic offered in erlangs.

Q 10 *(a) What characteristics of lead-acid cells make them particularly suitable for use in telephone-exchange power plant?*

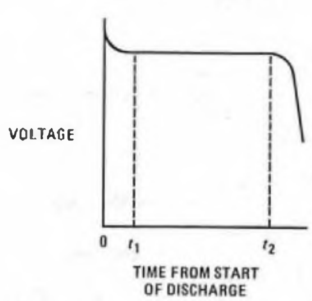
(b) For which types of power plant are 2 storage batteries used, and why?

(c) Explain why a hydrometer is normally used to indicate the state of charge of a lead-acid battery.

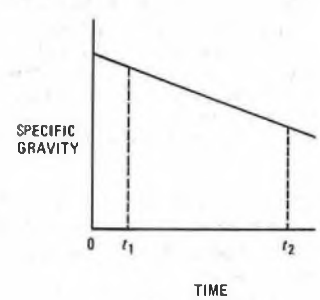
A 10 (*a*) The 2 principal characteristics of lead-acid cells which make them particularly suitable for telephone-exchange power plants are their

(*i*) steady voltage/time characteristic during discharge, and
(*ii*) low internal resistance.

The voltage/time characteristic is shown in sketch (*a*). When a fully-charged lead-acid cell starts to discharge through a constant load, the



(*a*)



(*b*)

voltage initially falls rapidly, until it stabilizes at approximately 2 V. It remains at this level during most of the discharge, and then falls away rapidly again as the cell approaches exhaustion.

The internal resistance of a lead-acid cell is very low, thus allowing heavy currents to be supplied without undue power loss in internal heating. The low resistance also minimizes crosstalk via the common battery.

(*b*) Charge-discharge systems consist of 2 batteries. At any time, one battery is discharging via the exchange load, and the other is being recharged while isolated from the load. At intervals, the batteries' roles are reversed. Two batteries are essential in such a system.

Float systems differ from charge-discharge systems in that the exchange load is supplied directly from a mains rectifier; the battery or batteries are connected in parallel with the rectifier. In the parallel-battery-float system, 2 such batteries are used, both in parallel with the rectifier. This permits one battery to be isolated for maintenance work, such as the periodic refresher charges which are normally given to floated batteries to maintain them in the fully-charged state.

(*c*) A hydrometer measures specific gravity, and it is the specific gravity of the acid in a cell which is usually used to indicate the state of charge of the cell. This is because there is a linear relationship between the specific gravity and the state of charge. The reason that the somewhat cumbersome mechanical measurement of specific gravity is used for this purpose, rather than an electrical method, is related to the voltage/time characteristic. If a cell supplies a constant current, its voltage/time characteristic is as shown in sketch (*a*). The charge remaining at any instant is decreasing at a constant rate. However, between times t_1 and t_2, the voltage of the cell remains constant. The current is defined as constant, and thus there is no simple electrical method of measuring the state of charge at any point. Sketch (*b*) shows, for comparison, the specific gravity over the same range of time, illustrating its linearly decreasing nature.

TELECOMMUNICATION PRINCIPLES C, 1976

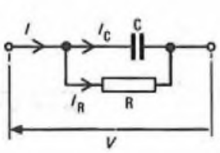
Students were expected to answer any 6 questions

Q 1 (*a*) *Explain, with the aid of a phasor diagram, the significance of the loss angle of a capacitor, and give a typical value for this angle.*

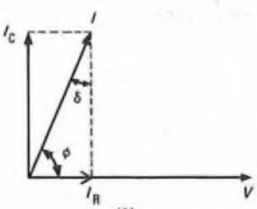
(*b*) *At 500 kHz, the equivalent circuit of a capacitor is 800 pF in series with 0·2 Ω. Calculate*

(*i*) *the reactance,*
(*ii*) *the loss angle,*
(*iii*) *the Q-factor,*
(*iv*) *the equivalent parallel resistance, and*
(*v*) *the power dissipated when the capacitor is connected to a 4 V source.*

A 1 (*a*) Sketch (*a*) shows the equivalent circuit of a capacitor, *C* farads, in which the losses are represented by a parallel resistance, *R* ohms. The phasor diagram is shown in sketch (*b*), in which δ is the loss angle.



(*a*)



(*b*)

In general, the loss angle of a capacitor is very small.

$$\therefore\ \delta \approx \frac{I_R}{I} \approx \frac{I_R}{I_C} = \frac{V/R}{V/X_C} = \frac{X_C}{R} \text{ radians,}$$

where I_R is the current in the parallel resistance, I is the supply current, I_C is the current in the capacitor, V is the supply voltage, and X_C is the reactance of the capacitor. (Also, for a low-loss capacitor, $\delta \approx \cos\phi$, where $\cos\phi$ is the power factor.)

A typical value for δ is between 10^{-3} and 10^{-4} rad.

(*b*) (*i*) Now, $X_C = \dfrac{1}{2\pi fC}$ ohms,

where f is the frequency (hertz).

$$\therefore\ X_C = \frac{1}{2\pi \times 500 \times 10^3 \times 800 \times 10^{-12}} = \underline{397 \cdot 9\ \Omega.}$$

(*ii*) For a series equivalent resistance of r ohms,

$$\delta = \frac{r}{X_C} \text{ radians,}$$

$$= \frac{0 \cdot 2}{397 \cdot 9} = \underline{5 \cdot 03 \times 10^{-4} \text{ rad.}}$$

(*iii*) The *Q*-factor is the ratio of the series reactance to the series resistance. Thus,

$$Q = \frac{X_C}{r} = \frac{1}{\delta} = \frac{1}{5 \cdot 03 \times 10^{-4}} = \underline{1988.}$$

(*iv*) Since $\quad \delta = \dfrac{r}{X_C} = \dfrac{X_C}{R},$

then $\quad R = \dfrac{X_C^2}{r}$ ohms,

$$= \frac{397 \cdot 9^2}{0 \cdot 2}\ \Omega = \underline{792\ \text{k}\Omega.}$$

(*v*) The power dissipated, *P* watts, is given by

$$P = \frac{V^2}{R} \text{ watts,}$$

$$= \frac{4^2}{792 \times 10^3}\ \text{W} = \underline{20 \cdot 2\ \mu\text{W}.}$$

Q 2 (*a*) *Explain the meaning of the terms characteristic impedance, wavelength and velocity of transmission, as applied to a transmission line.*

(*b*) *A 2 km loss-free transmission line is terminated in its characteristic impedance of (120 + j0) Ω. A source of e.m.f. 3 V and internal resistance 120 Ω, connected to the line input, produces a wavelength of 258 m on the line. Calculate*

(*i*) *the input voltage,*
(*ii*) *the input current,*
(*iii*) *the output voltage, and*
(*iv*) *the phase change along the line.*

(*c*) *Draw a phasor diagram relating the voltages at the input, mid-point and output of the line.*

A 2 (*a*) For an infinite transmission line, the characteristic impedance is the input impedance. If a section is cut from the infinite line and terminated in the characteristic impedance, the input impedance of the section is again the characteristic impedance.

The effect of the line reactances is to change the phase of the voltage (or current) along the transmission line. A wavelength is the distance between 2 adjacent points at which the voltages (or currents) are in phase.

The velocity of transmission, v, is the speed at which energy is propagated along the line, and is given by $v = f\lambda$, where f is the frequency and λ the wavelength.

(*b*) The transmission-line arrangements are shown in sketch (*a*), where V_s is the source voltage, Z_0 is the characteristic impedance, and R_s is the source resistance.

(*i*) Since the line is correctly terminated, the input impedance, Z_{in}, is 120 Ω, and the input voltage, V_{in}, is given by

$$V_{in} = \frac{Z_{in}}{R_s + Z_{in}} \times V_s = \frac{120 \times 3}{120 + 120} = \underline{1 \cdot 5\ \text{V}.}$$

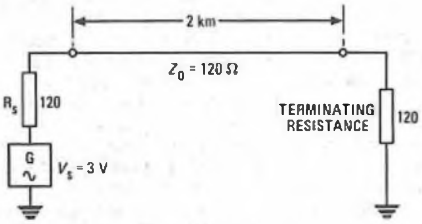
(*ii*) The input current

$$= \frac{V_{in}}{Z_{in}} = \frac{1 \cdot 5}{120}\ \text{A} = \underline{12 \cdot 5\ \text{mA}.}$$

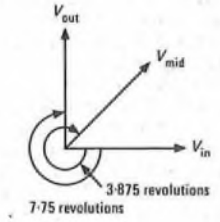
(*iii*) Since the line is loss free and correctly terminated, the output voltage, V_{out}, is equal to the input voltage.

$$\therefore\ \underline{V_{out} = 1 \cdot 5\ \text{V}.}$$

(*iv*) The line is $2 \times 10^3/258 = 7 \cdot 752$ wavelengths long. The total phase change is therefore $0 \cdot 752 \times 360° = \underline{270 \cdot 7°}$.



(*a*)



(*b*)

(*c*) A phasor diagram, relating the voltages at the input, mid-point and output, is shown in sketch (*b*). If the phase of the output voltage has made 7·75 revolutions with respect to the input voltage, that of the voltage at the mid-point, V_{mid}, has made 3·875 revolutions. Also, V_{mid} has the same magnitude as V_{in}.

Q 3 (*a*) *Explain the factors that cause power loss in iron-cored inductors.*

(*b*) *An iron-cored inductor has negligible conductor resistance and may be represented at a frequency of 1·59 kHz as an inductance of 500 mH, in parallel with a resistance of 10 kΩ representing the iron loss. By using complex numbers, or otherwise, determine the effective series resistance and series inductance of the inductor.*

(*c*) *If the inductor carries a current of 50 mA at a frequency of 1·59 kHz, determine*

(*i*) *the potential difference across the inductor, and*
(*ii*) *the power dissipated in the inductor.*

A 3 (*a*) Power loss in iron-cored inductors can be attributed to *copper loss* and *iron loss*.

Copper loss is the power dissipated by current flowing in the resistive winding of the inductor's coil.

Iron loss is the power dissipated in the magnetic circuit, and can be attributed to *eddy-current loss* and *hysteresis loss*. Eddy-current loss is the power dissipated by currents induced in the magnetic material. Hysteresis loss is the power dissipated in magnetizing the magnetic circuit. The flux-density/magnetic-field-strength (B/H) curve for a magnetic material is as shown in sketch (*a*). The energy expended per magnetization cycle is proportional to the area of the B/H loop.

(*b*) Sketch (*b*) shows the parallel equivalent circuit of the inductor, where L_p is the equivalent parallel inductance (henrys), and R_p is the equivalent parallel resistance (ohms). The admittance of the circuit, Y, represented in complex form, is given by

$$Y = \frac{1}{R_p} + \frac{1}{j2\pi f L_p}\ \text{siemens},$$

where f is the frequency (hertz).

$$\therefore\ Y = \frac{1}{10 \times 10^3} + \frac{1}{j2\pi \times 1{\cdot}59 \times 10^3 \times 500 \times 10^{-3}}\ \text{S},$$

$$= 10^{-4} - j2{\cdot}002 \times 10^{-4}\ \text{S}.$$

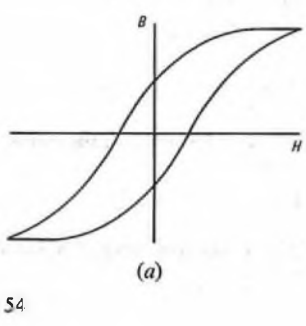
The impedance of the circuit, Z, is given by

$$Z = \frac{1}{Y} = \frac{10^4}{1 - j2{\cdot}002}\ \Omega,$$

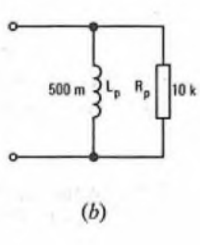
$$= \frac{(1 + j2{\cdot}002) \times 10^4}{1^2 + 2{\cdot}002^2} = \frac{(10 + j20{\cdot}02) \times 10^3}{5{\cdot}01}\ \Omega,$$

$$= 1{\cdot}996 + j3{\cdot}996\ \text{k}\Omega.$$

Thus, the effective series resistance is $\underline{1{\cdot}996\ \text{k}\Omega}$, and the effective series inductance is $3{\cdot}996 \times 10^3/(2\pi \times 1{\cdot}59 \times 10^3)$ H $= \underline{400\ \text{mH}}$.



(*a*)



(*b*)

(*c*) (*i*) In complex notation, the voltage across the inductor

$$= 50 \times 10^{-3} \times (1{\cdot}996 + j3{\cdot}996) \times 10^3\ \text{V},$$

$$= 99{\cdot}8 + j199{\cdot}8\ \text{V}.$$

The magnitude of the voltage

$$= \sqrt{(99{\cdot}8^2 + 199{\cdot}8^2)} = \underline{223{\cdot}3\ \text{V}}.$$

(*ii*) Power is dissipated only in the resistance. The power dissipated

$$= 50^2 \times 10^{-6} \times 1{\cdot}996 \times 10^3 = \underline{4{\cdot}99\ \text{W}}.$$

Q 4 (*a*) *With the aid of sketches, explain the principles of*

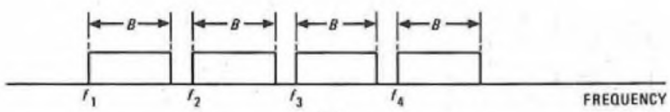
(*i*) *frequency-division multiplexing, and*
(*ii*) *time-division multiplexing.*

(*b*) *A time-division-multiplex system transmits telephone channels, each of which must be sampled 8000 times per second. If the minimum sample duration is 400 ns and the minimum switching time between one sample and the next is 100 ns, what is the maximum number of channels that can be transmitted?*

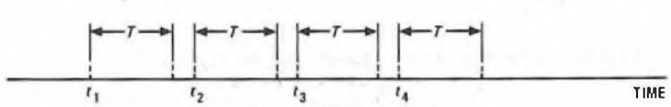
(*c*) *The signal on one channel in the system described in part* (*b*) *is sinusoidal, and has a peak value of 750 mV and a frequency of 1 kHz. If one sample on this channel occurs at the beginning of the sine wave, calculate the voltage of the next sample of the same signal.*

A 4 (*a*) (*i*) Sketch (*a*) shows the frequency spectrum of a 4-channel frequency-division-multiplex system. Each channel has the same bandwidth, B, but channel 1 extends from frequency f_1 to frequency $f_1 + B$, channel 2 extends from frequency f_2 to frequency $f_2 + B$, and so on. Multiplexing is achieved by allocating a different carrier frequency to each channel.

(*ii*) Sketch (*b*) illustrates a time-division-multiplex system in which the transmission time available is divided among 4 identical channels. Each channel is alloted a *time slot*, of duration T, during which a sample of the signal on that channel is transmitted. For channel 1, the sample begins at time t_1 and ends at time $t_1 + T$. Transmission of samples from all channels constitutes a *frame* and, at the end of one frame, the channels are again alloted successive time slots to form the next frame.



(*a*)



(*b*)

(*b*) The time between consecutive time slots for a particular channel is the reciprocal of the sampling rate, and is therefore 1/8000 s = 125 μs (that is, the duration of a frame is 125 μs).

The total time for one time slot and the guarding period between it and the next time slot is 400 + 100 = 500 ns.

Thus, the maximum number of channels

$$= \frac{125 \times 10^{-6}}{500 \times 10^{-9}} = \underline{250.}$$

(*c*) The signal can be expressed in the form

$$e = 750 \sin (2\pi \times 10^3 \times t)\ \text{millivolts},$$

where e is the instantaneous amplitude of the signal (volts), and t is the instantaneous time (seconds). If the first sample occurs at $t = 0$ s, the second occurs at $t = 125$ μs. Hence,

$$e = 750 \sin (2\pi \times 10^3 \times 125 \times 10^{-6})\ \text{mV},$$

$$= 750 \sin \pi/4 = \underline{530\ \text{mV}}.$$

Q 5 (*a*) *A 25 kΩ resistor, a 20 μH coil having a Q-factor of 150 at 3 MHz, and a capacitor are connected in parallel to form a circuit resonant at 3 MHz. Determine*

(*i*) *the capacitance required,*
(*ii*) *the impedance of the circuit at 3 MHz, and*
(*iii*) *the Q-factor of the circuit.*

(b) The parallel circuit is connected to a 3 MHz source having an e.m.f. of 20 V and an internal resistance of 10 kΩ. Calculate

(i) the total current taken,
(ii) the current in the capacitor, and
(iii) the current in the 25 kΩ resistor.

A 5 (*a*) (*i*) The circuit arrangement is shown in sketch (*a*). To a reasonable degree of accuracy, the resonant frequency occurs when the reactances of the inductor and capacitor are the same magnitude. Thus, $2\pi fL = 1/2\pi fC$, where f is the frequency (hertz), L is the inductance (henrys), and C is the capacitance (farads).

$$\therefore\ C = \frac{1}{4\pi^2 \times 3^2 \times 10^{12} \times 20 \times 10^{-6}}\,\text{F} = \underline{140{\cdot}7\ \text{pF}}.$$

(*ii*) The impedance, Z_T, of the circuit at 3 MHz is given by resistor R in parallel with the dynamic impedance, Z_D, of the resonant circuit. Now,

$$Z_D = \frac{L}{Cr} = \frac{\omega L}{\omega Cr}\ \text{ohms},$$

where r is the resistance of the inductor (ohms), and $\omega = 2\pi f$.

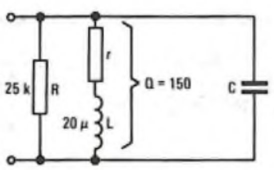
$$\therefore\ Z_D = \frac{\omega L}{r} \times \frac{1}{\omega C} = \frac{Q}{\omega C}\ \text{ohms, since } Q = \omega L/r.$$

$$\therefore\ Z_D = \frac{150}{2\pi \times 3 \times 10^6 \times 140{\cdot}7 \times 10^{-12}}\,\Omega = 56{\cdot}56\ \text{k}\Omega.$$

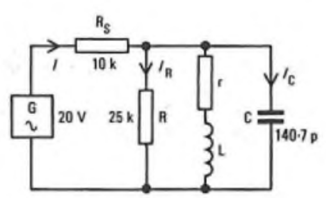
$$\therefore\ Z_T = \frac{25 \times 56{\cdot}56}{25 + 56{\cdot}56} = \underline{17{\cdot}34\ \text{k}\Omega.}$$

(*iii*) The value of Z_T obtained can be regarded as the dynamic impedance of the complete circuit. Thus, for the complete circuit,

$$Q = \omega C Z_T,$$
$$= 2\pi \times 3 \times 10^6 \times 140{\cdot}7 \times 10^{-12} \times 17{\cdot}34 \times 10^3 \approx \underline{46}.$$



(*a*)



(*b*)

(*b*) (*i*) The circuit arrangement is shown in sketch (*b*), in which resistor Rs represents the internal resistance of the source, I is the total current, I_R is the current in resistor R, and I_C is the current in the capacitor. The total current is given by

$$I = \frac{E}{R_S + Z_T}\ \text{amperes},$$

where E is the e.m.f. of the source (volts).

$$\therefore\ I = \frac{20}{10 \times 10^3 + 17{\cdot}34 \times 10^3}\,\text{A} = \underline{732\ \mu\text{A}.}$$

(*ii*) $I_C = QI = 46 \times 732 \times 10^{-6}\,\text{A} = \underline{33{\cdot}7\ \text{mA}.}$

(*iii*) $I_R = \dfrac{Z_D}{R + Z_D} \times I$ amperes,

$$= \frac{56{\cdot}56 \times 10^3 \times 732 \times 10^{-6}}{25 \times 10^3 + 56{\cdot}56 \times 10^3}\,\text{A} = \underline{508\ \mu\text{A}.}$$

Q 6 *(a) A 2-port (4-terminal) amplifier, having an open-circuit voltage gain of 400, is shown in Fig. 1. A 500 Ω load resistor is connected to the output port, and a 2 mV signal is applied to the input. Calculate*

(i) the input current,
(ii) the output current,
(iii) the current gain, and
(iv) the output voltage.

(b) Negative feedback is applied to the loaded amplifier described in part (a), such that 1% of the output voltage is fed back in series with the input. Determine the input current and output voltage for the same input voltage.

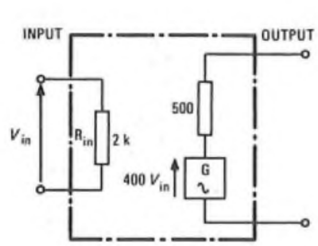
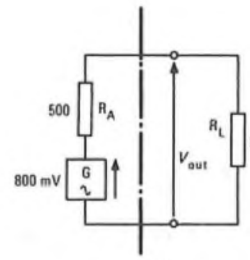


Fig. 1



(*a*)

A 6 (*a*) (*i*) The input current, I_{in}, is the ratio of the input voltage, V_{in}, to the input resistance, R_{in}.

$$\therefore\ I_{in} = \frac{V_{in}}{R_{in}} = \frac{2 \times 10^{-3}}{2 \times 10^3}\,\text{A} = \underline{1\ \mu\text{A}.}$$

(*ii*) The output circuit for the amplifier is shown in sketch (*a*). The voltage generator generates an e.m.f., E, of $400 \times 2 = 800$ mV. The output current, I_{out}, is given by

$$I_{out} = \frac{E}{R_A + R_L}\ \text{amperes},$$

where R_A is the output resistance of the amplifier (ohms), and R_L is the load resistance (ohms).

$$\therefore\ I_{out} = \frac{800 \times 10^{-3}}{500 + 500}\,\text{A} = \underline{800\ \mu\text{A}.}$$

(*iii*) The current gain $= I_{out}/I_{in} = \underline{800.}$

(*iv*) The output voltage, V_{out}, is given by

$$V_{out} = I_{out}R_L = 800 \times 10^{-6} \times 500\ \text{A} = \underline{400\ \text{mV}.}$$

(*b*) The voltage gain without feedback, A, is given by $V_{out}/V_{in} = 400/2 = 200$. From feedback theory, the voltage gain with feedback, A', is given by

$$A' = \frac{A}{1 + \beta A},$$

where β is the feedback fraction.

$$\therefore\ A' = \frac{200}{1 + 0{\cdot}01 \times 200} = 66{\cdot}7.$$

The output voltage with feedback applied

$$= A'V_{in} = 66{\cdot}7 \times 2 = \underline{133{\cdot}4\ \text{mV}.}$$

To determine the input current, it is necessary to find the modified input resistance, R_{in}'. Now,

$$R_{in}' = R_{in}(1 + \beta A),$$
$$= 2 \times 10^3 \times (1 + 0{\cdot}01 \times 200)\,\Omega = 6\ \text{k}\Omega.$$

Thus, the input current

$$= \frac{V_{in}}{R_{in}'} = \frac{2 \times 10^{-3}}{6 \times 10^3}\,\text{A} = \underline{333\ \text{nA}.}$$

Q 7 *(a) The current from a modulator is given by*

$$i = 5 \sin(2\pi \times 10^4 t) + 1{\cdot}5 \cos(2\pi \times 0{\cdot}95 \times 10^4 t) - 1{\cdot}5 \cos(2\pi \times 1{\cdot}05 \times 10^4 t)\ \textit{amperes}.$$

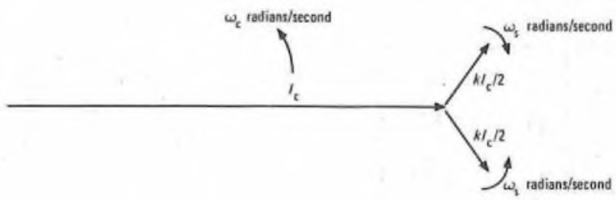
Identify the type of modulation, and determine

(i) the carrier frequency,
(ii) the signal frequency,
(iii) the r.m.s. value of the carrier component, and
(iv) the r.m.s. value of the total current.

(b) If the modulator current is passed through a 200 Ω resistor, determine

(i) the power due to the carrier component,
(ii) the total power, and
(iii) the peak instantaneous power.

A 7 (*a*) The equation is an example of the general form for an <u>amplitude-modulated</u> signal:



$$i = I_c \sin \omega_c t + \frac{kI_c}{2}\{\cos(\omega_c - \omega_s)t - \cos(\omega_c + \omega_s)t\},$$

where I_c is the peak carrier current (equal to 5 A in the given equation), $\omega_c = 2\pi f_c t$ where f_c is the carrier frequency, $\omega_s = 2\pi f_s t$ where f_s is the frequency of the modulating signal, t is the instantaneous time, and k is the modulation factor. The amplitude-modulated carrier is illustrated in the sketch.

(*i*) Since $\omega_c = 2\pi f_c t = 2\pi \times 10^4 t$ radians/second,

$f_c = 10$ kHz.

(*ii*) Since $f_c - f_s = 0{\cdot}95 \times 10^4$ Hz (or $f_c + f_s = 1{\cdot}05 \times 10^4$ Hz),

$f_s = 500$ Hz.

(*iii*) The r.m.s. value of the carrier current

$= 5/\sqrt{2} = 3{\cdot}54$ A.

(*iv*) The r.m.s. value of the total current

$$= \sqrt{\left\{\left(\frac{I_c}{\sqrt{2}}\right)^2 + \left(\frac{kI_c}{2\sqrt{2}}\right)^2 + \left(\frac{kI_c}{2\sqrt{2}}\right)^2\right\}} \text{ amperes,}$$

$$= \sqrt{\left(\frac{25}{2} + \frac{1{\cdot}5^2}{2} + \frac{1{\cdot}5^2}{2}\right)} = 3{\cdot}84 \text{ A.}$$

(*b*) (*i*) The power due to the carrier current

$= 3{\cdot}54^2 \times 200$ W $= 2{\cdot}5$ kW.

