SUPPLEMENT

TO THE POST OFFICE ELECTRICAL ENGINEERS' JOURNAL

CITY AND GUILDS OF LONDON
INSTITUTE EXAMINATIONS 1973–74

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 will, in future, be published in 32-page and 16-page sizes instead of the 24-page size hitherto maintained.

MATHEMATICS A, 1974 (*continued*)

Q. **10.** (*a*) *A flat is machined to a depth of* 10 *mm on a cylindrical shaft of diameter* 50 *mm. Calculate the perimeter of the remaining cross-section of the shaft.*

(*b*) *A closed cylindrical container,* 4 *m long and* 3 *m in diameter, is to be replaced by a spherical container of equal capacity. Neglecting the thickness of the casing, calculate*

(*i*) *the diameter of the sphere, and*
(*ii*) *the reduction in the total surface area.*

A. **10.** (*a*) The sketch shows the shaft in cross-section, where AB is the flat, and O is the axis of the shaft. Since the flat is 10 mm deep, DC = 10 mm, and the distance, CO, on the line perpendicular to AB from O is 15 mm, the radius, OD, being 25 mm.
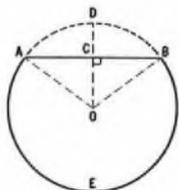
Now,

$$\cos \angle COB = \frac{CO}{OB} = \frac{15}{25} = 0 \cdot 6.$$

$$\therefore \angle COB = 53 \cdot 15° = 0 \cdot 9274 \text{ rad.}$$

Required perimeter, ACBE, = AB + arc AEB,

$$= 2 \times CB + 2 \times \pi \times 25 \times \frac{(2\pi - 2 \times 0 \cdot 9274)}{2\pi},$$

$$= 2 \times 25 \sin 53 \cdot 15° + 25(2\pi - 1 \cdot 8548),$$

$$= 50 \times 0 \cdot 8 + 25 \times 4 \cdot 4284,$$

$$= 150 \cdot 7 \text{ mm.}$$

(*b*) (*i*) Volume of cylinder = volume of sphere.

$$\therefore \pi \times 1 \cdot 5^2 \times 4 = \frac{4}{3}\pi r^3,$$

where r is the radius of the sphere.

$$\therefore r^3 = 3 \times 1 \cdot 5^2 = 6 \cdot 75 \text{ m}^3.$$

$$\therefore r = 1 \cdot 89 \text{ m.}$$

$$\therefore \text{ diameter of sphere} = \underline{3 \cdot 78 \text{ m.}}$$

(*ii*) Surface area of cylinder = $2 \times \pi \times 1 \cdot 5^2 + 2 \times \pi \times 1 \cdot 5 \times 4$,

$$= 51 \cdot 84 \text{ m}^2.$$

Surface area of sphere = $4 \times \pi \times 1 \cdot 89^2$,

$$= 44 \cdot 89 \text{ m}^2.$$

Therefore, reduction in surface area = $51 \cdot 84 - 44 \cdot 89$,

$$= \underline{6 \cdot 95 \text{ m}^2.}$$

RADIO AND LINE TRANSMISSION A, 1974

Students were expected to answer any six questions

Q. **1.** (*a*) *The envelope of a sinusoidally-modulated carrier-wave varies between a maximum value of* ± 8 *volts and a minimum value of* ± 2 *volts. Sketch the waveform and label the axes.*

(*b*) *With reference to your sketch, explain what is meant by "modulation envelope" and "modulation depth".*

(*c*) *Determine, for the above waveform,*

(*i*) *the amplitude of the unmodulated carrier,*
(*ii*) *the amplitude of the modulating signal, and*
(*iii*) *the percentage modulation.*

(*d*) *Briefly explain the purpose of the detector stage in a receiver used or the reception of amplitude-modulated signals.*

A. **1.** (*a*) The sketch shows the required waveform.

(*b*) (*i*) *Modulation Envelope*

When a carrier wave is amplitude-modulated, the amplitude of the carrier signal is caused to vary in accordance with the instantaneous
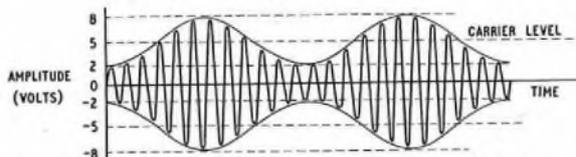
amplitude of the modulating signal. The outline of the modulated carrier waveform is known as the modulation envelope.