(*ii*) The total power $= 3{\cdot}84^2 \times 200$ W $= 2{\cdot}95$ kW.

(*iii*) To calculate the peak instantaneous power, it is necessary to find the peak instantaneous current. This occurs when the 3 current components are in phase, and thus has the value $5 + 1{\cdot}5 + 1{\cdot}5 = 8$ A. Hence, the peak instantaneous power

$= 8^2 \times 200$ W $= 12{\cdot}8$ kW.

Q 8 (*a*) *Positive feedback is applied to an amplifier to cause oscillation. What is the condition for the maintenance of these oscillations?*

(*b*) *An oscillator is to consist of an amplifier having a phase shift of 180°, the output of which is connected to a tuned circuit. Show 2 methods by which the necessary feedback can be arranged.*

(*c*) *Sketch a full circuit diagram of an oscillator, and briefly describe its operation.*

Q 9 (*a*) *Sketch a set of typical collector characteristics for a power transistor. For a power-supply voltage, marked V_{CC} on the characteristics, show typical load lines for*

(*i*) *a load resistor, and*
(*ii*) *a transformer-coupled resistive load.*

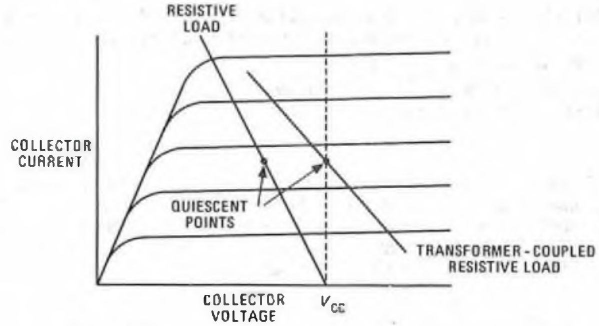
(*b*) *A power transistor is supplied from a 15 V power supply, and feeds a 10 Ω load resistance through a 2:1 step-down transformer. The collector current varies sinusoidally between 650 mA and 50 mA. Determine*

(*i*) *the output power,*
(*ii*) *the power taken from the supply,*
(*iii*) *the efficiency, and*
(*iv*) *the collector dissipation.*

A 9 (*a*) The sketch shows a set of typical collector characteristics, with load lines for a resistive load and a transformer-coupled load.

(*b*) (*i*) Since the transformer is a step-down transformer, the collector load is $10 \times 2^2 = 40$ Ω. The peak-to-peak collector current is 600 mA. Thus, the output power

$$= \left(\frac{300 \times 10^{-3}}{\sqrt{2}}\right)^2 \times 40 = 1{\cdot}8 \text{ W.}$$



(*ii*) The average current taken from the power supply is $(650 + 50)/2 = 350$ mA. Thus, the power taken from the supply

$= 15 \times 350 \times 10^{-3} = 5{\cdot}25$ W.

(*iii*) The efficiency is the ratio of the output power to the power taken from the supply. Hence, the efficiency

$$= \frac{1{\cdot}8}{5{\cdot}25} \times 100\% = 34{\cdot}3\%.$$

(*iv*) The collector dissipation is the difference between the power supplied and the output power. Hence, the collector dissipation

$= 5{\cdot}25 - 1{\cdot}8 = 3{\cdot}45$ W.

Q 10 (*a*) *A variable resistor, R, is connected to a source. The potential difference, V_R, across the resistor is measured for each resistor setting, and the results are given in the table.*

R (Ω)	*100*	*200*	*300*	*400*	*500*	*600*
V_R (*V*)	*2·7*	*4·4*	*5·5*	*6·4*	*7·1*	*7·6*

Assuming the source to have a resistive internal impedance, determine the Thévenin equivalent circuit of the source.

(*b*) *The above source is connected to terminals AB of the circuit shown in Fig. 2. Determine the turns ratio, N, of the transformer for maximum power in the 2 Ω load, and calculate that power.*

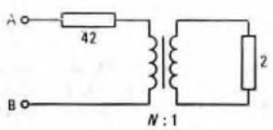
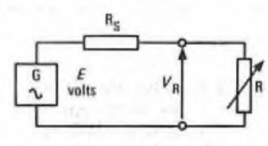


Fig. 2



(*a*)

A 10 (*a*) Sketch (*a*) shows the circuit arrangements, where resistor R_S represents the internal impedance of the source, and the source is assumed to generate E volts. Since resistors R_S and R form a potential divider,

$$V_R = E \times \frac{R}{R + R_s}.$$

$$\therefore \quad \frac{R}{V_R} = \frac{R}{E} + \frac{R_S}{E}.$$

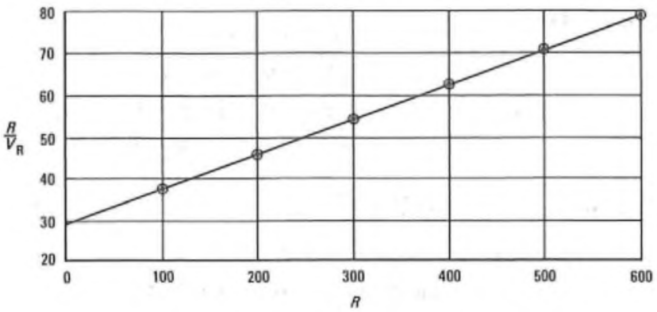
This equation is of the standard form for a straight line, $y = mx + c$, where $y = R/V_R$, $x = R$, $m = 1/E$, and $c = R_S/E$. The table shows values of R/V_R for the given values of R.

R	100	200	300	400	500	600
R/V_R	37	45·5	54·6	62·5	70·4	79

Sketch (*b*) shows the graph of (R/V_R) plotted against R.

From the graph, $1/E = (79 - 29)/600$. Therefore, the Thévenin equivalent voltage generator (E) is 12 V.

Also, $R_S/E = 29$, so that the Thévenin equivalent series resistance (R_S) is 348 Ω.



(*b*)

(*b*) The effective primary resistance of the circuit when the source is connected is 348 + 42 = 390 Ω. For maximum power in the load, the reflected load must equal the primary resistance. Hence,

$$N = \sqrt{\frac{390}{2}} = \underline{14:1.}$$

The load resistance reflected into the primary circuit is 390 Ω, so that the primary circuit consists of a 12 V generator feeding two 390 Ω resistors in series, one of which is the load. Hence, the power in the load

$$= \frac{6^2}{390}\text{ W} = \underline{92\cdot 3\text{ mW.}}$$

CORRECTION

TELECOMMUNICATION PRINCIPLES C, 1975

(*Supplement*, Vol. 69, Oct. 1976)

A 10 The vertical scale of sketch (*a*) is misleadingly shown as a logarithmic scale. The scale should be linear and calibrated in terms of the spectral density (watts/hertz). The upper point should have the value 1·0 W/Hz, and the lower, 0·1 W/Hz.

MATHEMATICS C, 1976

Students were expected to answer any 6 questions

Q 1 (*a*) *The second term of a geometric progression is* -48, *and the fifth term is* $+20\frac{1}{4}$. *Find the first term of the series, the common ratio, and the sum to infinity of the progression.*

(*b*) *The r.m.s. voltage measured at the receiving end of a transmission line, 20 km long, is found to be 5% of the voltage at the sending end. If the voltage, V, at a distance x kilometres is given by* $V = V_0e^{-kx}$, *where k is a constant, calculate the value of k to 2 significant figures.*

A 1 (*a*) If a denotes the first term of a geometric progression, and r is the common ratio, then

$$\text{the second term} = ar = -48, \qquad \ldots\ldots (1)$$

and

$$\text{the fifth term} = ar^4 = +20\tfrac{1}{4}. \qquad \ldots\ldots (2)$$

Dividing equation (2) by equation (1) gives

$$r^3 = -\frac{20\frac{1}{4}}{48} = -\frac{81}{192} = -\frac{27}{64}.$$

$$\therefore\ r = \left(-\frac{27}{64}\right)^{1/3} = \underline{-\frac{3}{4}}.$$

Substituting for r in equation (1) gives

$$-\frac{3a}{4} = -48.$$

$$\therefore\ a = \frac{4\times 48}{3} = \underline{64.}$$

The sum to infinity

$$= \frac{a}{1-r},$$

$$= \frac{64}{1-(-\frac{3}{4})} = \underline{36\tfrac{4}{7}.}$$

(*b*) When $x = 0$,

$$V = V_0e^{-kx} = V_0e^0 = V_0;$$

hence, V_0 is the r.m.s. voltage at the sending end.

When $x = 20$ km,

$$V = \frac{5V_0}{100} = V_0e^{-20k}.$$

$$\therefore\ e^{-20k} = 0\cdot 05.$$

$$\therefore\ -20k = \log_e 0\cdot 05 = -2\cdot 9957.$$

$$\therefore\ \underline{k = 0\cdot 15}, \text{ to 2 significant figures.}$$

Q 2 (*a*) *Expand* $(1+x)^{-1/2}$ *by the binomial series as far as the term in* x^4. *State the term involving* x^r, *and the limitations to be placed on the value of x for the series to be valid.*

(*b*) *The chance of throwing 3 sixes when 5 dice are thrown at the same time is calculated by evaluating the term involving* p^2q^3 *in the binomial series given by* $(p+q)^5$, *where* $p = 5/6$ *and* $q = 1/6$. *Show that there is approximately a 1-in-30 chance of this particular event occurring.*

A 2 (*a*)

$$(1+x)^{-1/2} = 1 + (-\tfrac{1}{2})x + \frac{-\frac{1}{2}(-\frac{1}{2}-1)x^2}{1\times 2} + \frac{-\frac{1}{2}(-\frac{1}{2}-1)(-\frac{1}{2}-2)x^3}{1\times 2\times 3} + \frac{-\frac{1}{2}(-\frac{1}{2}-1)(-\frac{1}{2}-2)(-\frac{1}{2}-3)x^4}{1\times 2\times 3\times 4} + \ldots,$$

$$= 1 - \frac{x}{2} + \frac{(\frac{1}{2}\times\frac{3}{2})x^2}{1\times 2} - \frac{(\frac{1}{2}\times\frac{3}{2}\times\frac{5}{2})x^3}{1\times 2\times 3} + \frac{(\frac{1}{2}\times\frac{3}{2}\times\frac{5}{2}\times\frac{7}{2})x^4}{1\times 2\times 3\times 4} - \ldots,$$

$$= \underline{1 - \frac{x}{2} + \frac{3x^2}{8} - \frac{5x^3}{16} + \frac{35x^4}{128} - \ldots.}$$

The term involving x^r can be written as

$$\underline{(-1)^r \times \frac{1\times 3\times 5\ldots\times(2r-1)\times x^r}{2^r\times r!}}.$$

This series is true provided that the numerical value of x is less than unity.

(*b*)

$$(p+q)^5 = p^5 + 5p^4q + \frac{(5\times 4)\,p^3q^2}{1\times 2} + \frac{(5\times 4\times 3)p^2q^3}{1\times 2\times 3} + \ldots.$$

When $p = 5/6$ and $q = 1/6$, the term involving p^2q^3 is

$$\frac{5\times 4\times 3}{1\times 2\times 3}\times\left(\frac{5}{6}\right)^2\times\left(\frac{1}{6}\right)^3 = \frac{10\times 25}{6^5},$$

$$= \frac{250}{7776} = \frac{1}{31\cdot 1},$$

so that the chance of throwing 3 sixes is 1 in 31·1, or approximately <u>1 in 30.</u>

Q 3 (*a*) *Solve approximately the equation* $e^x + e^{-x} = 2\cdot 4$, *in which x is positive. Use the exponential series to form a quadratic equation in* x^2, *and assume that* x^6 *and higher powers of x can be neglected.*

(*b*) *Show that the exact solution of the original equation is*

$$\log_e(1\cdot 2 + \sqrt{0\cdot 44}).$$

Evaluate this, and show that the percentage error arising from the assumption in part (*a*) *is less than 1%.*

A 3 (*a*) $$e^x = 1 + x + \frac{x^2}{2!} + \frac{x^3}{3!} + \frac{x^4}{4!} + \ldots,$$

and $$e^{-x} = 1 - x + \frac{x^2}{2!} - \frac{x^3}{3!} + \frac{x^4}{4!} - \ldots.$$

$$\therefore\ e^x + e^{-x} = 2 + x^2 + \frac{2x^4}{4!} + \ldots.$$

Neglecting terms in x^6 and higher powers,

$$2 + x^2 + \frac{x^4}{12} \approx 2\cdot4.$$

Let $y = x^2$; then,

$$\frac{y^2}{12} + y - 0\cdot4 = 0,$$

or $$\underline{y^2 + 12y - 4\cdot8 = 0.}$$

$$\therefore\ y = \frac{-12 \pm \sqrt{(144 + 19\cdot2)}}{2},$$

$$= \frac{-12 \pm 12\cdot77}{2} = 0\cdot39 \text{ or } -12\cdot39.$$

Since $x = \sqrt{y}$, and x is positive, the negative solutions can be discarded, so that

$$x = +\sqrt{0\cdot39} = \underline{0\cdot6245.}$$

(*b*) Let $y = e^x$; then,

$$y + \frac{1}{y} = 2\cdot4,$$

or $y^2 - 2\cdot4y + 1 = 0.$

$$\therefore\ y = \frac{2\cdot4 \pm \sqrt{(5\cdot76 - 4)}}{2}.$$

$$\therefore\ e^x = 1\cdot2 \pm \sqrt{0\cdot44}.$$

$$\therefore\ \underline{x = \log_e(1\cdot2 + \sqrt{0\cdot44}).} \qquad \text{QED.}$$

Note: Since $\sqrt{0\cdot44} = 0\cdot6633$, the alternative value for e^x, of $1\cdot2 - \sqrt{0\cdot44}$, is less than unity, giving a negative value of x.

The exact value of x

$$= \log_e(1\cdot2 + 0\cdot6633),$$

$$= \log_e 1\cdot8633 = \underline{0\cdot6223.}$$

The percentage error in the value of x obtained in part (*a*)

$$= \frac{0\cdot6245 - 0\cdot6223}{0\cdot6223} \times 100\%,$$

$$\approx \frac{0\cdot22}{0\cdot62}\%,$$

which is approximately $\underline{0\cdot\dot{3}\%}$, and hence is well under 1%.

Q 4 (*a*) *Assuming the expansions for* $\cos(A+B)$ *and* $\sin(A+B)$,

(*i*) *prove that* $\cos 2\theta = 2\cos^2\theta - 1$,
(*ii*) *express* $\cos 4\theta$ *in powers of* $\cos\theta$, *and*
(*iii*) *express* $\tan 2\theta$ *in terms of* $\tan\theta$.

(*b*) *The current output of a modulator is i milliamperes, and is given by*
$$i = 10\cos(2\pi t \times 10^3)\sin(\pi t \times 10^5).$$
(*i*) *Express this current as the sum or difference of 2 separate sinusoids.*
(*ii*) *Give the frequency, in hertz, of each of these sinusoids.*

A 4 (*a*) Now, $\cos(A+B) = \cos A\cos B - \sin A\sin B$, ... (1)

and $\sin(A+B) = \sin A\cos B + \cos A\sin B$. ...... (2)

(*i*) Now, $\cos 2\theta = \cos(\theta + \theta)$.

If $A = B = \theta$, then, using expansion (1),

$$\cos 2\theta = \cos^2\theta - \sin^2\theta.$$

But $\sin^2\theta + \cos^2\theta = 1$, or $\sin^2\theta = 1 - \cos^2\theta$.

$$\therefore\ \cos 2\theta = \cos^2\theta - (1 - \cos^2\theta),$$

$$= \underline{2\cos^2\theta - 1.} \qquad \text{QED.}$$

(*ii*) Now, $\cos 4\theta = \cos(2\theta + 2\theta)$,
which, from part (*i*),

$$= 2\cos^2 2\theta - 1,$$
$$= 2(2\cos^2\theta - 1)^2 - 1,$$
$$= 2(4\cos^4\theta - 4\cos^2\theta + 1) - 1,$$
$$= \underline{8\cos^4\theta - 8\cos^2\theta + 1.}$$

(*iii*) Now, $\tan 2\theta = \dfrac{\sin 2\theta}{\cos 2\theta}$.

If $A = B = \theta$, then, from expansion (2),

$$\sin 2\theta = \sin\theta\cos\theta + \cos\theta\sin\theta,$$
$$= 2\sin\theta\cos\theta.$$

From expansion (1),

$$\cos 2\theta = \cos^2\theta - \sin^2\theta.$$

$$\therefore\ \tan 2\theta = \frac{2\sin\theta\cos\theta}{\cos^2\theta - \sin^2\theta}.$$

Dividing the numerator and denominator by $\cos^2\theta$ gives

$$\underline{\tan 2\theta = \frac{2\tan\theta}{1 - \tan^2\theta}.}$$

(*b*) (*i*) If $B = -B$, then expansion (2) becomes

$$\sin(A - B) = \sin A\cos(-B) + \cos A\sin(-B).$$

Since $\cos B = \cos(-B)$, and $\sin B = -\sin(-B)$, then

$$\sin(A - B) = \sin A\cos B - \cos A\sin B. \qquad \ldots\ldots (3)$$

Adding expansions (2) and (3) gives

$$\sin(A+B) + \sin(A-B) = 2\sin A\cos B.$$

$$\therefore\ \sin A\cos B = \tfrac{1}{2}\{\sin(A+B) + \sin(A-B)\}.$$

Hence, if $A = \pi t \times 10^5$ and $B = 2\pi t \times 10^3$,

$$i = 10 \times \tfrac{1}{2}\{\sin\pi t(10^5 + 2\times10^3) + \sin\pi t(10^5 - 2\times10^3)\},$$

$$= \underline{5\sin(102\pi t\times10^3) + 5\sin(98\pi t\times10^3).}$$

(*ii*) For the first sinusoid, if f is the frequency,

$$2\pi ft = 102\pi t\times10^3.$$

$$\therefore\ \underline{f = 51\,000 \text{ Hz.}}$$

Similarly, for the second sinusoid, $2\pi ft = 98\pi t\times10^3$.

$$\therefore\ \underline{f = 49\,000 \text{ Hz.}}$$

Q 5 (*a*) *Sketch the curve given by the polar equation* $8/r = 2 + \cos\theta$, *over the range from* $\theta = 0$ *to* $\theta = 2\pi$ *rad. Mark on the curve the points A, B, C and D where* $\theta = 0$, $\pi/2$, $2\pi/3$ *and* π *rad respectively.*

(*b*) *Prove that the Cartesian equation of the same curve is*

$$3x^2 + 4y^2 + 16x = 64,$$

referred to the pole as origin and the line $\theta = 0$ *as the positive x-axis.*

A 5 See A6, Mathematics C, 1975, *Supplement*, Vol. 69, p. 62, Oct. 1976.

Q 6 (*a*) *Prove, from first principles, that* $\dfrac{d\tan\theta}{d\theta} = \sec^2\theta$.

(*b*) *The mass, m, of a particle increases with its velocity, v, such that*

$$m = \frac{m_0}{\sqrt{(1 - v^2/c^2)}},$$

where c is the velocity of light.

(*i*) *What does the constant,* m_0, *represent?*
(*ii*) *If the velocity varies with time, t, show that*

$$(c^2 - v^2)\frac{dm}{dt} = mv\frac{dv}{dt}.$$

A 6 (*a*) Let $y = \tan\theta = \dfrac{\sin\theta}{\cos\theta}$.

Let θ increase by a small amount, $\delta\theta$, and let δy be the corresponding change in y. Then,

$$y + \delta y = \frac{\sin(\theta + \delta\theta)}{\cos(\theta + \delta\theta)}.$$

Since $y = \sin\theta/\cos\theta$,

$$\delta y = \frac{\sin(\theta + \delta\theta)}{\cos(\theta + \delta\theta)} - \frac{\sin\theta}{\cos\theta},$$

$$= \frac{\sin(\theta + \delta\theta)\cos\theta - \cos(\theta + \delta\theta)\sin\theta}{\cos(\theta + \delta\theta)\cos\theta}.$$

Now, $\sin(A - B) = \sin A\cos B - \cos A\sin B$.

Hence, $\sin(\theta + \delta\theta)\cos\theta - \cos(\theta + \delta\theta)\sin\theta$

$$= \sin\{(\theta + \delta\theta) - \theta\} = \sin\delta\theta.$$

$$\therefore\ \delta y = \frac{\sin\delta\theta}{\cos(\theta + \delta\theta)\cos\theta}.$$

$$\therefore\ \frac{\delta y}{\delta\theta} = \frac{\dfrac{\sin\delta\theta}{\delta\theta}}{\cos(\theta + \delta\theta)\cos\theta}.$$

Now, $\lim\limits_{\delta\theta\to 0}\dfrac{\sin\delta\theta}{\delta\theta} = 1$ $\left(\text{since } \dfrac{\sin\theta}{\theta} \text{ always lies between unity and } \cos\theta\right)$.

$$\therefore\ \lim_{\delta\theta\to 0}\frac{\delta y}{\delta\theta} = \frac{dy}{d\theta} = \frac{1}{\cos^2\theta} = \underline{\sec^2\theta}. \qquad \text{QED.}$$

(*b*) (*i*) For $m = m_0$,

$$\sqrt{(1 - v^2/c^2)} = 1.$$

$$\therefore\ \sqrt{(c^2 - v^2)} = c.$$

$$\therefore\ c^2 - v^2 = c^2.$$

$$\therefore\ v = 0.$$

Thus, m_0 represents the mass of a particle having zero velocity.

(*ii*)

$$m = m_0(1 - v^2/c^2)^{-1/2}.$$

$$\therefore\ \frac{dm}{dv} = -\frac{m_0}{2}\left(1 - \frac{v^2}{c^2}\right)^{-3/2} \times \left(-\frac{2v}{c^2}\right),$$

$$= m_0 \times (1 - v^2/c^2)^{-1/2} \times (1 - v^2/c^2)^{-1} \times \frac{v}{c^2},$$

$$= m \times \frac{v}{c^2(1 - v^2/c^2)} = \frac{mv}{c^2 - v^2}.$$

But,

$$\frac{dm}{dt} = \frac{dm}{dv} \times \frac{dv}{dt}.$$

$$\therefore\ \frac{dm}{dt} = \frac{mv}{c^2 - v^2} \times \frac{dv}{dt}.$$

$$\therefore\ \underline{(c^2 - v^2)\frac{dm}{dt} = mv\frac{dv}{dt}}. \qquad \text{QED.}$$

Q 7 (*a*) *Obtain dy/dx for the following functions, simplifying each result:*

(*i*) $y = x^3e^{-3x}$, *and*
(*ii*) $y = (2x^2 - 7)^{3/2}$.

(*b*) *State and prove the rule for differentiating the quotient of 2 functions of x.*

(*c*) *Find the maximum and minimum values of*

$$y = \frac{(x + 1)^2}{2x - 5},$$

and sketch the graph of this function, displaying both positive and negative values of x.

A 7 (*a*) (*i*) $\dfrac{dy}{dx} = x^3e^{-3x} \times (-3) + e^{-3x} \times 3x^2 = \underline{3x^2e^{-3x}(1 - x)}.$

(*ii*) $\dfrac{dy}{dx} = \dfrac{3}{2}(2x^2 - 7)^{1/2} \times 4x = \underline{6x(2x^2 - 7)^{1/2}}.$

(*b*) For proof of the rule for differentiating a quotient, see A7, Mathematics C, 1973, *Supplement*, Vol. 67, p. 61, Oct. 1974.

(*c*)

$$\frac{dy}{dx} = \frac{(2x - 5) \times 2(x + 1) - (x + 1)^2 \times 2}{(2x - 5)^2},$$

$$= \frac{2(x + 1)(2x - 5 - x - 1)}{(2x - 5)^2}.$$

$$= \frac{2(x + 1)(x - 6)}{(2x - 5)^2}.$$

$$\frac{d^2y}{dx^2} = \frac{(2x - 5)^2 \times 2(2x - 5) - 2(x^2 - 5x - 6) \times 2(2x - 5) \times 2}{(2x - 5)^4},$$

$$= \frac{2}{2x - 5} - \frac{8(x^2 - 5x - 6)}{(2x - 5)^3}.$$

For maximum and minimum values of y, $dy/dx = 0$.

$$\therefore\ \frac{2(x + 1)(x - 6)}{(2x - 5)^2} = 0.$$

$\therefore\ (x + 1)(x - 6) = 0$, whence $x = -1$ or 6.

When $x = -1$, $\dfrac{d^2y}{dx^2} = -\dfrac{2}{7} - \dfrac{8(1 + 5 - 6)}{(-7)^3} = -\dfrac{2}{7}.$

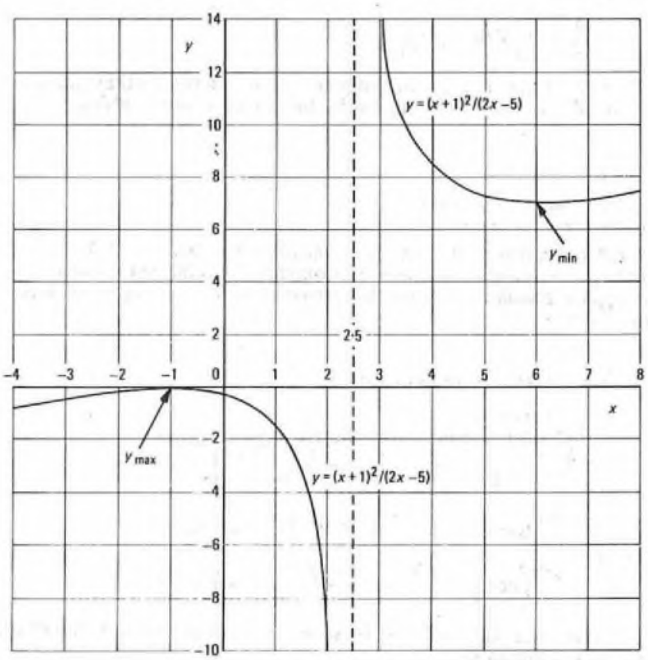
This is negative, and hence a maximum value occurs at $x = -1$. Substituting $x = -1$ into the expression for y gives the maximum value as $\underline{y_{max} = 0}$.

When $x = 6$, $\dfrac{d^2y}{dx^2} = \dfrac{2}{7} - \dfrac{8(36 - 30 - 6)}{7^3} = \dfrac{2}{7}.$

This is positive, and hence a minimum value occurs at $x = 6$. Substitution gives the minimum value as $\underline{y_{min} = 7}$.

The numerator of the function is positive for all values of x, except for $x = -1$, when it is zero. When $x = 2{\cdot}5$, the denominator is zero, and hence $y = \infty$. The graph is shown in the sketch, constructed from the following table of values.

x	-4	-2	-1	0	2	$2{\cdot}5$	4	6	8
$(x+1)^2$	9	1	0	1	9	12·25	25	49	81
$2x - 5$	−13	−9	−7	−5	−1	0	3	7	11
y	−0·69	$-0{\cdot}\dot{1}$	0	−0·2	−9	$\pm\infty$	$8{\cdot}\dot{3}$	7	$7{\cdot}\dot{3}\dot{6}$



The graph is asymptotic to the line $x = 2{\cdot}5$ for both parts of the curve. For larger negative and positive values of x, the respective parts of the curve descend and ascend at an increasing rate.

Q 8 (*a*) *By exact integration, evaluate* $\int_0^{\pi/3} \cos 2x\, dx$.

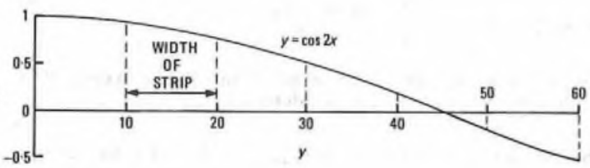
(*b*) *Use Simpson's rule, with 7 ordinates, to evaluate approximately the integral in part (a). Determine the degree of accuracy of this approximate method in this instance.*

A 8 (*a*) $$\int_0^{\pi/3} \cos 2x\, \mathrm{d}x = \left[\frac{\sin 2x}{2}\right]_0^{\pi/3},$$

$$= \frac{\sin(2\pi/3)}{2} = \frac{\sqrt{3}}{4} = \underline{0{\cdot}433.}$$

(*b*) The use of Simpson's rule with 7 ordinates demands division of the range of values of x into 6 equal parts of $\pi/18$ rad (10°) each. The values of the ordinates are derived in the table, from which the curve of the function over the stated range may be drawn, as shown in the sketch.

$x°$	0	10	20	30	40	50	60
$2x°$	0	20	40	60	80	100	120
$\cos 2x$	1	0·9397	0·766	0·5	0·1736	−0·1736	−0·5



By Simpson's rule, $\int_0^{\pi/3} \cos 2x\, \mathrm{d}x$

= ⅓ (width of strip) × (sum of first and last ordinates
+ twice the sum of remaining odd ordinates
+ 4 times the sum of the even ordinates),

$$= \frac{\pi}{3 \times 18}\{1 - 0{\cdot}5 + 2(0{\cdot}766 + 0{\cdot}1736) + 4(0{\cdot}9397 + 0{\cdot}5 - 0{\cdot}1736)\},$$

$$= \frac{\pi}{54}(0{\cdot}5 + 1{\cdot}8792 + 5{\cdot}0644),$$

$$= \frac{\pi}{54} \times 7{\cdot}4436 = \underline{0{\cdot}4331.}$$

To 3 significant figures, the answers agree exactly, so that the error is zero, within the limits imposed by the use of 4-figure tables.

Q 9 (*a*) *Calculate*

(*i*) *the mean value, and*
(*ii*) *the r.m.s. value*

of the function $v = 5 \sin 100\pi t$, *over the range* $t = 0$ *to* $t = 2{\cdot}5 \times 10^{-3}$.
(*b*) *By regarding a solid cone as a volume of revolution (or otherwise), use integral calculus to obtain the volume of such a cone of height h and base radius a.*

A 9 (*a*) (*i*) The mean value of v

$$= \frac{\int_0^{0{\cdot}0025} 5 \sin 100\pi t\, \mathrm{d}t}{2{\cdot}5 \times 10^{-3}} = \frac{5000}{2{\cdot}5}\left[-\frac{\cos 100\pi t}{100\pi}\right]_0^{0{\cdot}0025},$$

$$= -\frac{20}{\pi}\{\cos(100\pi \times 2{\cdot}5 \times 10^{-3}) - \cos 0\},$$

$$= -\frac{20}{\pi}\left(\cos\frac{\pi}{4} - 1\right) = -\frac{20}{\pi}\left(\frac{1}{\sqrt{2}} - 1\right) = \underline{1{\cdot}865.}$$

(*ii*) The r.m.s. value of v is the square root of the mean value of v^2. The mean value of v^2

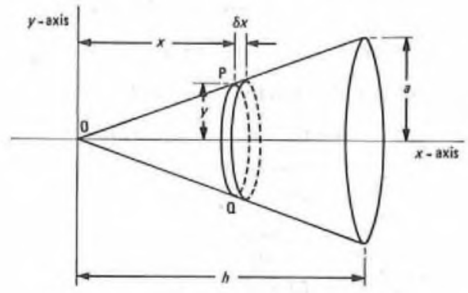
$$= \frac{\int_0^{0{\cdot}0025} 25 \sin^2 100\pi t\, \mathrm{d}t}{2{\cdot}5 \times 10^{-3}} = 10^4 \int_0^{0{\cdot}0025} \frac{1 - \cos 200\pi t}{2}\, \mathrm{d}t,$$

since $\cos 200\pi t = 1 - 2\sin^2 100\pi t$.

$$\therefore (v^2)_{\text{mean}} = \frac{10^4}{2}\left[t - \frac{\sin 200\pi t}{200\pi}\right]_0^{0{\cdot}0025},$$

$$= \frac{10^4}{2}\left(2{\cdot}5 \times 10^{-3} - \frac{\sin \pi/2}{200\pi}\right) = 12{\cdot}5 - \frac{25}{\pi}.$$

$$\therefore v_{\text{r.m.s.}} = \sqrt{(12{\cdot}5 - 7{\cdot}958)} = \underline{2{\cdot}131.}$$



(*b*) The sketch shows the cone with its vertex at the origin, O, of a set of Cartesian co-ordinates, and its axis set along the x-axis. Considering a thin slice, PQ, of the cone, if the ordinate, y, to point P is rotated about the x-axis, it generates a circle of area πy^2. If the thickness of the slice is δx (a small part of the abscissa, x, through point P), the volume of the slice is approximately $\pi y^2 \delta x$.
The total volume of the cone

$$= \lim_{\delta x \to 0} \sum_{x=0}^{x=h} \pi y^2 \delta x = \int_0^h \pi y^2\, \mathrm{d}x.$$

But y/x is the tangent of the semi-vertical angle of the cone, and is equal to a/h, so that $y = ax/h$. Thus, the volume of the cone

$$= \pi \int_0^h \frac{a^2x^2}{h^2}\, \mathrm{d}x = \frac{\pi a^2}{h^2}\left[\frac{x^3}{3}\right]_0^h = \underline{\frac{\pi a^2 h}{3}}.$$

Q 10 *The impedance, Z, of a circuit is given by the expression*

$$\frac{1}{Z} = \frac{1}{R + j\omega L} + j\omega C,$$

where $R = 40$, $L = 2 \times 10^{-3}$, $\omega = 10^4$ *and* $C = 4 \times 10^{-6}$.
(*a*) *Express 1/Z in polar form.*
(*b*) *Find the magnitude and phase angle of the impedance.*

A 10 (*a*) $$\frac{1}{Z} = \frac{1}{R + \mathrm{j}\omega L} + \mathrm{j}\omega C = \frac{R - \mathrm{j}\omega L}{R^2 + \omega^2 L^2} + \mathrm{j}\omega C.$$

Let $Z_{RL}{}^2$ represent the term $R^2 + \omega^2 L^2$.

$$\therefore \frac{1}{Z} = \frac{R - \mathrm{j}\omega L + \mathrm{j}\omega C Z_{RL}{}^2}{Z_{RL}{}^2} = \frac{R + \mathrm{j}\omega(C Z_{RL}{}^2 - L)}{Z_{RL}{}^2}.$$

The magnitude of $1/Z$ is given by

$$\left|\frac{1}{Z}\right| = \sqrt{\frac{R^2 + \omega^2(C Z_{RL}{}^2 - L)^2}{Z_{RL}{}^4}},$$

$$= \sqrt{\frac{R^2 + \omega^2 L^2 + \omega^2 C^2 Z_{RL}{}^4 - 2\omega^2 L C Z_{RL}{}^2}{Z_{RL}{}^4}},$$

$$= \sqrt{\frac{Z_{RL}{}^2(1 + \omega^2 C^2 Z_{RL}{}^2 - 2\omega^2 L C)}{Z_{RL}{}^4}},$$

$$= \sqrt{\frac{1 + \omega^2 C^2(R^2 + \omega^2 L^2) - 2\omega^2 L C}{R^2 + \omega^2 L^2}},$$

$$= \sqrt{\frac{1 - 2\omega^2 L C + \omega^4 L^2 C^2 + \omega^2 C^2 R^2}{R^2 + \omega^2 L^2}},$$

$$= \sqrt{\frac{(1 - \omega^2 L C)^2 + \omega^2 C^2 R^2}{R^2 + \omega^2 L^2}}.$$

The phase angle, θ, of $1/Z$ is given by

$$\tan \theta = \frac{\omega}{R}(CZ_{RL}^2 - L) = \frac{\omega}{R}(CR^2 + \omega^2 CL^2 - L),$$

$$= \frac{\omega}{R}\{CR^2 - L(1 - \omega^2 LC)\}.$$

Hence, in polar form, $1/Z$

$$= \sqrt{\frac{(1-\omega^2 LC)^2 + \omega^2 C^2 R^2}{R^2 + \omega^2 L^2}} \angle \tan^{-1}\frac{\omega}{R}\{CR^2 - L(1 - \omega^2 LC)\}.$$

(*b*) Substituting the given values,

$$\left|\frac{1}{Z}\right| = \sqrt{\frac{(1-0{\cdot}8)^2 + 2{\cdot}56}{1600 + 400}} = 3{\cdot}606 \times 10^{-2},$$

and $$\theta = \tan^{-1}\frac{10^4}{40}\{6{\cdot}4 \times 10^{-3} - 2 \times 10^{-3}(1 - 0{\cdot}8),$$

$$= \tan^{-1}(1{\cdot}6 - 0{\cdot}5 \times 0{\cdot}2) = \tan^{-1} 1{\cdot}5 = 56° 19'.$$

$$\therefore Z = \frac{1}{0{\cdot}03606\angle 56° 19'} = 27{\cdot}73\angle -56° 19'.$$

TELEPHONY C, 1976

Students were expected to answer any 6 questions

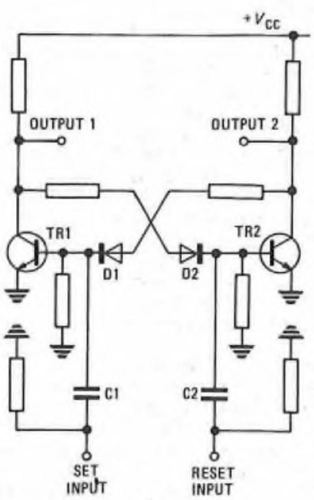
Q 1 (*a*) *Draw a typical transistor bistable circuit element* (*toggle*) *and explain its operation.*

(*b*) *With the aid of a block diagram, explain how such elements can be used to form a digit store suitable for use in a register.*

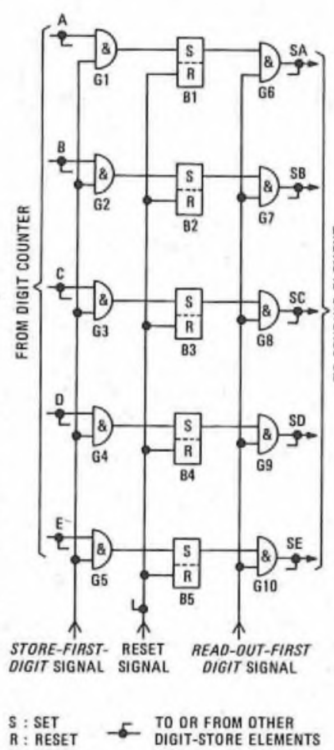
A 1 (*a*) Sketch (*a*) shows a typical transistor bistable circuit. Assuming transistor TR1 to be in its conducting state, then transistor TR2 is non-conducting since the collector of transistor TR1, and hence the base of transistor TR2, is at earth potential. Therefore, as the collector of transistor TR2 is at the positive supply potential, $+V_{CC}$, transistor TR1 is maintained in a stable conducting condition due to its base–emitter junction being forward-biased.

If a negative-going pulse of amplitude V_{CC} is applied to the SET input, capacitor C1 transfers the pulse to the base of transistor TR1, the potential of which falls from just above earth potential to approximately $-V_{CC}$ before rising exponentially back to earth potential. Hence, transistor TR1 turns off and, as its collector potential rises rapidly towards $+V_{CC}$, the base–emitter junction of transistor TR2 becomes forward-biased, and transistor TR2 conducts. The collector potential of transistor TR2 falls from $+V_{CC}$ to approximately earth potential, which then maintains transistor TR1 in the non-conducting state. Also, the potential at output 1 rises from earth to $+V_{CC}$, while that at output 2 falls from $+V_{CC}$ to earth.

To restore the toggle to its original state, it is necessary to apply a negative-going pulse of amplitude V_{CC} to the RESET input. The circuit functions as described above, except that transistor TR2 is now turned off and, in turn, turns on transistor TR1. Thus, the potential at output 1 reverts from $+V_{CC}$ to earth potential, and the potential at output 2 reverts from earth to $+V_{CC}$.



(*a*)



(*b*)

Diodes D1 and D2 raise the base threshold voltages at which the transistors turn on, thus rendering the circuit more immune to noise, and hence making it more stable.

(*b*) Sketch (*b*) shows 5 toggle elements arranged for use as a digit store in a register. For simplicity, only a single digit store is shown; thus, if it were necessary to store, say, 9 digits, the elements shown would need to be repeated 9 times.

To store a digit, a digit counter is needed to assess the value of the digit and then present this value in a 2-out-of-5 code as simultaneous signals on 2 of the leads A–E. To store the digit on toggles B1–B5, a momentary *store-first-digit* signal (as illustrated in sketch (*b*)) opens the respective gates, and the 2 appropriate toggles are set by the outputs from the gates. Typically, if the digit to be stored is 4, which is represented by signals on leads A and D, then when the *store-first-digit* signal appears, gates G1 and G4 open and their outputs set toggles B1 and B4. The digit counter is then reset in readiness for counting the next digit, which is stored in a similar fashion on another set of 5 toggles under the control of a *store-second-digit* signal.

When it is required to examine the value of the digit stored (for example, for digit-sending purposes), a *read-out-first-digit* signal opens gates G6 and G9. Hence, signals are passed over leads SA and SD (representing the value of the stored digit in a 2-out-of-5 code) to the sending element, which is then able to pulse out the digit. The second digit store is interrogated in a similar manner by a *read-out-second-digit* signal.

When the information stored is no longer required, the *reset* signal is applied to reset to normal any toggles that have been set, so that the toggles are available for storing digits on another call.

Q 2 (*a*) *With the aid of a trunking diagram, describe how the following classes of call are set up in a 200-line rural automatic exchange* (*that is, a unit automatic exchange* (*UAX*)):

(*i*) *own-exchange calls,*
(*ii*) *calls to a subscriber on a distant exchange* (*STD calls*), *and*
(*iii*) *calls incoming from an adjacent local exchange.*

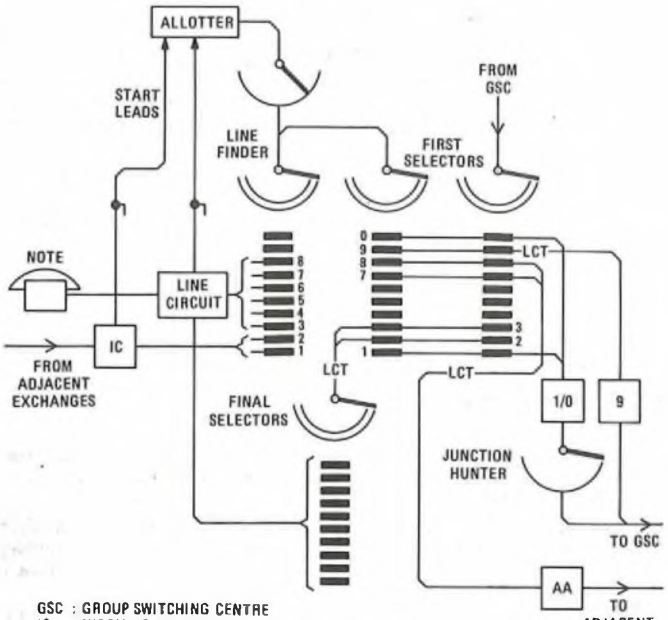
(*b*) *Describe the metering arrangements for call types* (*i*) *and* (*ii*).

A 2 (*a*) The sketch shows a typical trunking diagram for a 200-line UAX. The numbering range shown is 200–399, and access to 2 adjacent local exchanges on levels 7 and 8 of the first selectors is illustrated.

(*i*) When the subscriber originates an own-exchange call, a *start* signal is sent from the line circuit to the allotter, which allocates a line finder and first selector for the call. The line finder finds the calling line and extends the calling subscriber's loop to the first selector, which then returns dial tone. The first dialled digit steps the first selector to level 2 or 3, as appropriate, and a final selector is seized via a local-call timer (LCT). The final selector is positioned by the second and third digits, and ringing tone is returned from it.

(*ii*) The initial part of an STD call is as described in part (*i*) above, except that the first dialled digit steps the first selector to level 0 to seize a level-1-and-0 relay-set. When this relay-set detects the first *break* pulse after the digit 0, the associated junction hunter hunts for a free junction to the group switching centre (GSC). A first selector at the GSC is seized, and the relay-set sends forward the digit 0 to step the GSC first selector and seize a register-access relay-set and controlling register. The level-1-and-0 relay-set then forwards a *class-of-service* signal, in the form of a digit in the range 1–6, to the register-access relay-set at the GSC, and this digit indicates the charging group in which the call originated and whether the call is from an ordinary or coin-collecting-box line. The level-1-and-0 relay-set stores and repeats forward the rest of the dialled digits to enable the controlling register to set up the call through the network.

(*iii*) Incoming calls are set up exactly as described in part (*i*) above, except that the *start* signal to the allotter is originated by the incoming relay-set, and the line finder finds and extends the calling junction to a first selector.



GSC : GROUP SWITCHING CENTRE
IC : INCOMING RELAY-SET
1/0 : LEVEL-1-AND-0 RELAY-SET
9 : LEVEL-9 RELAY-SET
AA : AUTO-TO-AUTO RELAY-SET
LCT : LOCAL-CALL TIMER

Note: Coin-collecting-box lines routed via coin-and-fee-checking relay-sets

(*b*) When the called subscriber answers an own-exchange call, the final selector sends back an initial DC metering pulse to the calling subscriber's line circuit and meter via the LCT. When the LCT detects this initial metering pulse, it commences periodic metering by providing further metering pulses (at the appropriate rate for the time of day) for the duration of the call. It should be noted that the LCT is inoperative on incoming-junction and coin-collecting-box calls.

When the called subscriber answers an STD call, the register-access relay-set at the GSC sends back an initial *metering-over-junction* signal to the level-1-and-0 relay-set, which converts it into a metering pulse to operate the subscriber's meter. Following this, the level-1-and-0 relay-set receives further periodic *metering-over-junction* signals from the GSC at the appropriate rate, and for the duration of the call; these are again converted into metering pulses to operate the subscriber's meter.

Q 3 A traffic level of 1 erlang is offered to a full-availability group of 5 switches. Assuming an Erlang distribution, calculate

(*a*) *the grade of service given,*

(*b*) *the traffic lost, and*

(*c*) *the traffic carried by the second switch.*

A 3 (*a*) For an Erlang distribution, the grade of service, B, is given by

$$B = \frac{\dfrac{A^N}{N!}}{1 + A + \dfrac{A^2}{2!} + \dfrac{A^3}{3!} + \dots \dfrac{A^N}{N!}},$$

where A is the traffic offered (erlangs), and N is the number of circuits.

$$\therefore\ B = \frac{\dfrac{1^5}{120}}{1 + 1 + \dfrac{1^2}{2} + \dfrac{1^3}{6} + \dfrac{1^4}{24} + \dfrac{1^5}{120}} = \frac{\dfrac{1}{120}}{\dfrac{326}{120}},$$

$$= \frac{1}{326} = \underline{0 \cdot 003}.$$

(*b*) The traffic lost = traffic offered × B,

$$= 1 \times 0 \cdot 003 = \underline{0 \cdot 003 \text{ erlang}}.$$

(*c*) The traffic passed on from the first switch

$$= A\left(\frac{A}{1+A}\right) = 1\left(\frac{1}{1+1}\right) = 0 \cdot 5 \text{ erlang}.$$

The traffic passed on from the second switch

$$= A\left(\frac{\dfrac{A^2}{2!}}{1 + A + \dfrac{A^2}{2!}}\right) = 1\left(\frac{\dfrac{1^2}{2}}{1 + 1 + \dfrac{1^2}{2}}\right),$$

$$= 0 \cdot 2 \text{ erlang}.$$

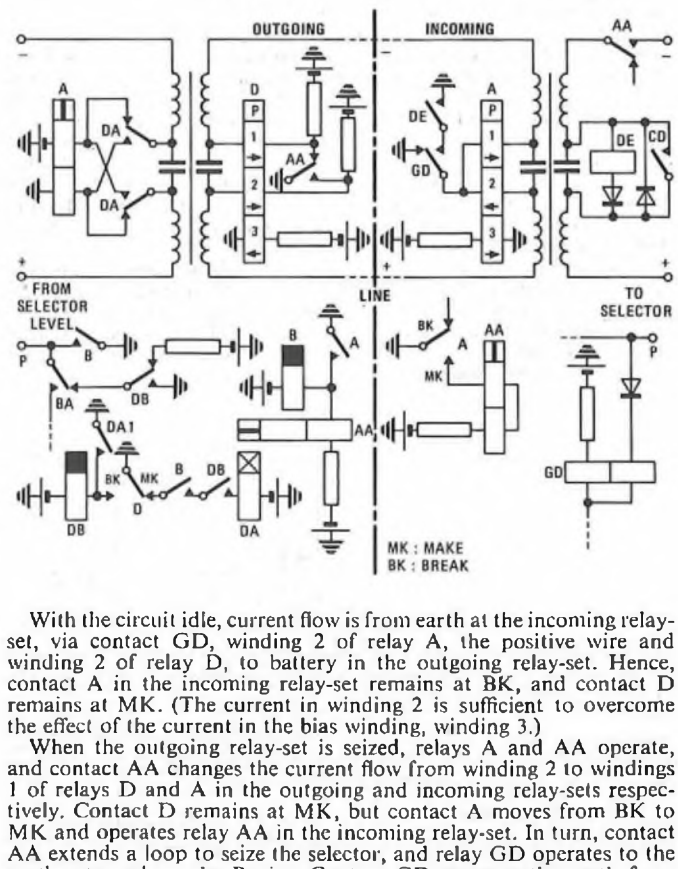
Thus, the traffic carried by the second switch

$$= 0 \cdot 5 - 0 \cdot 2 = \underline{0 \cdot 3 \text{ erlang}}.$$

Q 4 (*a*) *With the aid of sketches of circuit elements, describe the principle of operation of a DC signalling system suitable for pulsing over long lines up to about 160 km (100 miles) in length.*

(*b*) *State with reasons the factors which limit the length of line over which pulses can be satisfactorily received.*

A 4 (*a*) The sketch shows the signalling elements of a unidirectional DC signalling circuit using double-current working. Polarized relays D and A are of the Carpenter type and are side-stable; that is, they remain in the make (MK) or break (BK) position until current flows though the coils in such a direction as to change-over the contact. The windings of each relay are so connected that conventional current flow (earth to negative battery) in the direction of the arrow moves the contact to the MK position. If conventional current flow is in the opposite direction, the contact moves to the BK position.