(*ii*) *Modulation Depth*

The modulation depth is equal to the amplitude of the modulating signal divided by the amplitude of the unmodulated carrier signal. It is often expressed as the percentage modulation, which is the modulation depth times 100 per cent.

(*c*) Let V_C be the amplitude of the unmodulated carrier (volts), and V_M be the amplitude of the modulating signal (volts). Then, for the waveform in the sketch,

$$V_C + V_M = 8 \text{ volts}, \qquad \ldots \ldots (1)$$

$$\text{and } V_C - V_M = 2 \text{ volts}. \qquad \ldots \ldots (2)$$

(*i*) Adding equations (1) and (2) gives

$$2V_C = 10 \text{ volts}.$$

$$\therefore \underline{V_C = 5 \text{ volts.}}$$

(*ii*) Subtracting equation (2) from equation (1) gives

$$2V_M = 6 \text{ volts.}$$

$$\therefore \underline{V_M = 3 \text{ volts.}}$$

(*iii*) Percentage modulation $= \dfrac{V_M}{V_C} \times 100$ per cent,

$$= \frac{3}{5} \times 100 \text{ per cent,}$$

$$= \underline{60 \text{ per cent.}}$$

(*d*) Both the upper and lower envelopes of the modulated signal contain the required intelligence. It is the function of the detector stage in a receiver to recover the intelligence from the modulated signal, by means of an amplitude detector. In general, a linear detector is used, which performs the operation of detection by rectifying the amplitude-modulated signal such that the mean value of the output current varies at the modulating frequency.

Q. 2. (*a*) *Sketch a typical junction-diode characteristic curve, and label the axes to show the quantities concerned.*

(*b*) *Explain, with reference to the characteristic, the rectifying action of such a diode.*

(*c*) *Draw a diode detector circuit suitable for the demodulation of an amplitude-modulated signal.*

(*d*) *State the purpose of each of the circuit components.*

A. 2. (*a*) A typical junction-diode characteristic curve is shown in sketch (*a*).

(*a*)   (*b*)
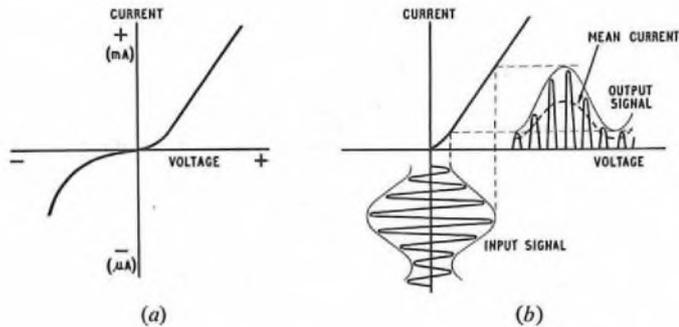
(*b*) A junction diode is known as a linear detector, because it has a nearly-linear current/voltage characteristic over its operating range. Thus, if an amplitude-modulated carrier signal is applied to such a diode, the resulting output is as illustrated in sketch (*b*). The amplitude-modulated signal is rectified by the diode, giving an output waveform having a mean value approximating to the shape of the original modulating signal.

(*c*) Sketch (*c*) shows a typical diode detector circuit suitable for the demodulation of an amplitude-modulated signal.

(*c*)
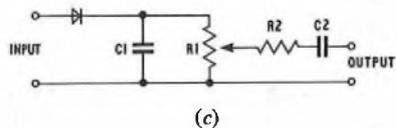
(*d*) The diode rectifies the incoming amplitude-modulated signal, as described in part (*b*). The rectified signal is applied to the resistor-capacitor combination, R1 and C1. By choosing suitable values for resistor R1 and capacitor C1, the envelope of the waveform appearing across them may be made to follow closely the envelope of the original modulated signal applied to the diode. This waveform, besides containing the required modulation envelope, also includes a d.c. component and a residual amount of the radio-frequency (r.f.) carrier signal. Capacitor C1, if it is of a large-enough value, tends to short-circuit the r.f. component, any residual r.f. signal being removed by resistor R2 and the input capacitance of the next stage. Capacitor C2 has a low reactance at the frequencies of modulation, but blocks the d.c. component. Hence, with the r.f. and d.c. components largely eliminated, the output waveform closely resembles the original modulating signal.