With the circuit idle, current flow is from earth at the incoming relay-set, via contact GD, winding 2 of relay A, the positive wire and winding 2 of relay D, to battery in the outgoing relay-set. Hence, contact A in the incoming relay-set remains at BK, and contact D remains at MK. (The current in winding 2 is sufficient to overcome the effect of the current in the bias winding, winding 3.)

When the outgoing relay-set is seized, relays A and AA operate, and contact AA changes the current flow from winding 2 to windings 1 of relays D and A in the outgoing and incoming relay-sets respectively. Contact D remains at MK, but contact A moves from BK to MK and operates relay AA in the incoming relay-set. In turn, contact AA extends a loop to seize the selector, and relay GD operates to the earth returned on the P-wire. Contact GD removes the earth from relay A and current now flows round the loop via the negative and positive wires and windings 1 and 2 of relays A and D. Contact A remains in the MK position, but contact D moves to the BK position as relay D comes under control of bias winding 3, windings 1 and 2 being in opposition and effectively cancelling each other. Contact D operates relay DB.

When pulsing takes place, contact AA in the outgoing relay-set moves to its normal position during each *break* pulse, thus reversing the direction of current in the loop, and moving contact A in the incoming relay-set to the BK position to release relay AA during each *break* pulse. Contact AA in the incoming relay-set repeats the *break* pulses to the selector.

When the called subscriber answers, relay DE operates in the incoming relay-set, and contact DE reconnects earth to windings 1 and 2 of relay A. Current then flows in the negative wire only, and contact A remains in the MK position, while contact D moves to the MK position. This operates relay DA, and contact DA1 holds relay DB. Further contacts of relay DA repeat the appropriate backward supervisory signals.

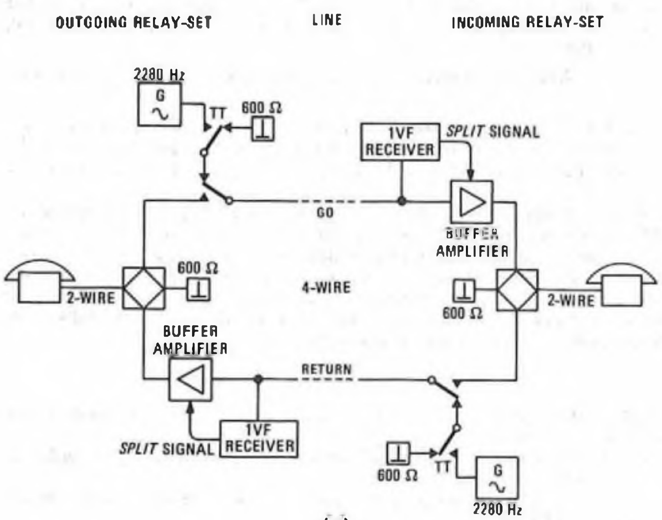
A backward-busy facility is provided since, if continuity of the line is broken while the circuit is idle, or relay GD operates to an earth on the P-wire, current in windings 1 and 2 of relay D either cease to flow or cancel each other, thus putting relay D under the influence of bias winding 3 so that contact D moves to the BK position. Relay DB operates, busying the circuit by putting an earth on the outgoing relay-set's P-wire.

(*b*) The pulsing limits placed on the system depend upon both the resistance and capacitance of the line. If the line resistance is too high, the current through the Carpenter relay windings is insufficient to (*i*) ensure satisfactory operation, and (*ii*) overcome the effect of current in the bias winding. If the line capacitance is too high, the time taken to discharge the line and charge it in the opposite direction becomes excessive and introduces unacceptable dial-pulse distortion. The pulsing limit for 50 V exchanges is 76 000 μF Ω, and the DC signalling limit is 8·2 kΩ. The line insulation resistance should be greater than 0·1 MΩ to ensure reliable working.

Q 5 *(a) With the aid of a block diagram of a voice-frequency signalling system, explain the function of the buffer amplifier.*

(b) Explain the principle employed to prevent misoperation of the receiver by far-end speech.

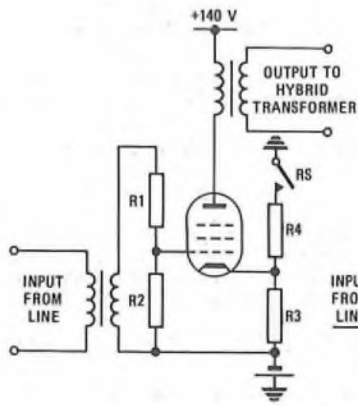
A 5 (*a*) Sketch (*a*) shows a block diagram of a typical single-voice-frequency (1VF) signalling system. The signalling frequency, 2280 Hz, is used to transmit signalling information in both directions, and the 1VF receivers convert this signalling information into DC signals to operate the exchange switching equipment as necessary. Typically, 10 pulses/s loop-disconnect signalling into the 2-wire side of the outgoing relay-set is converted to 10 pulses/s bursts of 2280 Hz tone by relay TT. At the incoming relay-set, the 1VF receiver converts the bursts of 2280 Hz tone back into 10 pulses/s loop-disconnect signalling on the outgoing 2-wire side. Signals in the reverse direction are sent and received in a similar manner, except that here the signals indicate, typically, *called-subscriber-answer* or other supervisory signals.



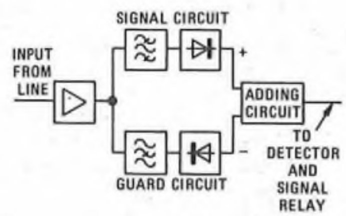
(*a*)

The buffer amplifiers are not required for amplification, but serve to prevent spill-over of 2280 Hz signalling tone into adjacent links via the normal 2-wire speech path; spill-over is prevented by splitting the line at the buffer amplifier. For instance, if the outgoing relay-set transmits 2280 Hz to the incoming relay-set, then the splitting signal from the 1VF receiver effectively cuts off the buffer amplifier. This prevents the tone going through the hybrid transformer and into, possibly, another similar signalling system in tandem, where the tone would be unwanted and may be recognized and give rise to false operation. The buffer amplifier also effectively isolates the associated 1VF receiver from the local 2-wire circuit, preventing false operation of the receiver by near-end interference, speech or noise.

Sketch (*b*) shows the basic principles of a buffer amplifier circuit. Resistors R1 and R2 form a potential divider such that, allowing for amplification through the tube, the overall gain from input to output is unity. Resistor R3 gives the normal automatic bias and negative feedback so that, typically, the grid bias is -2 V with respect to the cathode. When a 2280 Hz signal is received, relay RS operates and contact RS effectively increases the cathode potential from -48 V to, typically, -30 V by the potential-divider action of resistors R3 and R4. Thus, the grid bias becomes -20 V with respect to the cathode, and the tube is cut off; that is, signals on the input cannot get through to the output.



(*b*)



(*c*)

(*b*) Sketch (*c*) shows a block diagram of a typical 1VF receiver. There are 2 distinct elements: a signal circuit and a guard circuit. For genuine 2280 Hz signals, the signal circuit, which contains a 2280 Hz band-pass filter, responds by giving a strong positive output to the adding circuit to operate the signal relay via an amplifier. The guard circuit responds weakly to 2280 Hz signals, and its negative output to the adding circuit is ineffective. However, when speech appears on the input, the signal circuit responds weakly and its positive output is small; the guard circuit, however, responds more strongly to speech frequencies and gives a strong negative signal to the adding circuit to mask any positive output from the signal circuit, and thus prevents the signal relay from operating to any 2280 Hz frequencies that may occur in the speech.

Q 6 *An operator at an auto-manual centre is provided with a means of timing calls.*

(a) For a typical timing device, describe, with the aid of sketches of circuit elements, how

(i) the caller is given an indication of the elapsed time, and
(ii) the elapsed time is indicated to the operator.

(b) State 2 of the other principal facilities provided by the timing device, and outline how one of these facilities is provided.

A 6 (*a*) See A5, Telephony C, 1974, *Supplement*, Vol. 68, p. 65, Oct. 1975.

(*b*) Two other principal facilities provided by a chargeable-time clock are

(*i*) an indication of the actual time for which a call was effective, given when the calling subscriber clears down, and
(*ii*) an indication to the operator that 9 min (2·8 min on coin-collecting-box calls) have elapsed.

Descriptions of how each facility is provided are given in the above reference.

Q 7 *(a) Draw a trunking diagram for a TXE2 exchange, distinguishing clearly between speech and control paths.*

(b) With the aid of sketches, explain the use of the B-switch selector in the system.

(c) Assuming the register has received all the required digits, describe the setting-up sequence to the called subscriber's line.

A 7 (*a*) See A6, Telephony C, 1974, *Supplement*, Vol. 68, p. 66, Oct. 1975.

(*b*) and (*c*) See A10, Telephony C, 1975, *Supplement*, Vol. 69, p. 75, Oct. 1976.

Q 8 (a) Draw a trunking diagram to show the position of the controlling register-translator in an STD network.
(b) State the factors which influence the choice of the location.
(c) List the principal facilities of a controlling register-translator.
(d) Why is it usual to separate the register and translator functions?

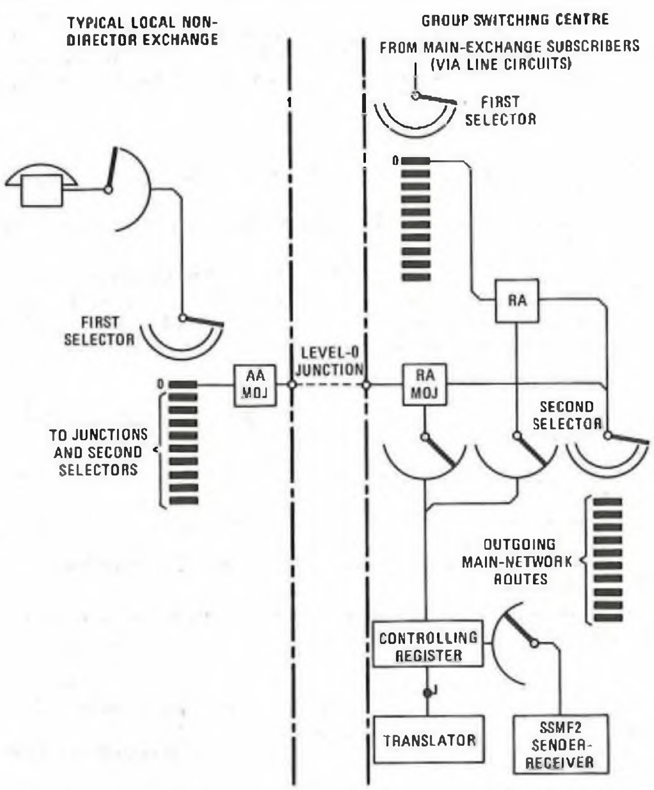
A 8 (*a*) The sketch shows the position of the controlling register-translator in a typical non-director-area system. Access to the register-access relay-set and controlling register is gained from level 0 of the first selectors in the main exchange or, typically, from level 0 of the first selectors in a dependent local exchange via a level-0 junction.

(*b*) Large amounts of STD equipment would be needed if such equipment were provided at each local exchange. It is more economic to provide the equipment at a centralized point. The group switching centre (GSC) is chosen because

(*i*) the exchange building has to be big enough to accommodate the STD equipment,
(*ii*) the GSC is provided with access to other main-network switching centres,
(*iii*) the GSC is near the centre of the charging group, so that junction routes from the dependent exchanges are as short as possible,
(*iv*) the GSC has the largest number of subscribers within the charging group, so that the number of metering-over-junction equipments is kept to a minimum, and
(*v*) the GSC has the largest number of business subscribers, who generate considerable STD traffic.

(*c*) A non-director controlling register-translator has facilities for

(*i*) storing up to 9 digits,
(*ii*) examining up to the first 5 digits for translation purposes,
(*iii*) sending a fee digit to the register-access relay-set to specify the charging rate for the call,
(*iv*) sending up to 6 translation digits, followed by the appropriate repeated digits,
(*v*) determining which of the stored digits should not be repeated,
(*vi*) changing the fee digit if necessary (and hence the charging rate for the call) on selected national-numbering-group codes should the call originate from a dependent charging group,
(*vii*) force-releasing the forward equipment and causing the return of number-unobtainable tone if a spare or barred code is received, or if there is excessive delay during the dialling of the first 7 digits,



AA: Auto-to-auto relay-set
MOJ: Metering-over-junction facilities
RA: Register-access relay-set
SSMF2: Signalling System Multi-Frequency No. 2

(*viii*) applying a 4 s time-out period after the seventh (and eighth, if received) digit, such that, if no further digits are received in that time, dialling is assumed to have finished,
(*ix*) delaying the sending of the penultimate digit until the final digit has been stored,
(*x*) associating a Signalling System Multi-Frequency No. 2 sender-receiver, if required, and
(*xi*) setting-up the transmission path through the register-access relay-set and disconnecting itself from the call when sending is complete.

(*d*) During the setting-up of a call by a register, the actual process of examining the initial digits dialled, and obtaining a translation for them, does not take more than about 0·5 s. Thus, if each register had its own integral translator, the translator portion would be used, typically, for only 2·5% of the time, 20 s being the average controlling-register holding time. As translators are expensive items of equipment, it is better to share them among a group of registers so that they are used more economically. This also means less manpower is required when changing translation strappings. Typically, one translator may be shared by 20 or more registers, and deals with around 5 register demands per second. A translator may be associated with a register up to 3 times during the setting-up of a call, as not all the translation information can necessarily be given on the first, or even the second, demand. Each association lasts for approximately 150 ms.

Q 9 (a) For what reason are trunk exchanges provided with access to a trunk transit network?
(b) Describe how a call is set up over the transit network.
(c) Approximately what proportion of the total trunk traffic would you expect to be carried by such a network? Give reasons for your answer.

A 9 (*a*) Trunk exchanges are provided with access to a trunk transit network so that STD calls that cannot be routed over the normal trunk network (due to insufficient routing digits being available from the controlling register, or to more than 2 links in tandem being required) can be set up over the transit network using a fast inter-register signalling system. The transit network, when fully implemented, will enable all group switching centres (GSCs) to be connected to each other via up to 5 transit links in tandem. It also allows GSCs to have access to international switching centres (ISCs) if direct GSC–ISC circuits are not provided. The transit network can also be used for alternative routing between GSCs, if so desired, when the normal trunk route is busy.

(*b*) See A10, Telephony C, 1973, *Supplement*, Vol. 67, p. 73, Jan. 1975.

(*c*) For a call routed from one GSC to another, the most economic routing is via a direct link provided there is sufficient traffic to justify a direct link; if an intermediate switching point is introduced, the cost rises.
As the transit network involves the use of at least one intermediate switching point (a TSC) on any call (with an average of 2–3 TSCs being used on any call), then the traffic routed over the transit network should be as small as possible. As the amount of traffic justifies normal 1-link or 2-link trunk routings between GSCs for about 94% of the total main-network traffic, it is only the residual 6% of traffic that needs to be routed via the transit network.

Q 10 (a) Outline the maintenance techniques used in step-by-step and crossbar exchanges.
(b) What features of the switching system influence the choice of maintenance technique?
(c) Which system would you expect to give a better service to the customer and why?

CORRECTION

TELEPHONY C, 1970 (*Supplement*, Vol. 64, July 1971)

A 4 The value of the expression

$$P(10) = \frac{32 \cdot 5^{10}}{2 \cdot 718^{32 \cdot 5} \times 10!}$$

is wrongly given as $3 \cdot 6 \times 10^{-5}$. The correct answer is $2 \cdot 782 \times 10^{-6}$.

TELEGRAPHY C, 1976

Students were expected to answer any 6 questions

Q 1 (a) What are the principal causes of telegraph signal distortion

(i) between a subscriber's teleprinter and a Telex exchange when the circuit is physically routed in a telephone cable, and

(ii) between 2 terminals of a circuit in a multi-channel voice-frequency system?

(b) Describe the equipment which would be used at a test desk to measure the distortion.

A 1 (*a*) A telegraph signal is said to suffer distortion when the received signal elements do not have exactly their theoretical durations.

(*i*) Telegraph signals between a subscriber's station and a Telex exchange are normally DC signals, and the principal causes of distortion when the circuit is routed over a telephone cable are listed below.

Line Capacitance The receiver current rises gradually as the capacitance of the line is charged, and this delays the response of the receiver to a change in transmitter polarity.

Filter Capacitance and Inductance A low-pass filter is provided to prevent interference to adjacent telephone circuits by the transmission of unwanted frequencies to line. The capacitance and inductance of the filter each slow the rise in line current and have a similar effect to the line capacitance.

Line Resistance The resistance of the line causes a reduction in the current available to operate the receiver; excessive resistance would prevent the current reaching a value sufficient to operate the receiver, and the character would be mutilated.

Noise Induced currents from neighbouring pairs in the cable could distort signals by aiding or opposing the line current; the induced currents are called *noise*.

(*ii*) Between 2 terminals of a voice-frequency circuit, the signals are carried by alternating currents in the speech-frequency band. The principal types of distortion and their causes are listed below.

Characteristic Distortion Characteristic distortion is the distortion inherent in the system, and is a measure of the electrical characteristics of the circuit and equipment; it is constant for any signal element or combination of elements. Characteristic distortion is measured when the system is in perfect working order, with perfect signals applied, and with the circuit free from bias and interference from any other source.

Bias Distortion Bias distortion is the consistent lengthening or shortening of elements of one polarity, with a corresponding shortening or lengthening of signals of the opposite polarity. Bias distortion can be caused by unequal signal voltages, earth currents, hysteresis effects or a badly adjusted relay. It may be corrected by adjusting the equipment until the received signals have neutral bias.

Fortuitous Distortion Fortuitous distortion is caused by random signals which affect the circuit or equipment. The signals arise from currents induced by neighbouring pairs in the same cable, or adjacent power wires, or are parasitic currents arising from equipment irregularities. Fortuitous interference is known as *crosstalk* or *noise*.

(*b*) See A7, Telegraphy B, 1972, *Supplement*, Vol. 66, p. 28, July 1973.

Q 2 (a) A local call has been connected through an automatic Telex exchange.

(i) Draw a diagram to show the main items of equipment in the circuit, both outside and inside the exchange.

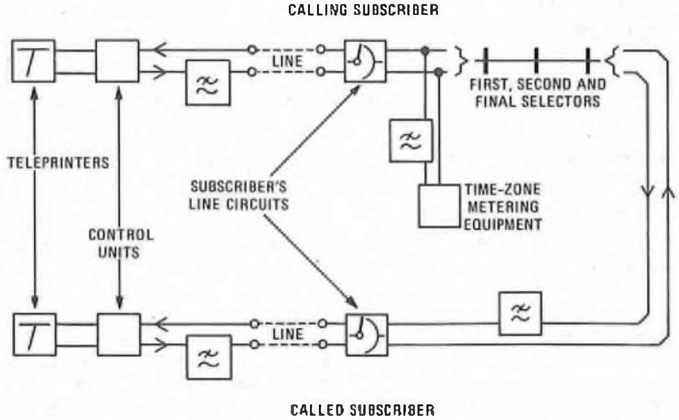
(ii) Show on the diagram where low-pass filters would be connected.

(b) With the aid of a diagram of a filter circuit and an attenuation/frequency graph, explain the purpose of the low-pass filter.

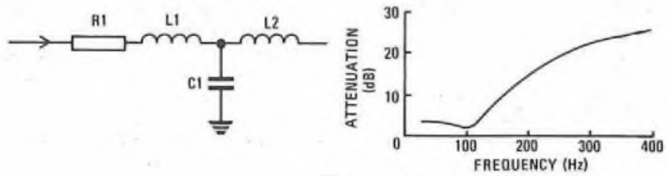
A 2 (*a*) Sketch (*a*) shows the main items of equipment in a Telex call and the location of the low-pass filters.

(*b*) Sketch (*b*) shows the components of the type of low-pass filter used in a Telex call, and the attenuation/frequency graph of such a filter. Pairs of wires routed in the same cable are affected by capacitive coupling, and energy from audio-frequency harmonics of a square-wave teleprinter signal can be transferred to a neighbouring telephone circuit, causing interference. As a square-wave telegraph signal contains all the odd harmonics to infinity, a low-pass filter, severely attenuating frequencies above about 140 Hz, would prevent crosstalk with other circuits and allow sufficient of the DC signal to pass unattenuated to line. For a 50 baud signal, the fundamental frequency is 25 Hz, the third harmonic is 75 Hz and the fifth harmonic is 125 Hz. The fundamental wave plus the 2 harmonics mentioned gives a reasonable signal.

The nominal values of the components are: $L_1 = L_2 = 1{\cdot}3$ H, $C_1 = 2\ \mu$F, and $R_1 = 200\ \Omega$. The design impedance is $1{\cdot}14$ kΩ,



(*a*)



(*b*)

which is an average match to the various types of line plant in use. The resistor is provided to damp oscillations that can occur due to mismatch with the modulator. The impedance of the modulator source is 100 Ω (the cold resistance of the current-limiting barretter), and oscillations can occur at each change of signal polarity, sufficiently severe to mutilate the signal.

Q 3 (a) A multi-channel voice-frequency (MCVF) telegraph system is to be brought into service on a circuit previously used for telephony. List the checks and the tests that should be performed on

(i) the bearer circuit, and

(ii) the MCVF equipment.

(b) When the system is in use, what regular tests should be applied to ensure that an adequate quality of service is maintained?

A 3 (*a*) (*i*) When a 4-wire speech circuit is used for an MCVF system, the circuit is split into two 2-wire channels to provide the GO and RETURN circuits, and any bridging equipment is removed. The voice-frequency signalling equipment is disconnected and, at each end of the circuit, the 2-wire-to-4-wire hybrid termination is detached. Echo suppressors are removed.

To ensure that the performance of all telegraph channels is similar, the circuits are adjusted so that the attenuation at any frequency does not vary from the attenuation at 800 Hz by more than 1 dB. The frequency drift must also be low, and a check is made of noise and intermodulation characteristics.

(*ii*) For the MCVF equipment, the general condition of the apparatus and wiring is observed, and a check made of fuses, supply voltages and alarms. The telegraph relay is checked, cleaned, adjusted and operated to each side by current from the distant terminal; this also ensures that the received signal is correctly filtered, amplified, rectified and applied to the coils of the relay. The gain compensation at each station is checked, adjusted and set at the mid-point position.

Reversal signals of *mark* : *space* ratios 1 : 1, 2 : 2 and 4 : 4, and mixed signals, are sent to and from each terminal, and the distortion is measured. The 2 : 2 signals at 50 baud are used to set the bias to the optimum value, and the mixed signals are not expected to require more than a small readjustment. The channel interference is checked in local operation; the 4-wire circuit is checked in local operation and with the distant terminal. The total distortion, including fortuitous and characteristic distortion, should not exceed 8% with mixed signals.

The mains voltages and supply voltages are then varied within the upper and lower limits and the tests repeated.

(*b*) When the system is in service, a 4-weekly check of alarms and voltage levels is made. At yearly intervals, the channel panels are checked, the relay is checked and cleaned, and the channel is checked in local operation. Correct operation of the SEND-wire guard relay is also verified. The system is lined-up from end to end and the distortion measured using 2 : 2 and mixed signals. The common equipment is functionally tested also at yearly intervals. Plugs and cords are checked every 3 months.

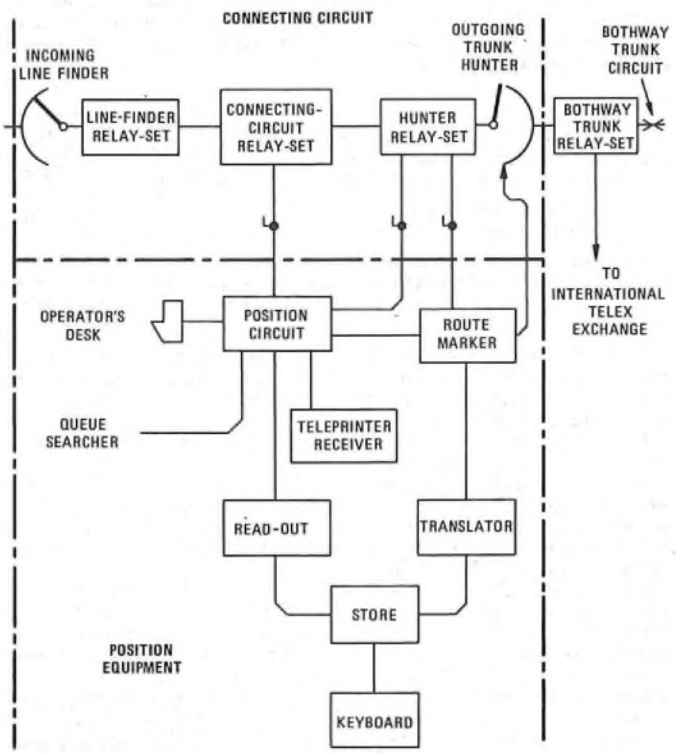
Q 4 *(a) With the aid of a block diagram, explain how an outgoing trunk circuit may be selected from a Telex cordless switchboard.*

(b) When a circuit has been seized, how are selection signals transmitted to the distant terminal for an exchange requiring

(i) dial pulses, and
(ii) teleprinter signals?

(c) For a call including a radio link, how may a subscriber recall the operator?

A 4 (*a*) The sketch shows the position equipment and a connecting circuit of a Telex cordless switchboard; several connecting circuits are provided for each position. When an incoming call has been accepted by the operator, and the destination determined, the SEND CALL key is operated and the character *figure shift*, followed by 2 digits, are transmitted from the keyboard, which is located at the operator's desk. The 2 digits represent a code appropriate to the outgoing route required, and are used as follows to locate the route on the banks of the outgoing trunk hunter. The 2 route-selection digits are selected by operating the keyboard; simultaneous 5-unit signals from the keyboard are passed to the store and to the translator, where the 5-unit signals are converted to dial-pulse signals. The function of the store is to retain the keyboard signals while the translator is converting the signals to dial pulses. The dial pulses operate the route marker and mark the banks of the outgoing trunk hunter; the hunter rotates until the marked route is located, and steps contact by contact over the route until a free circuit is found. The circuit is seized and, when the *call-connected* signal is received, the selection digits for the called subscriber are transmitted.



(*b*) (*i*) Simultaneous signals on the 5-wire output of the keyboard are not suitable for transmission over a trunk circuit where the distant exchange expects dial pulses. The 2 route-selection digits cause the store and translator to be retained in circuit so that the subscriber-selection digits, and further routing digits, are converted to dial pulses in a similar manner to the route-selection digits. The dial pulses are routed through the operator's position circuit and the outgoing-trunk-hunter relay-set to the trunk circuit.

(*ii*) When the 2 route-selection digits indicate that the distant terminal expects teleprinter signals as selection signals, the store and translator are not used. The 5-wire signals from the keyboard are connected to the read-out unit, which converts the signals to sequential teleprinter signals, and these are transmitted over the trunk circuit.

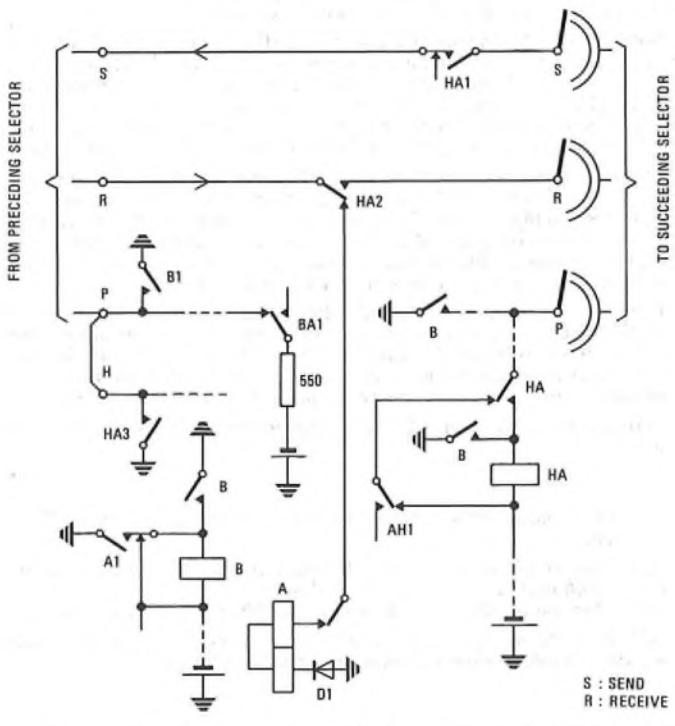
(*c*) At any time, either subscriber on a radio call can recall the operator by transmitting the *operator-recall* signal. This consists of a *line-feed* signal followed by 4 *carriage-return* signals. The *carriage-return* signals cause the THROUGH lamp to flash on the appropriate circuit. The *line-feed* signal prevents any subsequent signals being superimposed on text already received.

Q 5 *(a) With the aid of a circuit diagram, explain how, for a local call, a circuit between 2 ranks of group selectors in an automatic Telex exchange may be*

(i) seized,
(ii) held, and
(iii) released.

(b) Explain how the operation differs for a trunk circuit between 2 exchanges.

A 5 (*a*) (*i*) The sketch shows the circuit of a group selector. If the selector is free, a 550 Ω path to negative potential is applied to the P-wire of the preceding selector at contact BA1. The selector is seized by the preceding selector operating to this condition and connecting through the SEND and RECEIVE wires. Negative potential on the RECEIVE wire operates relay A through rectifier D1. Contact A1 operates relay B, and earth potential is applied to the P-wire from contact B1; another contact of relay B operates relay BA (not shown), which disconnects the 550 Ω testing potential from the P-wire. The earth potential on the P-wire prevents seizure by any other selector, and holds the preceding selector or selectors. Thus, the circuit between the 2 selectors is seized and protected from seizure by any other selector.



(*ii*) The selector steps vertically under the control of the calling subscriber's dial, and hunts for a free outlet with a rotary motion. Relay AH (not shown) operates to the 550 Ω testing potential on the P-wire, and contact AH1 removes the short-circuit from relay HA, which operates. Contacts HA1 and HA2 connect the SEND and RECEIVE wires to seize the next selector; the negative potential on the RECEIVE wire causes earth potential to be returned on the P-wire to hold relay HA when relays A and B release (due to the disconnexion at contact HA2). Contact HA3 holds the preceding selector. The group selector, the preceding selector and the circuit between them are thus held by earth potential on the P-wire from the succeeding equipment. The purpose of the strap between terminals P and H of the connexion to the preceding selector is that, if the preceding selector is a group selector, it is held by earth potential on the P-wire from contact HA3; if the preceding selector is a time-zone equipment, the holding earth is needed on the H-wire only, the strap being omitted.

(*iii*) The *clear* signal causes the removal of earth potential from the P-wire of the last selector in the train (for example, the final selector), which releases relay HA. Earth potential is disconnected at contact HA3, releasing the preceding selector and the circuit between the 2 selectors.

(*b*) When a trunk is provided between 2 exchanges, the circuit uses 2 wires in place of the 3 wires between selectors. This is to effect an economy in the use of costly long-distance circuits. A trunk relay-set is needed to convert the SEND, RECEIVE and P-wires in the exchange to the SEND and RECEIVE wires of the trunk. The relay-set is seized and holds the preceding group selector in the manner described above. Negative potential is connected to the SEND wire to busy the distant trunk relay-set, which returns a call-confirmation pulse of negative potential. The distant relay-set seizes a group selector, guards the distant P-wire and prepares to receive the incoming selection signals. Both relay-sets are held by negative potential on their RECEIVE wires during the call. (Bursts of positive potential during teleprinter signals are of insufficient length to release the supervisory relays.) When either terminal clears by connecting positive potential to the respective SEND wire, the holding earth is disconnected from the P-wires at each trunk relay-set, and the train of selectors at each end is released.

Q 6 (*a*) *What are the advantages and disadvantages of a fully-automatic routiner, as compared with a semi-automatic routine tester, for testing 2-motion selectors in a Telex exchange?*

(*b*) *With the aid of a block diagram of the principal elements of an automatic routiner installation (including ancillary equipment), explain the operation of the routiner.*

(*c*) *What factors would influence the decision whether to install a fully-automatic routiner in an exchange?*

Q 7 (*a*) *Explain the difference between a telegraph repeating relay and a telegraph regenerative repeater.*

(*b*) *Under what circumstances would each be used?*

(*c*) *With the aid of a block diagram, explain the operation of a regenerative repeater.*

(*d*) *If the selection period of a regenerative repeater is 0·1 ms, what is the margin of the repeater to 50 baud signals?*

A 7 (*a*), (*b*) and (*c*) See A2, Telegraphy C, 1973, *Supplement*, Vol. 67, p. 78, Jan. 1975.

(*d*) The selection period is 0·1 ms in an ideal element length of 20 ms. Assuming the selection period is at the centre of each element, the margin to early or late signals is 9·95 ms. Expressed as a percentage, the margin

$$= \frac{9 \cdot 95}{20} \times 100\% = \underline{49 \cdot 75\%}.$$

Q 8 (*a*) *Draw a block diagram of a fully-automatic message-relay switching centre.*

(*b*) *Explain the operation of the centre.*

(*c*) *What methods are used to deal with*

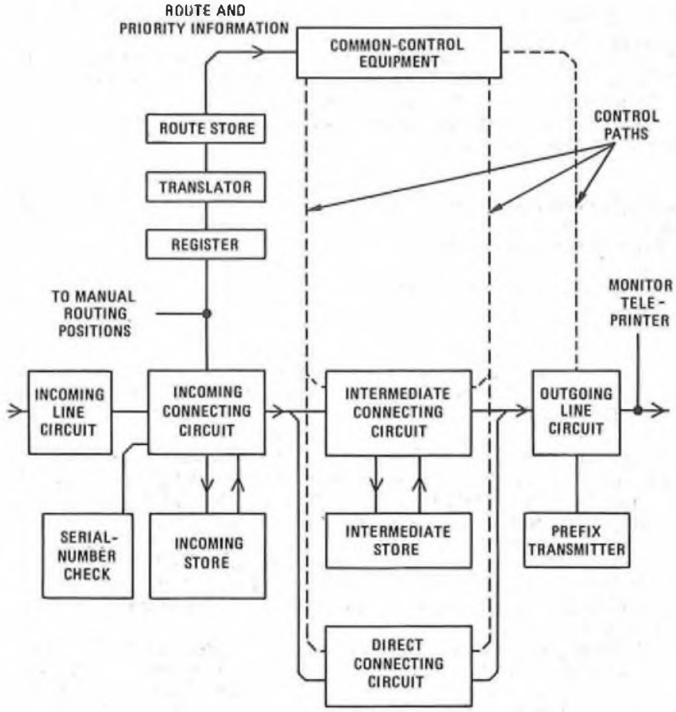
(*i*) *a lost message, and*

(*ii*) *a mutilated message?*

(*d*) *State a typical format for an incoming message, and explain the purpose of each section.*

A 8 (*a*) and (*b*) The sketch shows a block diagram of a fully-automatic message-relay centre. The incoming message is prepared at the distant station to conform with a fairly rigid format in order that the relay-centre equipment may easily recognize *start-of-message*, *serial-number*, *priority*, *destination-address* and *end-of-message* codes. Each message received is stored temporarily in the incoming store so that the destination indicator can be examined and the serial number checked; the channel remains connected to the store until the *end-of-message* sequence is received.

The routing and priority information is decoded by the register and translator and held in the route store until the common-control equipment is ready to process the call. A message destined for a route having one or more free channels is transmitted from the incoming store to the outgoing line circuit via a direct connecting circuit; the outgoing line circuit and the connecting circuit are allocated by the common-control equipment. If all the outgoing circuits on the required route are busy, an intermediate connecting circuit is used, and the message is held in the intermediate store. When the outgoing circuit becomes free, the common-control equipment selects the waiting message with the highest priority and connects the store to the line. Before any message is transmitted to line, a prefix generator



is connected to transmit a new *start-of-message* prefix together with a serial number appropriate to the outgoing channel.

(*c*) Three features of the system protect messages from irretrievable loss. The first is the use of incoming-channel sequential message-numbering equipment, which checks that the serial number carried as a prefix to each message is one integer greater than that of the previous message received on that channel. This enables missing messages to be identified and a request made to the originating station for the message to be retransmitted. Messages are safeguarded in transit through the system by measuring the delay time in storage, and by periodically checking the transmission paths through the system. Alarms are provided to indicate the faulty equipment, and any message in store is routed to the manual positions. The third safeguard is the provision of a monitor teleprinter for each outgoing circuit. Each outgoing message is recorded so that a lost or mutilated message can be retransmitted if requested by the distant terminal.

(*d*) A typical message format, to CCITT recommendations for an automatically-switched system handling telegrams is shown in the chart.

ZCZC GEB099 WY79
GBLD HL URWA 013
WASHINGTON 13/12 13 1205
3
LT
-MIDBANK LONDON-
3
TEXT
3
SIGNATURE
10
NNNN
10

Note: The figures between groups indicate the number of *line-feed* characters transmitted

The first line, known as the *numbering line*, contains the *start-of-message* signal, consisting of the characters ZCZC, the channel-sequence number, consisting of 3 letters identifying the channel and 3 numerals constituting the serial number, together with up to 12 characters giving the telegram identification group.

The second line is known as the *pilot line*, and contains the destination indicator, which consists of 4 characters, the first pair indicating the country of destination and the second pair characterizing the destination office in that country. These are followed by 2 characters indicating the priority and class of traffic of the message, 4 characters indicating the country and office of origin, and a 3-figure number giving the number of chargeable words.

The third line contains the preamble to the message. This may vary according to the requirements of the originator but, in general, the preamble consists of the office of origin in plain language (as opposed to the 4-letter code in the second line, which is used for automatic routing), the time and date of origin, the number of words in the message and the number of chargeable words.

The fourth line contains the address. The character – (corresponding to Combination No. 1, upper case) is inserted at the beginning and end of the address for easy recognition.

The text of the message follows, and then the signature and *end-of-message* signal, NNNN.

Q 9 Explain briefly the following terms as applied to high-frequency radio circuits:
(a) multi-path propagation,
(b) diversity working,
(c) error-detecting code, and
(d) error-correcting system.

A 9 (*a*) Long-distance radio communication at high frequencies depends on the existence of the ionosphere, which is concentric with the earth at an altitude of about 100 km. High-frequency radio waves are reflected from the ionosphere, or are refracted within it; the amount of reflection or refraction depends upon the frequency, the angle of incidence and the state of ionization, which may vary with sunlight or with the time of day. Over great distances, a receiver may detect multiple reflections of the same signal; one reflected signal may arrive earlier over one path than a reflection of the same signal over a second path, with differences of up to several milliseconds being possible. This is known as *multi-path propagation*, and the effect is to lengthen or shorten the signal element, possibly to the extent of mutilating the character.

(*b*) Diversity working is adopted to counter the effects of fading. The amount of fading over a particular path is unpredictable, and signals received some distance away on a second receiver may be stronger and perhaps not subject to fading. Diversity working uses 2 received signals from separate receiving aerials, or from dual-frequency transmissions, and the better signal is selected for onward transmission.

(*c*) An error-detecting code is a code which contains a high degree of redundancy; that is, the number of elements exceeds that required for transmission of the characters in the code. For example, the International Alphabet No. 2 uses a 5-unit code, giving $2^5 = 32$ possible combinations, all of which, except for Character No. 32 (all *space* elements), are used for characters. Should any element be transposed, such that a *mark* element is received in place of a *space* element, a different character is printed because the received signal must represent one of the characters in use. To overcome this, protective codes, called *error-detecting codes* are used; the receipt of an error is detected by the receiving equipment, and results in the printing of a special error sign or the initiation of retransmission. The error-detecting codes are chosen so that the effect of the more common forms of mutilation produce an unusual combination at the receiver which can be recognized as an error. A 6-unit code can be constructed to give $2^6 = 64$ combinations, which means that 32 combinations are not used; a mutilation resulting in the reception of a redundant combination is treated as an error. Alternatively, the sixth unit can be added to each combination as a parity bit, so that all combinations have an even (or odd) number of *mark* or *space* elements.

An improvement can be obtained by the use of a 7-unit code and, of the various forms which this may take, the code using 4 *space* elements and 3 *mark* elements is most favoured. The total number of usable combinations

$$= \frac{7!}{(7-3)!\,(7-4)!} = 35.$$

As the total number of combinations is $2^7 = 128$, a redundancy of 93 combinations results, and this gives good protection against mutilation. Error-detecting codes have the disadvantage that they need more signal elements than does the basic code for each combination and, therefore, either reduce the speed of signalling or require a higher-speed channel.

(*d*) An error-correcting system uses master-and-slave working between the home and distant stations. Transmission normally uses the 7-unit 4-plus-3 code described above, with synchronous transmission. The detection of an error by the distant equipment causes a *retransmit* signal (RQ) to be sent to the home station, which ceases normal transmission and retransmits the last 4 characters sent. If these are received correctly, normal transmission is resumed but, if an error is again received, the distant station retransmits the RQ signal and the cycle is repeated. Cycling continues until conditions improve and reception is normal. At the home terminal, 2 stores are needed: one for the 4 characters needed for error correction, and another for characters which may be received from a Telex circuit while cycling is in progress.

Q 10 (a) Draw a block diagram and explain the operation of equipment suitable for the automatic ticketing of international Telex calls.
(b) What information regarding the setting-up of the call is recorded on the calling-subscriber's teleprinter?
(c) What information is recorded in the exchange to form the basis of the charge for the call?
(d) Why is automatic ticketing provided in preference to operating the subscriber's meter?

A 10 (*a*) and (*d*) See A4, Telegraphy C, 1972, *Supplement*, Vol. 66, p. 55, Oct. 1973.

(*b*) The information recorded on the subscriber's teleprinter regarding the setting-up of a call is

(*i*) the reference code of the call (this enables the exchange equipment used for the connexion to be identified),

(*ii*) the time (this information is the same as recorded in the head sequence on the ticketing machine),

(*iii*) a local record of the answer-back code (this information is requested by the ticketing equipment to identify the caller),

(*iv*) the *proceed-to-select* signal, KEY+ (this signal is sent for 2 reasons: firstly, to request the subscriber to transmit the selection digits and, secondly, as a reminder to use the keyboard and not the dial), and

(*v*) a local record of the selection digits transmitted.

(*c*) The information recorded in the exchange is shown in the table.

Head Sequence	Tail Sequence
Entry-separation characters (4 *carriage returns*)	Entry separation characters (4 *carriage returns*)
Head marker (*letter shift* and *line feed*)	Tail marker (*line feed* and *letter shift*)
Reference code (access relay-set code and ticket-reperforator code)	Reference code (access relay-set code and ticket-reperforator code)
Time	Time
Calling subscriber's answer-back code	Duration (in self-checking code)
Selection signals	Duration (in digits of International Alphabet No. 2)
Route-indication character or transit-register code	

BASIC MICROWAVE COMMUNICATION C, 1976

Students were expected to answer any 6 questions

Q 1 (a) Explain why, at the surface of a perfect conductor, an electromagnetic wave can have no component of

(i) the E-field parallel to the surface, and
(ii) the H-field perpendicular to the surface.

(b) Fig. 1 represents a vertically-polarized wave propagated in the direction AB and meeting a horizontal perfectly-conducting surface at an angle θ.

(i) Explain, with the aid of sketches, any changes that occur in the E-field and the H-field on reflection.
(ii) Draw a diagram showing the direction of the E-field for the incident and reflected waves.

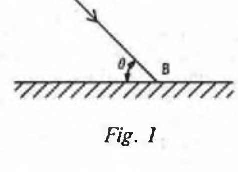


Fig. 1

A 1 See A1, Basic Microwave Communication C, 1975, *Supplement*, Vol. 69, p. 85, Jan. 1977.

Q 2 (a) Briefly describe a method of measuring the approximate values of the following parameters of a low-loss transmission line:

(i) the characteristic impedance, and
(ii) the propagation velocity.

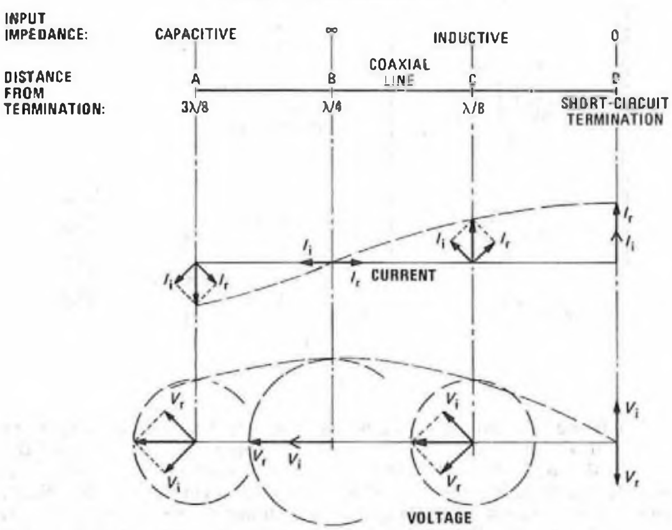
(b) A $3\lambda/8$ section of low-loss coaxial cable is terminated in a short-circuit. Explain, with the aid of phasor diagrams, why the input impedance of the section is a pure reactance.

A 2 (*a*) By considering the voltage and current at a distance l metres from a short-circuit on a low-loss transmission line, it can be shown that the input impedance of the line, Z_{SC}, is $jZ_0 \tan \beta l$ ohms, where Z_0 is the characteristic impedance, $\beta = 2\pi/\lambda$ and λ is the wavelength (metres). Similarly, the open-circuit impedance, Z_{OC}, is $-jZ_0 \cot \beta l$ ohms. The characteristic impedance can therefore be obtained by using a bridge to measure the input impedance of the line terminated by (*i*) a short-circuit and (*ii*) an open-circuit, both impedances being measured at the same frequency. From the 2 measurements,

$$Z_0 = \sqrt{(Z_{OC}Z_{SC})} \text{ ohms.}$$

(*ii*) The propagation velocity can be obtained by dividing Z_{SC} by Z_{OC}, giving $\tan \beta l = \pm j\sqrt{(Z_{SC}/Z_{OC})}$. At any point on the line, Z_{SC} and Z_{OC} are reactive and of opposite signs, so that $\tan \beta l$ always has a real value. As $\beta = 2\pi/\lambda$, the propagation velocity, v, is $2\pi f/\beta$ metres/second, where f is the frequency (hertz). It should be noted that β is multivalued; solutions occur in each quandrant of βl. The approximate value of β must therefore be known. Failing this, an approximate value can be readily obtained by measuring the input impedance of a line terminated in a short-circuit, and progressively increasing the frequency until the first high-impedance point is reached. Under this condition, the line is $\lambda/4$ metres long. If this length is called z, then $\lambda = 4z$. Hence,

$$v = 4zf \text{ metres/second.}$$



(*b*) Referring to the sketch, the reflected current, I_r, at the short-circuit termination (D) is in phase with the incident current, I_i, giving a current of twice the value which would be present in an infinite line of negligible loss. The reflected voltage, V_r, is in antiphase to the incident voltage, V_i, fulfilling the requirement that the voltage across the short-circuit must be zero. Moving back from the termination in the direction of the source to point C, the incident wave is advanced, but the reflected wave is retarded. The total current at point C is the vector sum of the 2 components, as shown. The incident and reflected voltages behave similarly, but the resultant is 90° in advance of the current, so that the impedance at point C is inductive. At point B, the impedance becomes infinite. At point A, the current leads the voltage by 90°. The input impedance at point A is therefore a pure capacitive reactance.

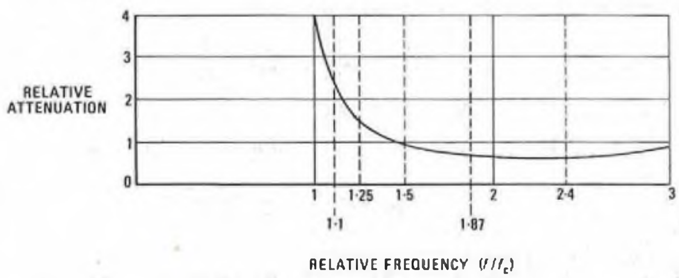
Q 3 (a) For dominant-mode propagation in a rectangular waveguide, it is recommended that frequencies of between 1·25 and 1·87 times the critical frequency should be used. What would be the disadvantage of energizing at a frequency of

(i) 1·1 times the critical frequency, and
(ii) 2·4 times the critical frequency?

(b) A rectangular waveguide has a cross-section of 4 × 2 cm. Calculate

(i) the lowest recommended frequency at which it should be energized, and
(ii) the guide wavelength if it is energized at a frequency of 5 GHz.

A 3 (*a*) (*i*) The relative-loss/relative-frequency graph for the dominant (TE_{10}) mode in a rectangular waveguide is shown in the sketch. It can be seen that the loss at a working frequency, f, of 1·1 times the critical (cut-off) frequency, f_c, is more than twice the loss in the middle of the useful frequency range at $1{\cdot}5f_c$. The group delay is also increased as the frequency approaches the cut-off point. The attenuation and differential delay for frequencies near f_c would therefore vary considerably over the range of frequencies transmitted along the waveguide.



(*ii*) At a frequency of $2{\cdot}4f_c$, the waveguide is more than 2 half-wavelengths wide, and can support both the TE_{10} and TE_{20} modes. This could cause dissipation of power in the unwanted mode, and could also cause problems of interference between the waves due to the 2 modes having different propagation velocities and attenuations per unit length.

(*b*) (*i*) The cut-off frequency corresponds to half a free-space wavelength (λ_c) at the cut-off frequency, equalling the wide dimension, a, of the waveguide; thus, $\lambda_c = 2a$. Since the frequency multiplied by the wavelength is equal to the velocity of light, c, then

$$f_c = \frac{c}{2a} = \frac{3 \times 10^{10}}{2 \times 4} = 3{\cdot}75 \text{ GHz.}$$

Therefore, the lowest recommended frequency

$$= 1{\cdot}25f_c = 1{\cdot}25 \times 3{\cdot}75 = 4{\cdot}688 \text{ GHz.}$$

(*ii*) Now,
$$\frac{1}{\lambda_g^2} = \frac{1}{\lambda_0^2} - \frac{1}{\lambda_c^2},$$

where λ_g is the guide wavelength, λ_0 is the free-space wavelength, and $\lambda_c = 2a$. At 5 GHz, $\lambda_0 = c/f = 3 \times 10^{10}/5 \times 10^9 = 6$ cm.

$$\therefore \frac{1}{\lambda_g^2} = \frac{1}{6^2} - \frac{1}{8^2}.$$

$$\therefore \lambda_g = 9{\cdot}07 \text{ cm.}$$

Q 4 Fig. 2 represents a cavity made by closing the ends of a 10 cm length of 6 × 3 cm waveguide. The cavity is to be excited at its lowest resonant frequency by a coaxial cable which may be connected at either X or Y.