Q. 3. (*a*) *Briefly explain the difference in the principles of operation of the moving-coil and the rocking-armature types of telephone receiver.*

(*b*) *List the relative advantages and disadvantages of each type.*

(*c*) *State one application for each type.*

A. 3. (*a*) Sketch (*a*) shows the essential features of the moving-coil type of telephone receiver. It consists of a coil, attached to a diaphragm, and suspended in the strong radial field formed by a permanent-magnet system. When signals from the transmitting end are received in the moving coil, a magnetic field is set up around the coil which interacts with the radial magnetic field. Hence, a force is exerted on the coil which makes it move axially within the annular air gap. This, in turn, causes the diaphragm to move in sympathy with the signal current, to produce sound waves. The moving-coil telephone receiver, therefore, operates on the principle of the electric motor.

(*a*)   (*b*)
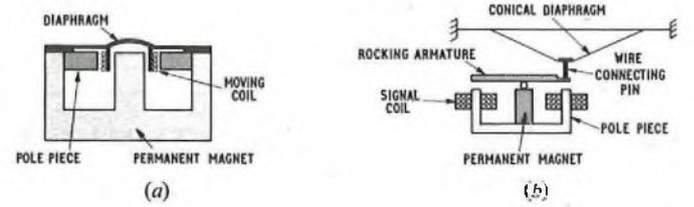
Sketch (*b*) shows the essential features of the rocking-armature type of telephone receiver. It consists of a rocking armature connected to a conical diaphragm. The armature is held in equilibrium, when no signals are present, by the equal attraction of the two pole pieces of a permanent-magnet system. When signals from the transmitting end are received in the two coils wound on the pole pieces, the magnetic balance is upset, and the armature tends to rock in sympathy with the variations in signal current. This movement is transmitted to the conical diaphragm, which is made from a non-magnetic material, to produce sound waves. The rocking-armature telephone receiver works on the principle of a magnetic unbalance resulting from the magnetic fields produced by signal currents flowing in two coils.

(*b*) The relative advantages and disadvantages of each type of receiver are given in the table.

Receiver Type	Moving coil	Rocking armature
Sensitivity	Low	High
Reliability	Good	Good
Frequency Response	Generally good over a wide range of frequencies	Good over commercial speech frequencies
Cost	Fairly expensive	Relatively cheap

(*c*) The rocking-armature type of receiver is used in public telephone systems.

The moving-coil type of receiver is used where higher-quality reproduction is required, and may also be used as a microphone.

Q. 4. (*a*) *List the factors which largely determine the choice of dielectrics for the various types of fixed and variable capacitors used in communication equipment.*

(*b*) *Give one application for each of the following types of capacitor in a radio receiver:*

   (*i*) *air dielectric,*
   (*ii*) *mica dielectric,*
   (*iii*) *paper dielectric, and*
   (*iv*) *electrolytic.*

(*c*) *State typical capacitance values and working voltages for each type.*

A. 4. (*a*) The choice of dielectric is largely determined by

   (*i*) the value of the capacitor,
   (*ii*) the working voltage,
   (*iii*) the permitted tolerance of the capacitance value,
   (*iv*) the permitted stability of the capacitance value,
   (*v*) the dielectric losses,
   (*vi*) the size and weight of the capacitor,
   (*vii*) the cost, and
   (*viii*) the function of the capacitor, which may involve special constructional features.

(*b*) (*i*) Air-dielectric capacitors are used in the radio-frequency (r.f.) tuning sections of a radio receiver.

(*ii*) Mica-dielectric capacitors are used for coupling and decoupling purposes in the r.f. sections of a radio receiver.

(*iii*) Paper-dielectric capacitors are used for coupling, d.c. blocking and decoupling purposes in the audio-frequency (a.f.) sections of radio receivers.