(a) (i) For each of these positions, state with reasons whether a loop or a probe would be required.
(ii) For the coupling requiring a loop, explain with the aid of a well-drawn diagram how the loop should be positioned.

(b) Calculate the lowest frequency at which the cavity resonates.

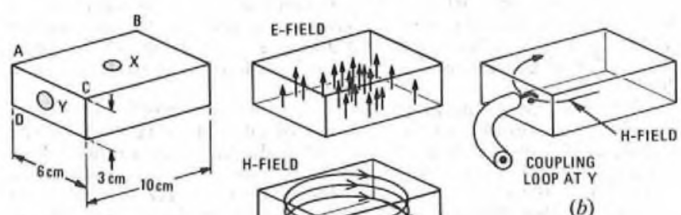


Fig. 2

(*a*)

A 4 (*a*) (*i*) The configuration of the electric (E) and magnetic (H) fields in the cavity at the lowest resonant frequency are as shown in sketch (*a*). At position X, the magnetic field is zero and the electric field is a maximum. A probe is necessary at this point to couple to the electric field. At position Y, the electric field is zero and the magnetic field is a maximum. A loop is necessary at this point to couple to the magnetic field.

(*ii*) The positioning of a loop at Y is shown in sketch (*b*). The loop is oriented so that the maximum magnetic field passes through it. The coupling can be reduced by rotating the loop. The orientation is adjusted to give a loop area sufficiently effective for adequate coupling, but not so great that the *Q*-factor of the cavity is unduly reduced by the shunting effect of the coupled load.

(*b*) The relationship between free-space wavelength (λ_0) and guide wavelength (λ_g), for propagation in a rectangular waveguide of broad dimension *a*, can be applied to the rectangular cavity, which has a length of $\lambda_g/2$ at its lowest resonant frequency.

$$\therefore \frac{1}{\lambda_0^2} = \frac{1}{\lambda_g^2} + \frac{1}{(2a)^2} = \frac{1}{20^2} + \frac{1}{12^2}.$$

$$\therefore \lambda_0 = 10 \cdot 29 \text{ cm.}$$

The frequency, *f*, is given by the speed of light divided by λ_0.

$$\therefore f = \frac{3 \times 10^{10}}{10 \cdot 29} \text{ Hz} = \underline{2 \cdot 915 \text{ GHz.}}$$

Q 5 (a) Fig. 3 shows the path of a wave from a source, T, located at the focus of a parabolic reflector. Using the dimensions shown, calculate the focal length of the reflector.

(b) Explain briefly why a parabolic reflector can convert a spherical wave into a plane wave.

(c) The transmitting aerial of a microwave link has a gain of 600 with respect to an isotropic source. Calculate the power that must be supplied to this aerial to ensure that a correctly aligned receiving aerial of effective aperture 4 m² will receive 1 μW at a range of 12 km.

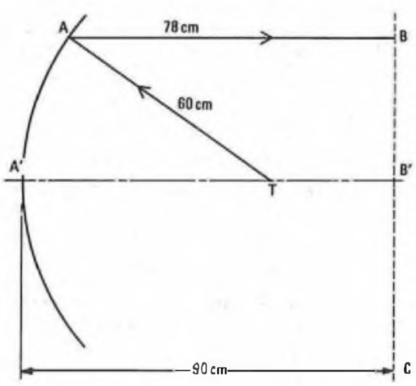


Fig. 3

A 5 (*a*) A property of a parabola is that the distance TAB is constant, regardless of the position of A on the parabola.

$$\therefore \text{TA'B'} = \text{TAB} = 60 + 78 = 138 \text{ cm,}$$

or $$2\text{A'T} + \text{TB'} = 138 \text{ cm.} \quad \ldots\ldots (1)$$

Also $$\text{A'T} + \text{TB'} = 90 \text{ cm.} \quad \ldots\ldots (2)$$

Subtracting equation (2) from equation (1) gives the focal length

$$= \text{A'T} = 138 - 90 = \underline{48 \text{ cm.}}$$

(*b*) The fact that distance TAB is constant for all positions of A means that a spherical wave radiated from the focal point and reflected at the parabolic surface arrives at all parts of plane BC in the same phase. The reflected wave-front is therefore plane and parallel to plane BC.

(*c*) In free space, the power radiated from an isotropic radiator is distributed uniformly over the surface of any sphere centred on the radiator. The surface area of a sphere is 4π times the radius. Thus, for a range of 12×10^3 m, the power is distributed over an area of $4\pi \times 12^2 \times 10^6$ m². The receiving aerial collects 1 μW from an area of 4 m², so that the transmitted power required from an isotropic radiator is

$$\frac{4\pi \times 12^2 \times 10^6 \times 1 \times 10^{-6}}{4} \text{ W.}$$

However, the transmitting aerial has a power gain of 600 (with respect to an isotropic radiator) in its correctly aligned position, so that the power required from the transmitter

$$= \frac{4\pi \times 12^2}{4 \times 600} \text{ W} = \underline{754 \text{ mW.}}$$

Q 6 (a) A modulated wave comprises a 10 V 1500 kHz carrier and 2 side-frequency components: namely, 4 V 1505 kHz and 4 V 1495 kHz. Explain, with the aid of phasor diagrams, how these components can be assembled to produce

(i) pure amplitude modulation, and

(ii) phase modulation with some amplitude modulation.

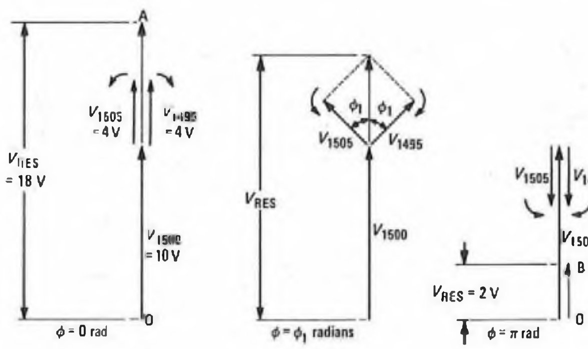
(b) From the diagram for part (a) (ii), calculate

(i) the maximum deviation in radians, and

(ii) the depth of the amplitude modulation.

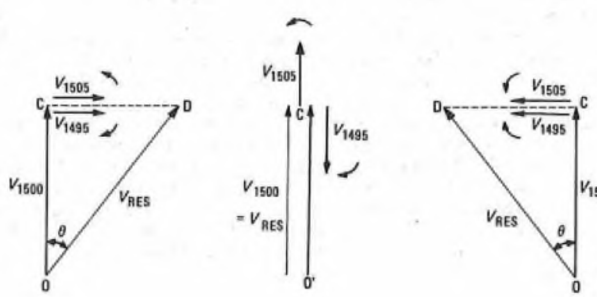
(c) Why, in practice, is frequency modulation of a carrier followed by several stages of frequency multiplication to obtain the required deviation?

A 6 (*a*) (*i*) Pure amplitude modulation can be produced by adding the carrier (V_{1500}) and side-frequency (V_{1505} and V_{1495}) components together, arranging that all three are in phase at some point $\phi = 0$ rad (see sketch (*a*)), where $\pm\phi$ is the phase angle between the side frequencies and the carrier. After a time such that phasor V_{1505} has advanced on phasor V_{1500} by ϕ_1 radians, phasor V_{1495} will lag V_{1500} by ϕ_1 radians. The sum of the 3 phasors, V_{RES}, is still in phase with the carrier, but its amplitude is reduced. This process continues, with the amplitude of V_{RES} varying between AO and BO at a frequency of 5 kHz. The modulation factor is half the excursion of V_{RES} divided by the carrier amplitude; that is, $(18 - 2)/(2 \times 10) = 80\%$.



(*a*)

(*ii*) If the 3 components are assembled with the side frequencies in phase at a point 90° to the carrier (see sketch (*b*)), the phase of the resultant varies through 2θ radians at a frequency of 5 kHz. The resultant is therefore phase-modulated, with a deviation of $\pm\theta$ radians, but it also contains some amplitude modulation because V_{RES} varies in length between CO and DO. This variation occurs twice for each cycle of the phase modulation, and therefore has a frequency of 10 kHz.



(*b*)

(*b*) (*i*) From sketch (*b*), the maximum deviation

$$= \theta = \tan^{-1}\frac{8}{10} = \underline{0 \cdot 675 \text{ rad.}}$$

(*ii*) The modulation factor is half the excursion of V_{RES} divided by the mean length of V_{RES}. Thus, the modulation factor

$$= \frac{0\cdot5\{\sqrt{(10^2+8^2)}-10\}}{0\cdot5\{\sqrt{(10^2+8^2)}+10\}},$$

$$= \underline{12\cdot3\%.}$$

(*c*) The process of frequency multiplication multiplies both the carrier frequency and the deviation, and gives the required frequency deviation at the transmitted frequency for a much smaller deviation at the exciter stage. For example, a typical ultra-high-frequency transmission can be obtained from a 4 MHz exciter by multiplying (in stages) by 108 to give 432 MHz. If a maximum deviation of $\pm$15 kHz is required, it is obtained by a 15/108 kHz = 139 Hz deviation at the exciter. This can be achieved by simple phase modulation of a crystal-controlled 4 MHz source, simplifying frequency-stability problems. A reduced deviation at the exciter also reduces non-linear distortion and unwanted amplitude modulation, as is apparent from sketch (*b*) if θ is considered to be small.

Q 7 (a) Explain briefly what is meant by

(i) white noise,
(ii) thermal-agitation noise,
(iii) shot noise,
(iv) noise temperature, and
(v) impulsive noise.

(b) The input stage of a receiver has a gain-bandwidth product of 50 MHz and a noise factor of 4 dB. Taking kT as 4 × 10⁻²¹ J, calculate the noise power generated in the stage.

A 7 (*a*) (*i*) Noise is said to be white when the noise power per unit bandwidth is distributed uniformly over the frequency range concerned.

(*ii*) Thermal-agitation noise is white noise caused by the thermal agitation of electrons in a conductor. The noise power available is kTB watts, where k is Boltzmann's constant ($1\cdot38 \times 10^{-23}$ J/K), T is the absolute temperature (kelvins), and B is the bandwidth (hertz).

(*iii*) Shot noise is caused by the random emission of individual electrons from, for example, the cathode of an electronic tube, or by the random movements of individual charge carriers in a semiconductor device.

(*iv*) Noise temperature is the equivalent temperature of a resistor producing the same noise power as is present in a given circuit. If N is the noise power in the circuit (watts), the temperature of an equivalent noise source is $T = N/kB$ kelvins, from the formula given in part (*ii*).

(*v*) Impulsive noise is caused by interference from such sources as power switching and telephone dial pulses, and can be defined as random transient disturbances.

(*b*) Noise factor, F, which is the signal-to-noise ratio at the input to a stage divided by that at the output, can also be defined as

$$F = \frac{GN_i + N_r}{GN_i} = 1 + \frac{N_r}{GN_i},$$

where G is the gain of the stage, N_i is the input noise power (watts), N_r is the noise power generated in the stage (watts), and F and G are expressed as linear numerical ratios.

Since $N_i = kTB$ watts,

$$F = 1 + \frac{N_r}{kTBG},$$

or $$N_r = kTBG(F-1) \text{ watts.}$$

Now, $kT = 4 \times 10^{-21}$ J, $BG = 50$ MHz, and $F = 4$ dB, which is a linear ratio of $10^{0\cdot4} = 2\cdot512$.

$$\therefore N_r = 4 \times 10^{-21} \times 50 \times 10^6 \times (2\cdot512 - 1) \text{ W},$$

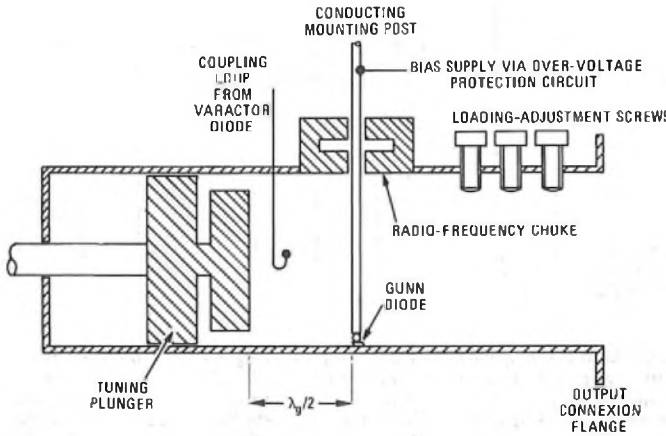
$$\simeq \underline{0\cdot302 \text{ pW}.}$$

Q 8 (a) Briefly describe, with the aid of a sketch, the construction and action of a Gunn-diode oscillator.

(b) Under the following headings, compare the Gunn-diode oscillator with the reflex klystron for use as a local oscillator in a receiver:

(i) size (including power supplies),
(ii) controllable tuning range,
(iii) power output,
(iv) power efficiency, and
(v) noise.

A 8 (*a*) The construction of a typical Gunn-diode oscillator is shown in the sketch. The Gunn diode is made by growing an epitaxial layer, about 10 μm thick, of crystalline n-type gallium arsenide on a substrate of more heavily doped n-type material. An alloyed contact is made to the top surface of the active expitaxial layer. When an electric field exceeding about 350 V/mm is applied, a high-electric-field domain developes at the cathode. The domain drifts across to the anode at a speed of about 10^8 mm/s; another domain develops at the cathode, and the process is repeated. The current thus contains a component whose frequency is related to the transit time; for the velocity and distance figures given above, the corresponding frequency is 10 GHz. The current and voltage fluctuations are coupled via the mounting post to the waveguide cavity, which resonates at half a guide wavelength, $\lambda_g/2$. The oscillations are coupled to the output via a matching-screw section.



The cavity can be mechanically tuned by adjusting the short-circuit plunger. Electrical tuning can be effected by altering the bias on a varactor diode, which acts as a variable capacitance and is loop-coupled into the cavity. A radio-frequency choke is provided in the DC supply to the diode at the point where the mounting post enters the cavity. This restricts the oscillations to the cavity, and inhibits any spurious oscillations in the bias circuit; such oscillations can be excited by the negative resistance presented by the Gunn diode to the biasing circuit under working conditions. Protection circuits are usually incorporated in the bias lead to avoid damage by voltage spikes or misuse.

(*b*) (*i*) The Gunn-diode oscillator is very compact, and can be made smaller than the equivalent klystron. It requires only about 1 W of power at a few volts, and its power unit is therefore much smaller than that of a klystron, which must deliver about 10 W at 300 V with, in addition, a low-current supply adjustable between −350 V and −650 V.

(*ii*) A Gunn oscillator can be mechanically tuned by up to about 50% whereas, in a klystron, mechanical tuning is limited to about 10%. A Gunn oscillator can be electronically tuned by a varactor diode by up to 20%, whereas a klystron can be electrically tuned via the reflector voltage by only about 0·5% between half-power points.

(*iii*) Modern medium-power Gunn oscillators, and small local-oscillator klystrons, both produce an output of about 50 mW (at 10 GHz), which is more than adequate for local-oscillator requirements.

(*iv*) The klystron requires about 10 W of power, and the Gunn oscillator about 1 W to produce the same output power, so that the modern Gunn oscillator is correspondingly more efficient.

(*v*) The frequency-modulation noise generated in a Gunn oscillator is less than in a typical klystron.

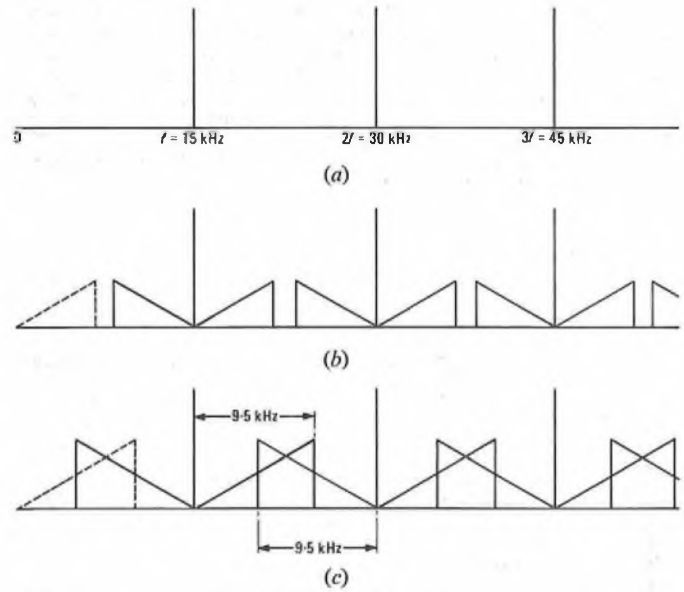
The amplitude-modulation noise output of either source in a local oscillator is less important, especially if a balanced mixer is used. However, the low-noise performance of a Gunn oscillator is superior to that of a klystron.

Q 9 (a) Band limitation of the modulating signal is required before pulse modulation takes place. Explain this statement, using as an example a baseband from 100 Hz to 9·5 kHz and a sampling rate of 15 kHz.

(b) A 6 GHz microwave link carries a train of equal mark and space pulses of duration 4 μs. State, with reasons, the following requirements of the receiver:

(i) the minimum value of intermediate-frequency (IF) bandwidth,
(ii) the order of the IF, and
(iii) the upper limit of the frequency response of the video amplifier.

A 9 (*a*) Distortion of the received signal results if the highest modulating frequency exceeds half the pulse-repetition frequency, f, of the sampling pulse train. This can be avoided by limiting the bandwidth of the modulating frequency so that the maximum modulating frequency is less than $f/2$.



The frequency spectrum of the sampling frequency, assuming a very short pulse duration, is as shown in sketch (*a*). Modulation by a complex signal results in each line carrying a pair of sidebands as shown in sketch (*b*). When the highest modulating frequency (9·5 kHz in the example) exceeds half the sampling frequency, the sidebands overlap, as shown in sketch (*c*), and the original modulating signal cannot be extracted without distortion.

(*b*) (*i*) The pulse-repetition frequency is the reciprocal of the duration of one cycle. Thus, $f = 1/(2 \times 4 \times 10^{-6})$ Hz = 125 kHz. The square-wave modulation contains components of f, $3f$, $5f$ etc. The minimum requirement for the IF bandwidth is that it should pass the first pair of sidebands. The minimum bandwidth is therefore $2f = 250$ kHz. This figure makes no allowance for instability of the carrier and local-oscillator frequencies, which would offset the signal from the centre of the IF band.

(*ii*) An IF of, say, 70 MHz would meet the design requirements, which are that:

(*1*) the IF must be large compared with the 250 kHz bandwidth to simplify the design of the IF stages,

(*2*) the IF must be a sufficiently large percentage of the carrier frequency of 6 GHz to simplify the design of the input filter, which must reject the image frequency, this being separated from the wanted frequency by twice the IF, and

(*3*) the IF must be sufficiently large to avoid the need for severe restrictions on the frequency stability of the carrier or the local oscillator.

(*iii*) The frequency of the demodulated signal is 125 kHz; that is, the difference between the carrier and the extremity of one of the sidebands. The upper limit of frequency response of the video amplifier must therefore also be 125 kHz.

Q 10 *The circuit shown in Fig. 4 is triggered by positive-going pulses.*

(a) With the aid of time-related waveform diagrams showing emitter and output-voltage variations, explain the action of the circuit.

(b) State, with reasons, the effect on performance of

(i) making time-constant C_1R_1 *three times the interval between trigger pulses, and*

(ii) reducing the ratio R_2/R_1.

(c) Explain briefly how distortion of the output waveform could be reduced.

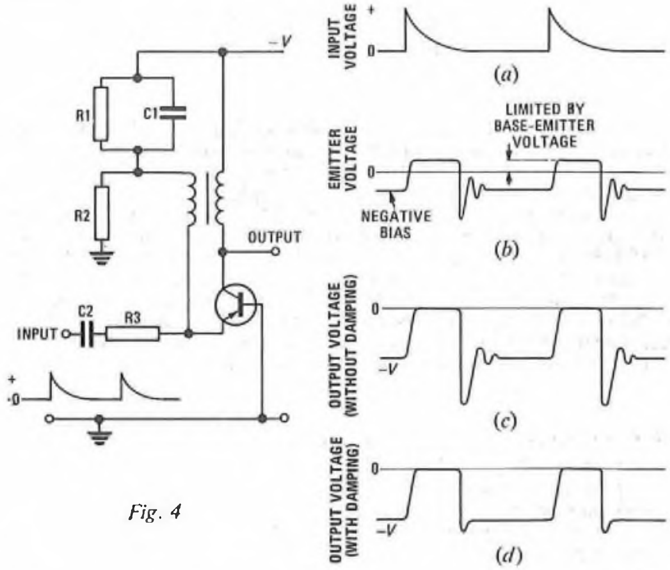


Fig. 4

A 10 (*a*) In the idle state, a small negative bias (sketch (*b*)) is applied to the transistor's emitter by the potential divider formed by resistors R2 and R1. Capacitor C1 is charged according to the ratio R_2/R_1. The collector current is zero, and the voltage at the output is $-V$ volts (the negative supply voltage), as shown in sketch (*c*). The positive input pulse (sketch (*a*)) is passed to the emitter through coupling capacitor C2 and emitter-source-impedance limiting resistor R3. The transistor conducts, the output voltage rises to 0 V, and the collector current rises at a constant rate via the inductance of the transformer's primary winding. The voltage across the primary winding induces a voltage in the secondary winding that assists, and then replaces, the positive voltage applied to the emitter by the incoming pulse (sketch (*b*)). The transistor rapidly attains the bottomed condition, and capacitor C1 loses its charge via the emitter current. Discharge of capacitor C1 allows the junction of resistors R1 and R2 to go negative for as long as the opposing voltage from the transformer is sufficient to keep the transistor conducting. Alternative factors determining the pulse length are saturation of the transformer core, or limitation of the collector current.

The emitter bias becomes negative and, through regenerative action, the transistor rapidly switches off. The sudden decrease of current through the inductive load generates a negative voltage spike at the collector, followed by damped oscillations as the output voltage settles back to $-V$ volts (sketch (*c*)). In the meantime, capacitor C1 recharges through resistor R2, and the circuit is restored for the next input pulse. (If the negative bias is small, $R_1 \gg R_2$ and the relevant time constant for recharging is C_1R_2.)

(*b*) (*i*) If time constant C_1R_2 is 3 times the interval between the trigger pulses, the first pulse will trigger the circuit as described above but, when the second pulse arrives, capacitor C1 will be only partially recharged, and the emitter will have a heavy negative bias. The input-pulse voltage will be insufficient to overcome the bias. By the time the fourth pulse arrives, capacitor C1 will be charged to 63% of its final value and, provided that the input pulse is large enough, the circuit will be triggered once again. Thus, the circuit will behave as a frequency divider.

(*ii*) The ratio R_2/R_1 determines the static emitter bias as a fraction of $-V$ volts. A reduction of this ratio will make the circuit easier to trigger, but will also make it more susceptible to false operation to variations in the amplitude of the triggering pulse.

(*c*) The large negative voltage spike on the output waveform can be reduced by adding a diode in series with a damping resistor across the primary winding. This dissipates the energy from the transformer, and allows critical damping to be achieved, as shown in sketch (*d*).

LINE PLANT PRACTICE C, 1976

Students were expected to answer any 6 questions

Q 1 (a) Explain how a radio transmitter can cause audible interference with a telephone circuit.

(b) Describe the measures which can be taken to overcome this type of interference.

A 1 See A1, Line Plant Practice C, 1971, *Supplement*, Vol. 65, p. 84, Jan. 1973.

Q 2 A beam of cross-section as shown in Fig. 1 is made of steel. The safe working stress for steel is 140 MPa and the density of the steel is 7250 kg/m³.

(a) If the beam is simply supported over a span of 6 m, calculate the safe working load which could be uniformly distributed over the span.

(b) Draw a bending-moment diagram for the loaded beam.

(c) What is the maximum bending moment of the beam?

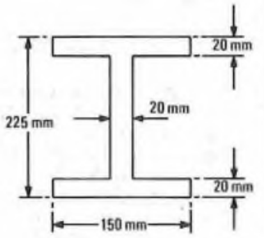
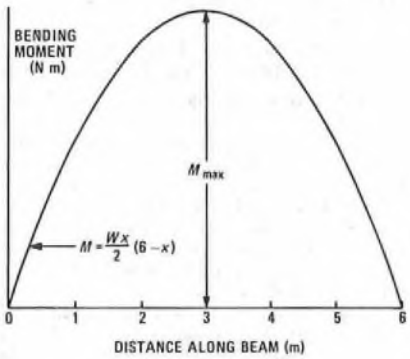


Fig. 1



(a)

A 2 (*a*) The cross-sectional area of the beam

$$= 2 \times (150 \times 20) + (185 \times 20) = 9700 \text{ mm}^2.$$

Thus, the volume of the beam

$$= 9700 \times 6 \times 10^{-6} = 0{\cdot}0582 \text{ m}^3.$$

Thus, the mass of the beam

$$= 0{\cdot}0582 \times 7250 = 422 \text{ kg}.$$

Now, for stresses in a loaded beam,

$$\frac{M}{I} = \frac{f}{y} \text{ newtons/metre}^3,$$

where M is the bending moment (newton metres), I is the moment of inertia (metres⁴), f is the stress in the beam (pascals), and y is the depth of the neutral axis (metres).