(*iv*) Electrolytic capacitors are used for coupling and decoupling purposes in the a.f. sections of radio receivers, and as reservoir and smoothing capacitors in the power supplies of mains-operated receivers.

(*c*) Typical capacitance values and working voltages are given in the table.

Dielectric	Capacitance Range	Working Voltage
Air	10–2,000 pF	10 kV
Mica	10–1,000 pF	10 kV
Paper	0·0001–40 μF	25 volts
Electrolyte	1–10,000 μF 1–50 μF	25 volts 350 volts

Note: In a radio receiver, working voltages as high as those quoted for air and mica dielectrics would not be expected.

Q. **5.** (*a*) *Briefly explain why all radio transmissions, and most long-distance line transmissions, involve the use of a carrier.*

(*b*) *What particular advantage arises from the use of a carrier in line transmission?*

(*c*) *Explain, with reference to an impedance/frequency or a voltage/frequency characteristic curve, how parallel resonant circuits may be used to select narrow-band transmissions centred on carriers of different frequencies.*

(*d*) *A parallel tuned circuit consists of an inductance of 490 mH and a capacitance of 1,600 pF. Determine the frequency of resonance.*

A. **5.** (*a*) If two or more audio signals are passed over a single pair of wires, there will be mutual interference. If, therefore, it is required to send more than one audio signal over a pair of wires, the different signals must appear in separate frequency bands. A method of achieving this is to amplitude-modulate a series of radio-frequency (r.f.) carriers with the signals, so translating the signals to different parts of the frequency spectrum.

For radio transmission, in addition to the need to separate signals to avoid mutual interference, it is necessary to use a high-frequency carrier because audio frequencies are propagated over only very short distances in free space. Therefore, the carrier frequencies are chosen, among other reasons, to provide optimum propagation conditions for the service concerned.

(*b*) The particular advantage which arises from the use of a carrier, in line transmission, is that a single pair of wires may be used to carry a large number of communication circuits.

(*c*) The sketch shows a typical impedance/frequency characteristic curve for a parallel tuned circuit near resonance. The impedance is a maximum at the frequency of resonance, F_R, and decreases in value each side of F_R.
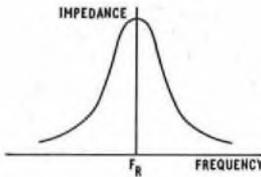
This characteristic of a parallel tuned circuit may be used, in an r.f. amplifier, to select narrow-bandwidth transmissions centred on carriers of different frequencies. The parallel tuned circuit is connected as the anode load of the amplifier. Since the voltage-amplification factor of an r.f. amplifier is almost directly proportional to the value of its anode load impedance, any desired signal is selected by tuning the parallel tuned circuit to resonance at the frequency of the desired signal, thus providing maximum amplification at that frequency.

(*d*) Now, $F_R = \dfrac{1}{2\pi\sqrt{(LC)}}$ hertz, where L is the inductance (henrys), and C is the capacitance (farads).

$$\therefore F_R = \frac{1}{2\pi\sqrt{(490 \times 10^{-3} \times 1,600 \times 10^{-12})}} \text{ Hz},$$

$$= \underline{5\cdot68 \text{ kHz}.}$$

Q. **6.** (*a*) *Briefly explain why logarithmic units are used for expressing the ratios of powers, currents and voltages in radio and line communication.*

(*b*) *Define the decibel.*

(*c*) *Use this definition to derive an expression for voltage ratios in decibel notation. State the condition for which the expression is valid.*

(*d*) *An amplifier has input and output load resistances of 600 ohms. The input signal voltage is +18 dB relative to 1 μV, and the amplifier has a gain of 25 dB. Determine*

(*i*) *the input and output voltages, and*

(*ii*) *the output power in dB relative to 1 mW.*

A. **6.** (*a*) The principal advantages of using logarithmic units for expressing the ratios of powers, currents and voltages, in radio and line communication, are as follows.

(*i*) The enormous ranges of powers, currents and voltages involved in radio- and line-communication work can be expressed in simpler numbers.