For a beam, $I = \frac{bd^3}{12}$ metres⁴,

where b is the breadth of the beam (metres), and d is the depth (metres).

$$\text{Thus,} \quad I = \frac{0{\cdot}150 \times 0{\cdot}225^3}{12} - \frac{0{\cdot}130 \times 0{\cdot}185^3}{12} \text{ m}^4,$$

$$= 0{\cdot}0000738 \text{ m}^4.$$

$$\therefore M_{max} = \frac{f_{max} I}{y} \text{ newton metres},$$

$$= \frac{140 \times 10^6 \times 0{\cdot}0000738}{112{\cdot}5 \times 10^{-3}} \text{ N m},$$

$$= 91{\cdot}84 \text{ kN m}.$$

But, for a uniformly distributed load,

$$M_{max} = \frac{WL^2}{8} \text{ newton metres},$$

where W is the total weight per unit length (newtons/metre) and L is the length of the beam (metres).

$$\therefore W = \frac{8M_{max}}{L^2} = \frac{8 \times 91{\cdot}84}{36} = 20{\cdot}41 \text{ kN/m}.$$

Now, the weight per unit length of the unloaded beam

$$= \frac{422 \times 9{\cdot}81}{6} = 690 \text{ N/m}.$$

Thus, the safe working load $= 20{\cdot}41 - 0{\cdot}69$ kN/m,

$$= \underline{19{\cdot}72 \text{ kN/m}}.$$

(*b*) The bending-moment diagram is shown in sketch (*a*), where x is the distance along the beam (metres).

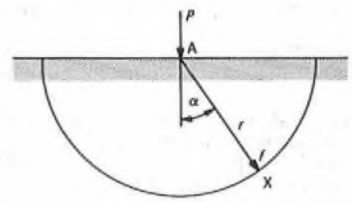
(*c*) The maximum bending moment is $\underline{91{\cdot}84 \text{ kN m}}$.

Q 3 (a) Describe the distribution of pressure within the ground caused by a traffic load. Explain how this is used to calculate the forces on a manhole.

(b) Calculate the horizontal load on a manhole wall at a point 0·83 m below the ground due to a static load of 90 kN applied at the surface at a horizontal distance of 1·18 m from the wall.

A 3 (*a*) The distribution of pressure resulting from an external load applied at the surface of a carriageway can be estimated using Boussinesq's theory. The sketch shows a load, P newtons, applied to a surface at point A. The stress produced at X is f pascals; r is the radial distance between A and X (metres), and α is the angle between AX and the vertical through A. Boussinesq's theory gives the relationship

$$f = \frac{3P}{2\pi r^2} \cos^2 \alpha \text{ pascals}.$$



Horizontal and vertical earth pressures can be obtained by resolving the equation for f into its vertical and horizontal components, f_v and f_h, giving

$$f_v = \frac{3P}{2\pi r^2} \cos^3 \alpha \text{ pascals},$$

$$\text{and} \quad f_h = \frac{3P}{2\pi r^2} \cos^2 \alpha \sin \alpha \text{ pascals}.$$

Thus, the horizontal and vertical components of the force can be used to estimate the load on a manhole roof or walls.

Simpler methods can be used to evaluate the stresses due to traffic loads in manhole design. These loads are assessed by assuming the traffic load to be equivalent to an extra depth of soil. This surcharge of soil is then introduced into the Rankine formula.

(*b*) The information given implies that the Boussinesq formula should be used to obtain the load.

$$\text{Now,} \quad r = \sqrt{(1{\cdot}18^2 + 0{\cdot}83^2)} = 1{\cdot}44 \text{ m}.$$

$$\text{Also,} \quad \cos \alpha = \frac{0{\cdot}83}{1{\cdot}44}, \text{ and } \sin \alpha = \frac{1{\cdot}18}{1{\cdot}44}.$$

$$\text{Thus,} \quad f_h = \frac{3 \times 90 \times 10^3}{2\pi \times 1{\cdot}44^2} \times \frac{0{\cdot}83^2}{1{\cdot}44^2} \times \frac{1{\cdot}18}{1{\cdot}44} \text{ Pa},$$

$$= \underline{5{\cdot}642 \text{ kPa}}.$$

Q 4 For a pressurized cable system,

(a) give details of 8 advantages of using such a system,

(b) give details of 3 disadvantages of using such a system, and

(c) briefly outline the effect that different operating pressures have on stresses in a cable-joint sleeve.

A 4 (*a*) The advantages to be obtained from a pressurized cable system are given below.

(*i*) Cable pressurization provides a higher standard of electrical insulation. This is due to the reduction in cable sheath defects and the desiccating properties of the gas.

(*ii*) Cable sheath faults become apparent immediately the sheath is damaged. Thus, sheath repairs can be carried out before moisture enters the cable to cause circuit faults.

(*iii*) The internal gas pressure provides protection for the cable core against water or damp air under fault conditions. Protection of this type results in considerable saving in the number of length renewals required, and eliminates "piecing out" sections of damaged cable.

(*iv*) Cable breakdowns, which are costly in terms of time, money and service disruption, are reduced.

(*v*) The responsibility for damage to telecommunications cables by other authorities is more readily determined. The audible hiss of escaping gas from a damaged pressurized cable provides an indication to the authorities concerned that a cable has been pierced. Also, the consequent drop in pressure is noticed by the local maintenance staff, enabling the damage to be localized and the damaging party notified immediately.

(*vi*) Test equipment for localizing sheath faults is robust and is able to withstand adverse weather conditions.

(*vii*) Inaccuracies in fault location measurements, due to difficulty in obtaining a satisfactory insulation-resistance ratio between a good and faulty wire, are eliminated.

(*viii*) The reduction in cable breakdowns and insulation faults provides a considerable improvement in the service provided by the telecommunications network.

(*b*) The disadvantages of cable pressurization are as follows.

(*i*) Pressurization requires more records to be kept; the positions of monitoring equipment and testing points on cables must be recorded. The accuracy of such records must be maintained to assist in the rapid and successful location of cable faults.

(*ii*) More work is involved when diversions and rearrangements have to be carried out on pressurized schemes. Maintenance staff controlling test panels on the cables involved must be forewarned of the proposed work and, on completion, the cable has to be repressurized.

(*iii*) While electrical tests can be carried out from a maintenance test room, testing on pressurized cables normally requires access to jointing chambers.

(*c*) The maximum safe pressure which can be resisted by a cable-joint sleeve can be calculated by likening the cable sheath to a thin cylindrical shell which is subject to longitudinal and hoop stresses.

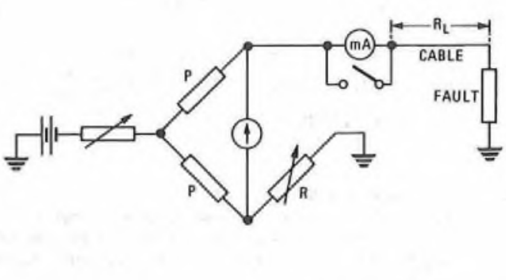
The longitudinal stress $= Pd/4t$, and the hoop stress $= Pd/2t$, where P is the operating pressure, d is the sleeve diameter and t is its thickness.

The operating pressure is chosen so that the sleeve design meets the hoop stress, which is twice the longitudinal stress.

Q 5 (*a*) *Describe in detail, with the aid of a diagram, a 2-current test method of locating a break in a submarine-cable sheath.*

(*b*) *Explain why special methods of DC testing have to be used when locating a submarine cable break with conductors exposed to the sea.*

A 5 (*a*) Kennelly's 2-current test is made using a Wheatstone bridge with a milliammeter in series with the line as shown in the sketch. When the bridge is balanced, a "false zero" is used on the galvanometer scale to eliminate the effects of any spurious current. This zero is determined by removing the test battery, which allows current to flow through the galvanometer due to cable capacitance discharge and also due to the polarization EMF at the fault. The effect on the galvanometer is a "flick", changing to a steadily-decreasing deflexion. The point at which the change takes place is regarded as the false zero to which bridge-resistance adjustments are made.



A balance, R_1, is obtained with a current to line of I_1, which should not exceed 25 mA. The current is adjusted by means of the variable resistor in series with the battery, and a new balance, R_2, is found using a second, lower, value of current, I_2. By Kennelly's law,

$$\frac{R_{F1}}{R_{F2}} = \frac{\sqrt{I_2}}{\sqrt{I_1}},$$

where R_{F1} and R_{F2} are the apparent fault resistances introduced at the fault point by line currents I_1 and I_2 respectively.

Also, $R_1 = R_L + R_{F1}$ and $R_2 = R_L + R_{F2}$,

where R_L is the resistance of the conductor to the fault.

$$\therefore \frac{R_{F1}}{R_{F2}} = \frac{\sqrt{I_2}}{\sqrt{I_1}} = \frac{R_1 - R_L}{R_2 - R_L}.$$

$$\therefore R_1\sqrt{I_1} - R_L\sqrt{I_1} = R_2\sqrt{I_2} - R_L\sqrt{I_2}.$$

$$\therefore R_L = \frac{R_1\sqrt{I_1} - R_2\sqrt{I_2}}{\sqrt{I_1} - \sqrt{I_2}}.$$

If the current ratio is 4 : 1, for example, then $R_L = 2R_1 - R_2$. To determine the actual fault position in nautical miles, it is first necessary to make an allowance for

(*i*) any known fixed resistances, such as repeaters, and

(*ii*) the estimated "fixed" fault resistance,

and then to divide the residual by the ohm/nautical-mile constant for the cable.

(*b*) When a break occurs in a submarine cable and the copper conductor is exposed to sea water, a simple cell is formed between the copper (positive) and the iron armour wires (negative), with the sea water acting as the electrolyte. This cell generates a current which flows from the conductor to the sea and makes DC fault localization extremely difficult. If a DC potential is applied to the conductor during tests, electrolytic action takes place at the break, provided that the potential difference at the fault due to the applied potential is greater than 1·4 V. The conductor is coated with cupric chloride when a potential positive with respect to earth is applied, and with hydrogen bubbles when negative potential is applied. This results in a variable resistance between the conductor and the sea water depending upon the polarity and magnitude of the applied current. This, and the effects of spurious line currents, are not taken into account in the straightforward measurement of resistance and could, therefore, give rise to errors in the localization of the fault. Most DC methods of locating faults in submarine cables are, therefore, designed to reduce errors due to the polarization EMF at the fault, varying earth potential and spurious currents in the line.

Q 6 *An 80 m span of dropwire from a pole to a house is tensioned at 36 N. As a result of ice forming on the wire, the tension increases to 181 N. The weight of the wire without ice is 0·15 N/m, the total conductor cross-section is 0·8 mm² and Young's modulus (E) is 205 GPa. Assume that the ice coating is uniform over the length of the dropwire and neglect the strength of the ice.*

Calculate the mass of ice coating per metre of dropwire.

A 6 The original length of the wire, l_1, is given by $l_1 = L + \dfrac{8d^2}{3L}$ metres, where L is the span length (metres), and d is the dip at the centre of the span (metres).

But $\quad d = \dfrac{WL^2}{8T_1}$ metres,

where W is weight of wire per unit length (newtons/metre), and T_1 is the original tension in the wire (newtons).

Thus, $\quad l_1 = 80 + \dfrac{0{\cdot}15^2 \times 80^3}{24 \times 36^2} = 80{\cdot}370$ m.

If l_2 is the length of the wire when carrying ice, and T_2 is the new tension,

$$\frac{l_2 - l_1}{l_1} = \frac{T_2 - T_1}{A} \times \frac{1}{E},$$

where A is the cross-sectional area of the wire (metres²), and E is Young's modulus (pascals).

$$\therefore l_2 = \frac{l_1(T_2 - T_1)}{AE} + l_1 \text{ metres,}$$

$$= \frac{80{\cdot}370 \times (181 - 36)}{0{\cdot}8 \times 10^{-6} \times 205 \times 10^9} + 80{\cdot}370 \text{ m,}$$

$$= 80{\cdot}441 \text{ m.}$$

If W_i is the weight of the wire with the ice coating, then

$$l_2 = L + \frac{W_i^2 L^3}{24T_2^2} \text{ metres.}$$

$$\therefore\ W_i = \sqrt{\frac{24T_2^2(l_2 - L)}{L^3}} \text{ newtons/metre,}$$

$$= \sqrt{\frac{24 \times 181^2 \times (80{\cdot}441 - 80)}{80^3}} = 0{\cdot}823 \text{ N/m.}$$

Thus, the ice coating weighs

$$0{\cdot}823 - 0{\cdot}15 = 0{\cdot}673 \text{ N/m.}$$

Hence, the mass of ice per unit length

$$= \frac{0{\cdot}673}{9{\cdot}81} = \underline{0{\cdot}07 \text{ kg/m.}}$$

Q 7 (a) Explain the term bond strength as applied to reinforced concrete design.

(b) Quote a typical value for a bond strength which could be used in reinforced concrete design calculations.

(c) Using the value quoted for (b), calculate the minimum length of embedment of a 12 mm diameter reinforcing rod needed to achieve a bond with the concrete. (The maximum tensile stress in steel is 140 MPa.)

(d) Describe briefly 2 methods of improving the bond between reinforcing bars in concrete.

A 7 (*a*) Bond strength in reinforced concrete refers to the magnitude of the adhesion between the concrete and the steel reinforcing bars, and is the bonding force per unit surface area of the reinforcement.

Concrete shrinks when setting and hardening, thereby creating a firm bond between concrete and steel. It is necessary to embed the steel bars in the concrete for a length which provides sufficient adhesion between concrete and steel; otherwise, the steel bars may slip and fail to take up their proportion of the load.

Reinforcement of concrete must be such that the overall bonding force must be at least equal to the permissible tension in the reinforcing bars. The bonding force is the bond strength multiplied by the surface area of the reinforcing bars in contact with the concrete.

(*b*) A typical value for bond strength is 1 MPa.

(*c*) If L is the length of the steel bar (metres), and d is the diameter (metres),

the bonding force $= \pi d L \times$ (bond strength) newtons.

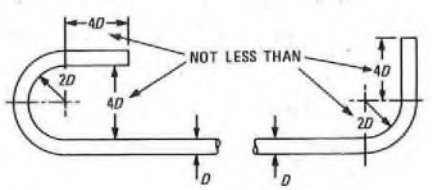
Also, the tension in the bar at maximum stress

$$= 140 \times 10^6 \times \frac{\pi d^2}{4} \text{ newtons.}$$

Equating the bonding force and the limiting tension in the bar gives

$$12\pi \times 10^{-3} \times L \times 1 \times 10^6 = \frac{140 \times 12^2 \pi}{4} \text{ newtons.}$$

$$\therefore\ L = \frac{140 \times 12}{4 \times 10^3} \text{ m} = \underline{420 \text{ mm.}}$$

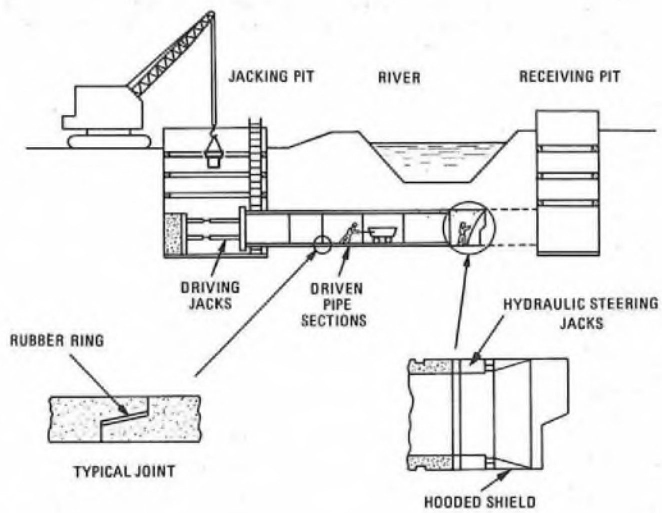


(*d*) Adhesion between the steel and concrete can be increased by providing hooks on the steel rods, as shown in the sketch. The U-hook gives an anchorage equivalent to that provided by a length of straight rod equal to 16 times the diameter of the bar. Another method of improving the bond is the use of deformed steel bars, which give a greater adhesion than can be obtained using straight rods.

Q 8 (a) Describe, with the aid of sketches, the method that would be used to pipe-jack a 1·2 m concrete pipe under a river 20 m wide.

(b) Describe a method of installing 90 mm diameter PVC pipes in the concrete pipe after it has been laid and the manholes built at each end.

A 8 (*a*) It is first necessary to determine the depth of the river bed and to make a ground survey in the areas where the drive and receive pits will be dug. The ground survey includes trial excavations and the sinking of boreholes. If the ground is found to be unstable due to the presence of silt, sands or ballasts, it may be necessary to carry out stabilization by one of the accepted methods (for example, chemical consolidation or freezing). The site for the driving (jacking) and receiving pits having been agreed, the excavations are commenced and the pits are dug to a depth which allows for a clearance of at least 3 m between the top of the pipe and the river bed. The size of the drive pit for the jacking of a 1·2 m pipe is of the order of 7 m long, 4 m wide, and deep enough to allow for adequate clearance of the river bed. A thrust wall of reinforced concrete is built on the pit wall facing the direction of thrust. The thrusting jacks (at least 2, depending on the soil) are then set up and aligned to give the correct direction of thrust to ensure that the pipes emerge in the correct position at the other side of the river.



The sketch shows the site layout for the operations. It is norma practice to fit the leading concrete pipe with a hooded shield. The shield provides a suitable cutting edge and includes jacks to allow control of the line and level of the drive as it proceeds. The pipes are jacked into the ground one at a time while the miner at the cutting face within the shield excavates by hand and removes spoil by means of a small trolley system. When the end of the thrust has been reached, the cutting shield is removed. In difficult ground, it may be necessary to lubricate the pipe during jacking by injecting bentonite, or similar materials, through the grout holes. The pits may be used for the construction of manholes if required, or the duct route may be extended to suitable manhole positions. If manholes are not constructed, reinforced concrete collars must be constructed at the ends of the pipe to minimize differential movement. Depending on ground conditions, the concrete pipe may be stabilized in the ground by pumping cement grout through prepared holes in the pipes to fill the overbreak and eliminate the possibility of subsidence at a later date.

(*b*) A 1·2 m pipe accommodates about eighty 90 mm PVC ducts, and these can be laid by hand in 1·5 m lengths within the concrete pipe or drawn-in, either singly or in groups of up to 10 pipes at a time. In the case of 6 m ducts, joints are made as the ducts enter the pipe or, if sufficient clear space is available, the ducts can be laid out and jointed on the ground before pulling in. To ensure reasonable rigidity of a large number of pipes during subsequent cabling operations, concrete or cement grout is placed between the interstices of the duct, and between the ducts and the inside of the concrete pipe. Where duct is laid in short lengths by hand, concrete of quality F (1 : 2 : 3 mix using 9·5 mm aggregate) is hand packed approximately every 1·5 m. In the case of the 6 m lengths of duct drawn into the concrete pipe, a sand/cement or pulverised-fly-ash/cement grout is gravity fed into the pipe to fill most voids.

Q 9 (a) Describe, with the aid of sketches, 4 different types of foundation that could be used in a steel self-supporting aerial tower.

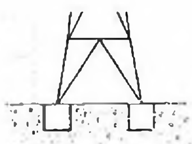
(b) A square-sided concrete block is used for one leg of a tower foundation. The depth of the block must not exceed 4·75 m and the density of concrete is 2200 kg/m³. Calculate the size of a concrete block which would be required to withstand an uplift of 1175 kN with a factor of safety of 1·5.

A 9 (*a*) The simplest of foundations for a self-supporting tower is an individual reinforced concrete block for each leg of the tower (see sketch (*a*)). The weight of each block, together with any support that the surrounding soil may be considered to provide, must be adequate to resist the maximum uplift that can be experienced in that leg, and provide a factor of safety. The size of the base of the block must be sufficient to limit the pressure on the soil due to downthrust to within the allowable value for the soil.

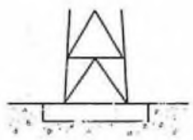
An alternative is the slab foundation, in which all the legs of a tower are founded in a single slab of reinforced concrete (see sketch (*b*)). The overturning moment is resisted by the couple generated between the total structure and foundation weight acting through the centroid of the structure, and the resultant upthrust of the soil pressure distribution beneath the slab acting through the centre of pressure.

In some soils, it is necessary to resort to piled foundations, or a combination of block or slab and piles (see sketch (*c*)). In this case, the pile is driven until it meets sound bearing soil or until adequate friction is developed along the surface of the pile to resist the downthrust from the tower.

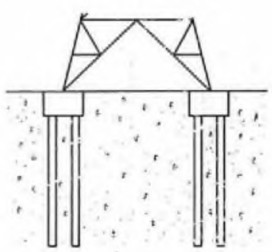
Where the tower is to be founded on sound rock, good use can be made of the rock by using rock anchors to resist the uplift force in a leg of the tower (see sketch (*d*)).



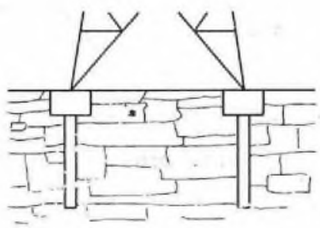
Individual Blocks
(*a*)



Slab
(*b*)



Piled
(*c*)



Rock Anchored
(*d*)

(*b*) The volume of the block is $4 \cdot 75S^2$ metres^3, where S is the side length of the block (metres). The weight of the block is $4 \cdot 75S^2 \times 2200 \times 9 \cdot 81$ newtons.

For equilibrium, the weight of the block must equal the uplift multiplied by the safety factor.

$$\therefore\ 4 \cdot 75S^2 \times 2200 \times 9 \cdot 81 = 1175 \times 10^3 \times 1 \cdot 5 \text{ newtons.}$$

$$\therefore\ \underline{S = 4 \cdot 15 \text{ m.}}$$

Q 10 (a) State the factors which influence the coefficient of friction for cables being pulled into ducts.

(b) (i) Derive a relationship between the limiting pull, N newtons, on a cable, the number of pairs, n, and the diameter, d millimetres, of each conductor if the stress on copper must not exceed 60 MPa.

(ii) If the cable contained aluminium conductors, what would be the relationship if the stress on aluminium must not exceed 30 MPa?

A 10 (*a*) The sliding coefficient of friction for cables being pulled into ducts varies over a wide range and depends upon several factors, not all of which are always known or controllable; for example,

(*i*) the type of cable,
(*ii*) the type of duct,
(*iii*) correct construction of the duct route,
(*iv*) the cleanliness of the ducts, and
(*v*) the type of lubricant used (if any).

If close accuracy in calculation were required then, due to the above factors, the coefficient of friction (μ) would need to be determined for each case. When pulling-in lead-sheathed cable under normal conditions, values of μ between $0 \cdot 17$ and $0 \cdot 83$ are experienced (average value $0 \cdot 48$). It is safe to assume a value of μ of $0 \cdot 4$–$0 \cdot 5$ for lubricated lead-sheathed cables, and $0 \cdot 3$–$0 \cdot 4$ for unlubricated polyethylene cables.

The actual value of μ depends on the type of duct used. It is general practice to lubricate lead-sheathed cables with petroleum jelly before pulling-in.

(*b*) (*i*) The cross-sectional area of the cable core

$$= \frac{\pi d^2 n}{4} \times 10^{-6} \text{ metres}^2.$$

The stress on the cable

$$= 60 \times 10^6 = \frac{N}{\dfrac{\pi d^2 n}{4} \times 10^{-6}} \text{ pascals.}$$

$$\therefore\ N = \frac{60\pi d^2 n}{4} = \underline{47 \cdot 12 d^2 n \text{ newtons.}}$$

(*ii*) If the cable is aluminium and the limiting stress is 30 MPa, the relationship is

$$\underline{N = 23 \cdot 56 d^2 n \text{ newtons.}}$$

LINE TRANSMISSION C, 1976

Students were expected to answer any 6 questions

Q 1 (a) Write down the expression for the characteristic impedance of a uniform transmission line in terms of its primary coefficients.

(b) Calculate the characteristic impedance of a uniform line at $\omega = 5000$ rad/s, given that $R = 30\ \Omega$/km, $L = 1$ mH/km, $C = 0 \cdot 1\ \mu$F/km and $G = 1\ \mu$S/km.

(c) Show how the characteristic impedance of a line can be obtained from 2 separate impedance measurements.

A 1 (*a*) The characteristic impedance, Z_0, of a uniform transmission line is given by

$$\underline{Z_0 = \sqrt{\frac{R + j\omega L}{G + j\omega C}} \text{ ohms,}}$$

where R is the loop resistance (ohms/kilometre), L is the loop inductance (henrys/kilometre), G is the loop leakance (siemens/kilometre), C is the loop capacitance (farads/kilometre), and ω is the angular frequency (radians/second).