(*ii*) In complicated circuits, involving several sections in tandem, such as amplifiers, attenuators and lines, it is a relatively simple matter to determine the overall output-to-input ratios of the powers, voltages and currents, by taking the sum of the logarithmic ratios for the different parts of the circuits.

(*iii*) A further point, of academic interest, is that the effect of sound intensity on the ear is approximately proportional to the logarithm of the sound intensity.

(*b*) If the ratio of two powers, P_1 and P_2, is to be expressed in decibels, the number of decibels, N, is given by

$$N = 10 \log_{10} \frac{P_1}{P_2} \text{ dB}.$$

(*c*) The power, P, dissipated in a resistance, R, is given by

$$P = \frac{V^2}{R} \text{ watts, where } V \text{ is the voltage.}$$

Hence, for two resistances, R_1 and R_2, having voltages V_1 and V_2, respectively,

$$N = 10 \log_{10} \frac{V_1^2/R_1}{V_2^2/R_2} \text{ dB}.$$

If $R_1 = R_2$, then $N = 10 \log_{10} \dfrac{V_1^2}{V_2^2}$ dB,

$$= 10 \log_{10} \left(\frac{V_1}{V_2}\right)^2 \text{ dB},$$

$$= \underline{20 \log_{10} \frac{V_1}{V_2} \text{ dB}.}$$

(*d*) (*i*) Let V_{in} be the input voltage in microvolts.

Then, $\qquad 18 = 20 \log_{10} \dfrac{V_{in}}{1}.$

$\therefore \log_{10} V_{in} = 0\cdot9.$

$\therefore \underline{V_{in} = 7\cdot943 \ \mu\text{V}.}$

Now, output $= 18 + 25 = 43$ dB relative to 1 μV.

Let V_{out} be the output voltage in microvolts.

Then, $\qquad 43 = 20 \log_{10} \dfrac{V_{out}}{1}.$

$\therefore \log_{10} V_{out} = 2\cdot15.$

$\therefore \underline{V_{out} = 141\cdot3 \ \mu\text{V}.}$

(*ii*) Output power $= \dfrac{(141\cdot3 \times 10^{-6})^2}{600}$ watts,

$$= 3\cdot33 \times 10^{-8} \text{ mW},$$

$$= 10 \log_{10} \frac{3\cdot33 \times 10^{-8}}{1} \text{ dB, relative to 1 mW},$$

$$= \underline{-74\cdot78 \text{ dB, relative to 1 mW}.}$$

Q. **7.** (*a*) *Draw the circuit diagram of a tuned-collector transistor oscillator.*

(*b*) *Describe the function of each section of the circuit in maintaining oscillations of constant amplitude.*

(*c*) *Briefly explain one method of obtaining self-bias in oscillators.*

A. 7. (*a*) Sketch (*a*) shows the circuit of a tuned-collector transistor oscillator.

(*b*) The frequency-determining tuned circuit, inductor L2 and capacitor C2, is connected in the collector circuit of the transistor. The frequency of oscillation, *f*, is approximately given by

$$f = 1/\{2\pi\sqrt{(L_2C_2)}\}.$$

Part of the output from the collector is fed back into the base circuit via inductor L1. The transistor, acting as a common-emitter amplifier, is arranged, by means of the potential divider R1 and R2 and the emitter resistor R3, to compensate for any circuit losses and, so, produce continuous oscillations of constant amplitude.

Capacitor C4 is a d.c. blocking capacitor, and provides the point at which the output may be obtained.

(*a*)   (*b*)
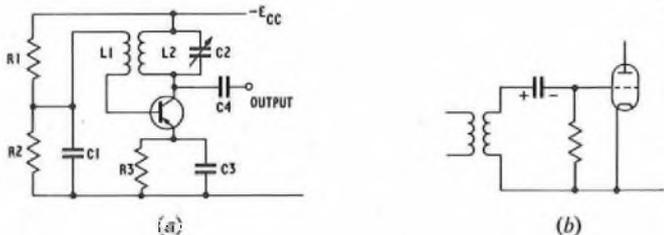
(*c*) The principle of self-bias in oscillators may be explained by reference to the *leaky-grid* circuit shown in sketch (*b*).