(*b*) From the values given,

$$R + j\omega L = 30 + j5000 \times 10^{-3} = 30 + j5\ \Omega,$$

$$= 30 \cdot 41 \underline{/\tan^{-1} 0 \cdot 1667} = 30 \cdot 41 \underline{/9 \cdot 462^\circ}\ \Omega,$$

and $G + j\omega C = 10^{-6} + j500 \times 10^{-6} = 10^{-6}(1 + j500)\ \Omega$,

$$= 500 \times 10^{-6} \underline{/\tan^{-1} 500} \approx 500 \times 10^{-6} \underline{/90^\circ}\ \Omega.$$

$$\therefore\ Z_0 = \sqrt{\left(\frac{30 \cdot 41}{500 \times 10^{-6}}\right)} \underline{\Big/ \frac{9 \cdot 462^\circ - 90^\circ}{2}}\ \Omega,$$

$$= \underline{246 \cdot 6 \underline{/-40 \cdot 27^\circ}\ \Omega.}$$

(*c*) The characteristic impedance of a transmission line can be calculated from bridge measurements of the input impedance when the far end is short-circuited (Z_{SC}), and when it is open-circuited (Z_{OC}). The characteristic impedance is the geometric mean of the 2 measurements; that is,

$$\underline{Z_0 = \sqrt{(Z_{SC} \times Z_{OC})}.}$$

Q 2 (a) Explain how the attenuation of a transmission line can be reduced by coil loading.

(b) Derive an expression for the cut-off frequency of a loaded line in terms of the inductance of the coils, the spacing between them and the capacitance of the line.

A 2 (*a*) The attenuation of a transmission line is a measure of the power loss along that line, and is expressed as the ratio of the input power to the output power.

The 2 primary coefficients of a cable pair which cause loss of power are the loop resistance, R, and the loop leakance, G. At any point in a cable pair, the voltage, V, and the current, I, are related by the equation $Z_0 = V/I$, where Z_0 is the characteristic impedance of the pair.

The power losses are equivalent to I^2R and V^2G and, in modern pair-type cables, the I^2R losses predominate because G is negligible over the frequency range used.

Therefore, if Z_0 can be increased without increasing the loop resistance, the current will be reduced, so that the I^2R losses, and hence the attenuation, will decrease. (The voltage would rise and increase the V^2G losses but, as previously stated, these losses are negligible.)

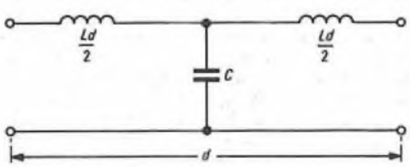
The characteristic impedance is given by the expression

$$Z_0 = \sqrt{\frac{R + j\omega L}{G + j\omega C}},$$

where L is the loop inductance, C is the loop capacitance, and ω is the angular frequency.

Thus, by inserting loading coils at intervals along the line and so increasing the inductance, the characteristic impedance is increased and the attenuation reduced.

(*b*) The sketch shows a transmission line for which the spacing of the loading points is d kilometres. Half the added inductance appears at each end of the line, and L and C are the total loop inductance (henrys/kilometre) and loop capacitance (farads/kilometre) respectively.



At the cut-off frequency, the characteristic impedance is zero. Thus, equating to zero the expression for the characteristic impedance of a T-network:

$$Z_0 = \sqrt{\left\{\frac{Ld}{Cd}\left(1 - \frac{\omega^2 LCd^2}{4}\right)\right\}} = 0.$$

$$\therefore \quad \frac{\omega^2 LCd^2}{4} = 1.$$

$$\therefore \quad \omega = \sqrt{\frac{4}{LCd^2}} = \frac{2}{d \times \sqrt{(LC)}} \text{ radians/second.}$$

$$\text{Hence, the cut-off frequency} = \frac{1}{\pi d \times \sqrt{(LC)}} \text{ hertz.}$$

In practice, the cut off does not occur sharply, and the maximum usable frequency is therefore somewhat lower than the theoretical cut-off frequency.

Q 3 (a) Sketch the circuit arrangement of a bridge suitable for measuring the impedance of a cable pair at audio frequencies.

(b) Derive the modulus and angle of the measured impedance in terms of the bridge components.

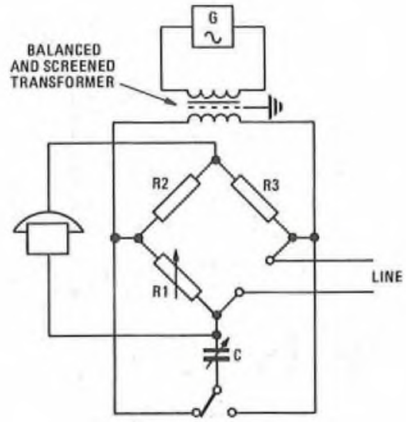
(c) Explain the precautions necessary to ensure a reasonable degree of accuracy in making the measurements.

A 3 (*a*) The circuit diagram of a bridge suitable for making impedance measurements on a cable pair at audio frequencies is shown in the sketch. Variable capacitor C can be connected either across variable resistance R1, when the line impedance has a small negative angle, or across the line itself when its impedance has a small positive angle. A high-impedance telephone receiver acts as the detector, and the signal source is connected to the bridge through a balanced and screened transformer.

(*b*) Consider the bridge to be balanced with capacitor C connected across resistance R1, and let Z_L be the impedance of the line, and Z_{CR1} be the impedance of capacitor C and resistance R1 in parallel. At balance,

$$R_2 Z_L = R_3 Z_{CR1} \text{ or } Z_L = \frac{R_3 Z_{CR1}}{R_2}. \qquad \ldots\ldots (1)$$

$$\text{Now, } Z_{CR1} = \frac{\dfrac{R_1}{j\omega C}}{R_1 + \dfrac{1}{j\omega C}} = \frac{R_1}{1 + j\omega CR_1}.$$



Rationalizing this expression gives

$$Z_{CR1} = \frac{R_1}{1 + j\omega CR_1} \times \frac{1 - j\omega CR_1}{1 - j\omega CR_1},$$

$$= \frac{R_1}{1 + \omega^2C^2R_1^2} - \frac{j\omega CR_1^2}{1 + \omega^2C^2R_1^2}.$$

Substituting for Z_{CR1} in equation (1) gives

$$Z_L = \frac{R_3}{R_2}\left(\frac{R_1}{1 + \omega^2C^2R_1^2} - j\frac{\omega CR_1^2}{1 + \omega^2C^2R_1^2}\right).$$

Expressing this in polar form gives

$$|Z_L| = \frac{R_3}{R_2}\sqrt{\frac{R_1^2 + \omega^2C^2R_1^4}{(1 + \omega^2C^2R_1^2)^2}},$$

$$= \frac{R_3}{R_2}\left\{\frac{R_1}{\sqrt{(1 + \omega^2C^2R_1^2)}}\right\},$$

and $\quad \arg Z_L = \tan^{-1} \omega CR_1.$

If the bridge is balanced with capacitor C connected across the line, it can be shown that

$$Z_L = \frac{R_2R_3R_1}{R_2^2 + \omega^2C^2R_1^2R_3^2} + j\frac{\omega CR_3^2R_1^2}{R_2^2 + \omega^2C^2R_1^2R_3^2},$$

or, in polar form,

$$|Z_L| = \frac{R_3R_1}{\sqrt{(R_2^2 + \omega^2C^2R_1^2R_3^2)}} \text{ and } \arg Z_L = \tan^{-1}\frac{\omega CR_3R_1}{R_2}.$$

(*c*) To ensure the best degree of accuracy, the following precautions are necessary:

(*i*) each component must be screened, and all screens must be connected to a common earth point,

(*ii*) the oscillator must be connected to the bridge through a balanced and screened transformer,

(*iii*) the oscillator must be free from harmonics,

(*iv*) the telephone receiver used as the detector must be of high quality, and

(*v*) all variable components must be properly calibrated.

Q 4 (a) Draw a block diagram showing the requirements for a CCITT 12-channel group for carrier telephony.

(b) Explain the function of each block.

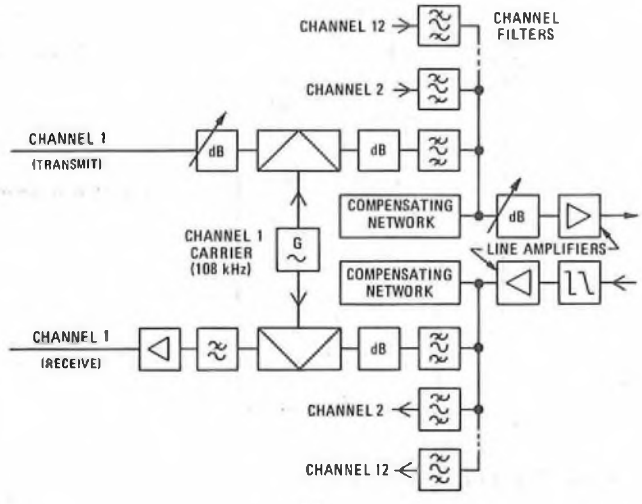
(c) Sketch the attenuation/frequency characteristic of a typical channel filter, and show how it relates to that of an adjacent channel filter.

(d) Explain why each channel filter needs to have a sharp cut-off, and say how this is achieved.

A 4 (*a*) A block diagram of a 12-channel CCITT group for carrier telephony is shown in sketch (*a*).

(*b*) Channel 1 is shown in detail, and channels 2–12 are similar except for the carrier-supply frequency, which changes in steps of 4 kHz from 64 kHz for channel 12 to 104 kHz for channel 2.

For transmission from channel 1, audio signals, applied to the TRANSMIT side, reach the modulator which is fed from the 108 kHz carrier supply. The modulation process gives rise to 2 sidebands, but the channel filter allows only the lower one to pass; the upper side-



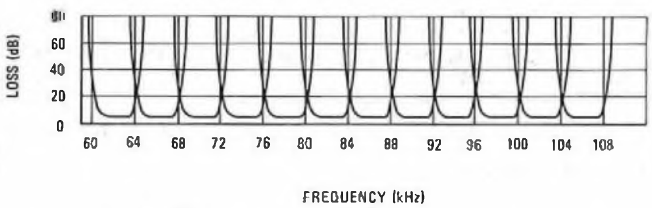
(*a*)

band is rejected. The lower sidebands of all the channels pass to line through the line amplifier.

On the RECEIVE side, each lower sideband is selected by the appropriate channel filter. The channel-1 sideband reaches the demodulator which is fed from the 108 kHz supply, and the demodulation process yields the original audio signal plus an unwanted band far above it in frequency. The unwanted band is rejected by the low-pass filter.

The attenuators ensure that no item of equipment is overloaded, and the amplifiers ensure a satisfactory signal-to-noise ratio. The equalizer ensures that the overall attenuation is independent of frequency. The compensating networks simulate the loading effects of filters for frequencies either side of the first and last channel carriers, so that all the band-pass filters have an equal loading.

(*c*) The attenuation/frequency characteristics of channel filters, and how they relate to each other, are shown in sketch (*b*).



(*b*)

(*d*) The spacing between adjacent pass bands is only 0·9 kHz, and the rate of rise of the attenuation/frequency characteristic of each channel filter has to be very large to allow a discrimination of 72 dB to be achieved. For this reason, crystal filters are used.

Q 5 (*a*) *Explain the need for equalizers in line-transmission systems.*

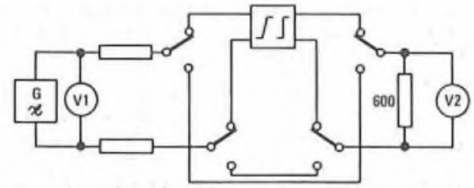
(*b*) *With the aid of sketches, show how the insertion-loss/frequency characteristic of a line equalizer can be measured.*

(*c*) *List the precautions necessary to ensure a reasonable degree of accuracy in measurement.*

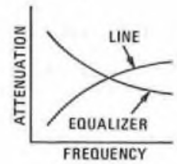
A 5 (*a*) Equalizers are necessary in a line-transmission system to enable the desired attenuation/frequency characteristic to be achieved when flat-gain amplifiers are used. The attenuation of a cable pair rises with frequency. An equalizer is a network whose attenuation/frequency characteristic is the inverse of that of the line. Provided that the equalizer and the line are both properly matched in impedance, the combination of the two in series gives an overall attenuation which is substantially independent of frequency. The principle is illustrated in sketch (*a*). (The same effect can also be achieved by using sloping-gain amplifiers, thus eliminating the need for equalizers.)

(*b*) The insertion loss of a line equalizer is the ratio of the power delivered to the load before the equalizer is connected to the power delivered to the load after it is connected. Sketch (*b*) shows an arrangement by which the loss can be measured directly. At each frequency, 2 measurements are made: one via the equalizer and one via the direct path. Then, assuming the oscillator output to be adjusted to give a constant output voltage, the insertion loss

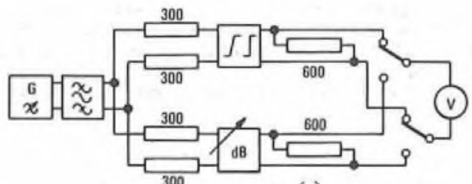
$$= 20 \log_{10} \frac{V_2\,(\text{via direct path})}{V_2\,(\text{via equalizer})} \text{ decibels.}$$



(*b*)



(*a*)



(*c*)

The accuracy of this method depends upon the stability of both voltmeters, and the calibration of voltmeter V2.

Sketch (*c*) shows a method by which the equalizer is compared directly with an adjustable attenuator. The voltmeter is switched alternately between the equalizer path and the attenuator path, while the attenuator is adjusted until the same reading is given for both paths. The accuracy thus depends only upon the calibration of the attenuator; the long-term stability and calibration of the voltmeter are unimportant.

(*c*) For both methods of measurement, it is important to avoid errors caused by harmonics of the testing frequency, and this can be done by putting a band-pass filter in the oscillator's output circuit, or by having a frequency-selective detector in place of the voltmeter. The voltmeter (or detector) should have a high impedance relative to that for which the equalizer is designed.

Q 6 *An audio-frequency 4-wire circuit with 2-wire terminations is to be set up between terminal repeater stations 20 km apart. There is an intermediate station mid-way between them. The circuit is to be lined up to a loss of 3 dB between the 2-wire points. Each pair of the interconnecting cables has an attenuation of 2·5 dB/km at 3·4 kHz and 0·5 dB/km at 0·3 kHz.*

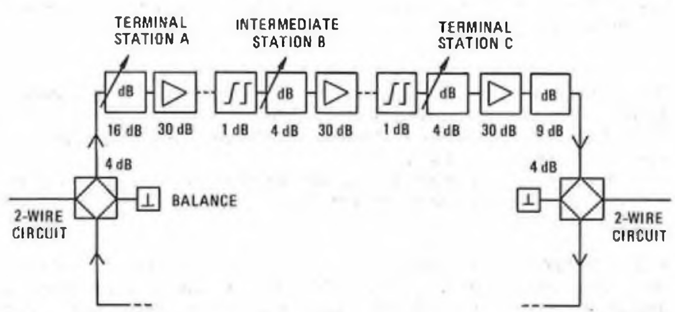
(*a*) *Draw a block diagram for the whole circuit.*

(*b*) *Explain the function of each block.*

(*c*) *Draw a level diagram for one direction of transmission.*

(*d*) *Explain the need to plan for both maximum and minimum signal levels.*

A 6 (*a*) A block diagram of the circuit arrangements for one direction of transmission is shown in sketch (*a*).



(*a*)

(*b*) Connecting the 2-wire circuits to the 4-wire audio circuit are hybrid transformers, the purpose of which is to combine the unidirectional paths of the 4-wire section with the bi-directional paths of the 2-wire sections. The theoretical loss of a hybrid transformer is 3 dB, but an additional 1 dB has been allowed to cater for the transformer loss.

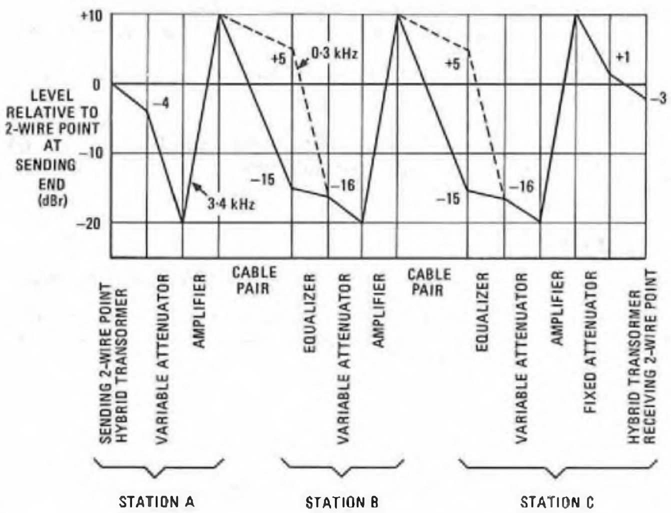
Each repeater comprises a 30 dB fixed-gain amplifier and an adjustable attenuator, the latter set to the values shown.

At stations B and C, line equalizers are provided to make the attenuation of the circuit substantially independent of frequency. Each cable pair can then be considered to have its maximum value of loss (that is, 25 dB) over the entire frequency range. Each equalizer has been assumed to introduce a basic loss of approximately 1 dB at 3·4 kHz.

The 9 dB attenuator between the amplifier and hybrid transformer at station C is necessary to reduce the signal level at the 2-wire point to the specified value.

The whole arrangement is designed to ensure that planned +10 dBr upper and −20 dBr lower limits of signal level are not exceeded. During the lining-up process, minor adjustments to the attenuator settings may have to be made.

(*c*) A level diagram for one direction of transmission is shown in sketch (*b*). The cable loss between stations is 25 dB at 3·4 kHz and 5 dB at 0·3 kHz.



(*b*)

(*d*) It is necessary to limit the maximum signal level at any point to minimize interference with other circuits in the same cable. A minimum signal level is necessary to avoid a poor signal-to-noise ratio.

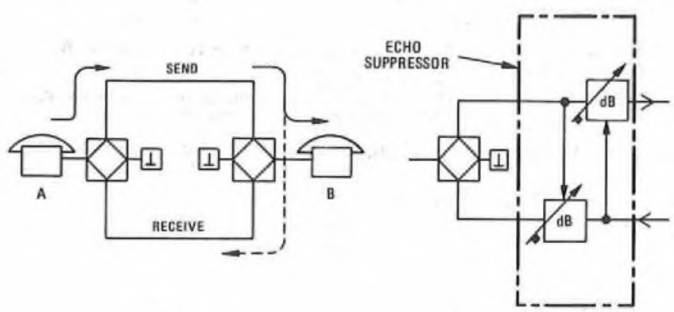
Q 7 (*a*) *Why are echo suppressors essential for some types of circuit?*

(*b*) *Outline the principle of operation of an echo suppressor.*

(*c*) *Give one example of the application of an echo suppressor.*

A 7 (*a*) A 4-wire connexion between 2 subscribers, A and B, is shown in sketch (*a*). Ideally, the loss between the 4-wire RECEIVE and SEND pairs across a terminating unit should be infinite but, in general, this is not the case because of the practical difficulty of making the balancing network a perfect match for the 2-wire line.

Assuming subscriber A to be speaking, and signals to be passing to subscriber B over the path shown by the full-line arrows, any imperfection in the matching of the balancing network at the receiving end causes part of the received signal to be passed back to subscriber A over the path shown by the dashed line. For an electrically-short connexion (that is, one where the delay time is small), subscriber A merely hears this signal as sidetone. However, where the delay time is long (for example, of the order of 200 ms), the signal appears as a distinct echo. Similar conditions apply when subscriber B is talking, and such echoes can be a serious impediment to conversation.



(*a*) (*b*)

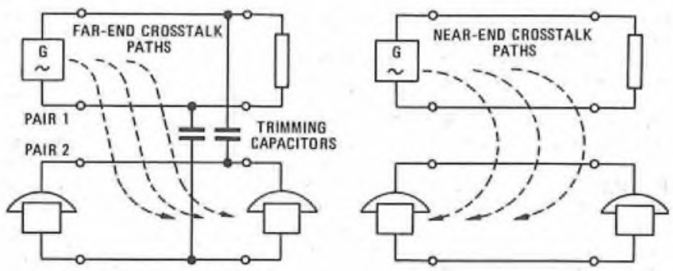
(*b*) Echo suppressors attenuate the return path when activated by speech signals on the forward path. The principle is illustrated in sketch (*b*).

(*c*) Echo suppressors were originally used on long 4-wire audio circuits, but these are now largely superseded by high-velocity carrier circuits, where the delay time is small. The need for echo-suppressors remains, however, on transoceanic-cable and satellite circuits.

Q 8 (*a*) *Explain what is meant by near-end crosstalk and far-end crosstalk.*

(*b*) *Explain how crosstalk can be reduced to an acceptable level on a carrier telephone system using star-quad cables.*

A 8 (*a*) Sketches (*a*) and (*b*) both illustrate 2 pairs in a cable, with a signal in pair 1 causing interference signals in pair 2. The interference signal appearing at the right-hand end of pair 2 is called *far-end crosstalk* (sketch (*a*)), and that appearing at the left-hand end is called *near-end crosstalk* (sketch (*b*)).



(*a*) (*b*)

(*b*) Crosstalk can be caused by

(*i*) capacitance unbalance,
(*ii*) resistance unbalance,
(*iii*) inductive coupling,
(*iv*) low insulation resistance, or
(*v*) wire-to-wire contacts.

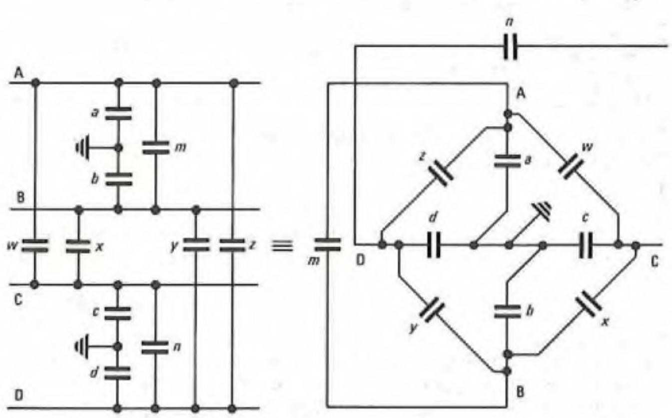
Some of these factors can be minimized by sound design, and by ensuring uniformity during manufacture; others by careful installation and maintenance procedures.

The 4 wires of each quad are taken from the same reel, thus eliminating any possible differences in diameter due to the drawing die, and all are insulated with paper ribbons cut from the same roll. The identification marks, printed on the paper in the form of 1, 2, 3 or 4 rings, are spaced in such a way as to use the same amount of ink in a given length for each wire of the quad, thus obviating any difference in insulation resistance attributable to the ink itself.

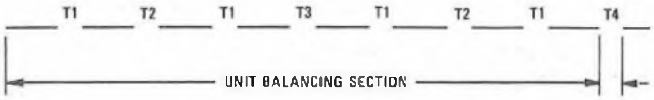
All quads have different lengths of lay, some being laid clockwise and some counter-clockwise. The cable is built up in layers of quads, each layer being laid helically over the one below it and in the opposite direction of lay. Thus, all quads in a layer have the same direction of lay, but adjacent layers have clockwise and counter-clockwise quads. This arrangement of layers and quads minimizes the possibility of inductive coupling between quads. The completed cable core is thoroughly dried in an oven before the sheathing is applied, and it is important that moisture should not be allowed to enter the core, and hence reduce the insulation resistance, during the subsequent jointing operations.

In practice, the main source of crosstalk is capacitance unbalance. It is necessary to make measurements on each cable length, and then to select quads for jointing in such a way as to even out and minimize the overall unbalance figures. Such joints are called *test-selected joints*.

Sketch (*c*) shows the various capacitances involved in a star quad with wires A, B, C and D. No crosstalk occurs within the quad if $w = x = y = z$, $a = b$, $c = d$, and $m = n$.



(*c*)



(*d*)

Sketch (*d*) shows a typical unit balancing section for a carrier cable system. It comprises 8 lengths of cable, each about 160 m long, and every joint is test selected. The cable lengths are first jointed in pairs (joints T1) and then in groups of four (joints T2). Selected joint T3 completes the unit balancing section. A number of such sections are then selectively jointed together (joints T4) to build up the cable to the required length. It is possible to carry out much of the measurement and selection work in the factory before the cable lengths are delivered to the site. However, it is usually necessary to test and select joints T3 and T4 *in situ*.

For a carrier system, it is usual to use separate cables for each direction of transmission. On completion of the balancing process, small trimming capacitors are added at the receiving end, as required, to compensate for any residual unbalance, as shown in sketch (*a*). For this purpose, a special balancing frame is provided at the terminal station. Although there are various paths over which far-end crosstalk can occur, they all have the same electrical length. Thus, all components of the total far-end crosstalk arrive in the same phase, and their effect can be substantially eliminated by the use of simple capacitors. This contrasts with the conditions for near-end crosstalk, where the various paths have different lengths (see sketch (*b*)); thus, the components of the total near-end crosstalk cover a range of phase relationships. Near-end crosstalk is avoided by making use of the screening effect of the lead sheaths.

Q 9 (*a*) *Outline a typical transmission plan for a national telephone network, showing the loss allowable in each part of a subscriber-to-subscriber trunk connexion.*

(*b*) *Explain the advantage of 4-wire switching in the trunk network.*

Q 10 (*a*) *With the aid of sketches, show how power can be provided for the transmission equipment in a repeater station.*

(*b*) *Explain how interruptions due to failure of the mains power supply can be avoided.*

(*c*) *List the precautions necessary to minimize interference from the power supplies.*

A 10 See A7, Line Transmission C, 1975, *Supplement*, Vol. 69, p. 78, Oct. 1976. (See also the correction below.)

CORRECTION

LINE TRANSMISSION C, 1975 (*Supplement*, Vol. 69, Oct. 1976)

A 7 A battery was omitted from the starter-motor circuit in sketch (*b*). The arrangement should have been similar to that shown for the starter motor in sketch (*a*).

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