Initially, the grid-bias voltage is zero. On the first positive half-cycle of input signal, the resulting grid current charges the capacitor in the direction shown. During the negative half-cycle, no grid current flows, and the capacitor starts to discharge. By choosing suitable values of capacitance and resistance, a steady condition is achieved, where the grid current flowing during the positive half-cycle is just sufficient to restore the charge lost during the negative half-cycle. Thus, the circuit becomes self-biased, and oscillations of constant amplitude are produced.

Q. **8.** (*a*) *Show, by means of a circuit diagram, the equipment required to determine the static characteristics of a transistor in the common-emitter configuration.*

(*b*) *Describe the procedure for obtaining the collector-voltage/ collector-current characteristics for various values of base current. Draw a typical family of characteristics.*

(*c*) *Briefly explain how to derive the output resistance and short-circuit current gain from the above characteristics.*

Q. **9.** (*a*) *The data in the table refers to a thermionic triode.*

Anode Voltage	Anode Current (mA) for grid voltage =		
	0	−5	−10
80	17	1	—
120	27	11	—
160	37	22	6
200	47	32	16
240	—	—	25

Plot the static anode characteristics for grid voltages of 0 volts, −5 volts, and −10 volts.

(*b*) *The triode is used with an h.t. supply of 200 volts, a pure resistance load of 4 kohms and a grid bias of 5 volts. Draw the load line, and determine the power dissipated at the anode.*

(*c*) *For a sinusoidal voltage input, which has a peak value of 5 volts, determine*

(*i*) *the maximum and minimum values of anode current, and*

(*ii*) *the maximum and minimum values of anode voltage.*

Q. **10.** (*a*) *Draw circuit diagrams showing how a junction transistor may be used in a single-stage, resistance-loaded, audio-frequency amplifier in the*

(*i*) *common-base configuration, and*

(*ii*) *common-emitter configuration.*

In each case, mark clearly the biasing arrangements and the directions of the currents.

(*b*) *Prepare a table comparing the current gain, input impedance and output impedance of the two types of amplifier in part (a) above.*

(*c*) *Briefly outline the principle of operation of one of the circuits.*

A. 10. (*a*) (*i*) and (*ii*) Sketches (*a*) and (*b*) show the circuit diagrams of a single-stage, resistance-loaded, audio-frequency amplifier in the common-base and common-emitter configurations, respectively. The arrows indicate the directions of the emitter, collector and base currents (I_E, I_C and I_B, respectively).

(*a*)   (*b*)
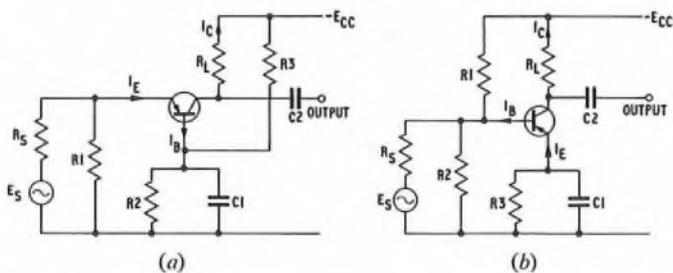
(*b*) The table compares the current gain, input impedance and output impedance of the common-base and common-emitter amplifier configurations.

	Common-Base Configuration	Common-Emitter Configuration
Current Gain	Approximately unity	30–100
Input Impedance	Low (30–100 ohms)	Medium (200–5,000 ohms)
Output Impedance	High (10^5–10^6 ohms)	High (10^4–10^5 ohms)

(*c*) *Common-Base Configuration*

Referring to sketch (*a*), the emitter–base junction is forward biased, and the collector–base junction is reverse biased, by the resistors R2 and R3, which form a potential divider across the collector supply voltage, −E_{CC}. When an alternating source voltage, E_S, is connected to the input, it varies the forward bias applied to the emitter–base junction. During positive swings of E_S, the forward bias is increased, and, during negative swings, it is decreased. Thus, the emitter current varies about its nominal value, and, since the collector current is approximately equal to the emitter current ($I_C/I_E = 0.98$, typically), it also varies in accordance with E_S.

The current gain, α, is given by the change in collector current, δI_C, divided by the change in emitter current, δI_E, and is marginally less than unity. However, since approximately the same current flows in the input resistance, R_{in}, and the load resistance, R_L, then, if $R_L \gg R_{in}$, a considerable voltage gain is obtained.

$$\text{Voltage gain} = \frac{\delta I_C R_L}{\delta I_E R_{in}} = \alpha \frac{R_L}{R_{in}} = 0.98 \frac{R_L}{R_{in}}, \text{ typically.}$$

Similarly, large power gains are also obtained.

$$\text{Power gain} = \frac{(\delta I_C)^2 R_L}{(\delta I_E)^2 R_{in}} = \alpha^2 \frac{R_L}{R_{in}} = 0.96 \frac{R_L}{R_{in}}, \text{ typically.}$$

TELECOMMUNICATION PRINCIPLES B, 1973

Students were expected to answer any 6 questions

Q. **1.** (*a*) *Explain the meanings of the terms reactance and impedance.*

(*b*) *An impedance, carrying a current at a frequency of 12 kHz, consists of a negative reactance of 20 Ω in series with a resistance of 15 Ω. Calculate the value of the reactive component.*

(*c*) *Draw the phasor diagram to represent this impedance, and express it in the notation a + jb.*

(*d*) *For a current of 30 mA, find*

(*i*) *the voltage across the impedance, and*

(*ii*) *the phase angle between the current and this voltage.*

A. 1. (*a*) When a pure inductance or capacitance is connected across a sinusoidal supply, the ratio of the applied voltage to the current in the circuit is termed the reactance. Reactance is the opposition offered by an inductance or a capacitance to the flow of alternating current; it is frequency-dependent and is measured in ohms. Inductive reactance, X_L, is proportional to frequency, and is given by $X_L = 2\pi f L$, where *f* is the frequency (Hz), and *L* is the inductance (H). Capacitive reactance, X_C, is inversely proportional to frequency, and is given by $X_C = 1/(2\pi f C)$, where *C* is the capacitance (F).

When a circuit consisting of resistance and reactance is connected across a sinusoidal supply, the ratio of the applied voltage to the current taken by the circuit is termed the impedance of the circuit. Impedance is the opposition offered to the flow of alternating current by such a circuit, and is measured in ohms. The modulus of the impedance, $|Z|$, of a circuit consisting of a resistance, R, in series with a reactance, X, is given by $|Z| = \sqrt{(R^2 + X^2)}$, and the phase angle, ϕ, between the current and the voltage is given by $\phi = \tan^{-1}(X/R)$.

(*b*) Since the reactance is negative, the reactive component is a capacitance.

From part (*a*), $C = \dfrac{1}{2\pi f X_C}$,

$$= \frac{1}{2 \times \pi \times 12 \times 10^3 \times 20} \text{ F} = \underline{0 \cdot 663 \ \mu\text{F}.}$$

(*c*) The sketch shows the phasor diagram. OA represents the resistance, in phase with the current, I, and OB represents the capacitive reactance, lagging the current by 90°. OC then represents the impedance of the circuit.
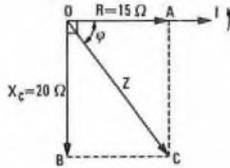
In $a + jb$ notation, $\underline{Z = 15 - j20 \ \Omega.}$

(*d*) (*i*) Now, $\quad |Z| = \sqrt{(15^2 + 20^2)} = 25 \ \Omega.$

The voltage across the impedance $= IZ$,

$$= 30 \times 10^{-3} \times 25 = \underline{0 \cdot 75 \text{ V}.}$$

(*ii*) $\qquad\qquad \phi = \tan^{-1}\dfrac{20}{15} = \underline{53° \ 8'.}$

Q. 2. (*a*) *Define the h-parameters for a transistor connected in the common-base configuration.*

(*b*) *Describe an experiment to obtain the static characteristic curves relating these parameters for a transistor connected in the common-base configuration. Give a circuit diagram and details of the equipment needed, and describe the procedure. Include typical curves with values on the axes, and derive from these curves values for input and output resistances.*

A. 2. (*a*) The *h*-parameters (hybrid parameters), for a transistor connected in the common-base configuration, are related to the gradients of the 4 main d.c. static characterististic curves for that transistor. They describe the low-frequency, small-signal performance of the transistor in terms of the input and output alternating signal voltages (V_{eb} and V_{cb} respectively) and currents (I_e and I_c respectively).

The 4 *h*-parameters are the input parameter (h_i), the output parameter (h_o), the forward current gain (h_f), and the reverse voltage ratio (h_r). A second suffix, *b*, to each parameter denotes that the transistor is connected in the common-base configuration.

The *h*-parameters are defined as follows.

(*i*) $h_{ib} = \dfrac{V_{cb}}{I_e}$, with the output short-circuited; i.e., $V_{cb} = 0$.

(*ii*) $h_{ob} = \dfrac{I_c}{V_{cb}}$, with the input open-circuited; i.e., $I_e = 0$.

(*iii*) $h_{fb} = \dfrac{I_c}{I_e}$, with the output short-circuited; i.e., $V_{cb} = 0$.

(*iv*) $h_{rb} = \dfrac{V_{eb}}{V_{cb}}$, with the input open-circuited; i.e., $I_e = 0$.

(*b*) The static characteristics, for a transistor in the common-base configuration, are plotted from measurements of the direct input and output voltages (V_{EB} and V_{CB} respectively) and currents (I_E and I_C respectively). The 4 characteristics, with their related static *h*-parameters, are listed below.

(*i*) The input characteristic shows the graph of I_E/V_{EB} for a fixed value of V_{CB}. The gradient of the curve gives the input admittance of the transistor, the reciprocal of which is the input parameter, h_{IB}.

(*ii*) The output characteristic shows graphs of I_C/V_{CB} for a series of fixed values of I_E. The gradient of the curves gives the output admittance of the transistor, or the output parameter, h_{OB}.

(*iii*) The transfer characteristic shows the graph of I_C/I_E for a

fixed value of V_{CB}. The gradient of the curve gives the forward current gain, h_{FB}.

(*iv*) The feedback characteristic shows graphs of V_{EB}/V_{CB} for a series of fixed values of I_E. The gradient of the curves gives the reverse voltage ratio, h_{RB}.

Sketch (*a*) shows a circuit diagram suitable for obtaining the static characteristics of a transistor connected in the common-base configuration.

(*a*)
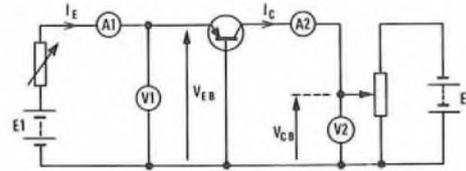
The emitter–base junction is biased by battery E1, the bias being adjusted by means of the variable resistor. The collector–base junction is biased by battery E2, the bias being adjusted by means of the potentiometer. Voltmeters V1 and V2 record V_{EB} and V_{CB} respectively, and must be high-impedance instruments, so that the circuit conditions are not disturbed. Ammeters A1 and A2 record I_E and I_C respectively.

To obtain the input characteristic, V_{CB} is set to a convenient value and maintained at that value by adjusting the potentiometer as necessary. Values of I_E are recorded for a range of values of V_{EB}, obtained by means of the variable resistor.

To obtain the output characteristic, I_E is set to a convenient value and maintained at that value by adjusting the variable resistor as necessary. Values of I_C are recorded for a range of values of V_{CB}, obtained by means of the potentiometer. The measurements are repeated for a range of fixed values of I_E.

To obtain the transfer characteristic, V_{CB} is maintained at a constant value, while values of I_C are recorded for a range of values of I_E.

To obtain the feedback characteristic, I_E is maintained at a constant value, while values of V_{EB} are recorded for a range of values of V_{CB}. The measurements are repeated for a range of fixed values of I_E.

Typical input, output, transfer and feedback characteristics are shown in sketches (*b*), (*c*) (*d*) and (*e*) respectively. (Note that, by
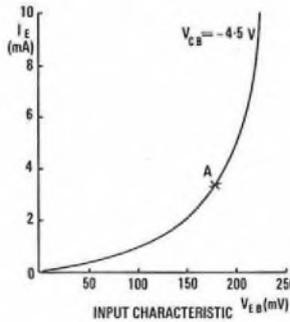

(*b*)
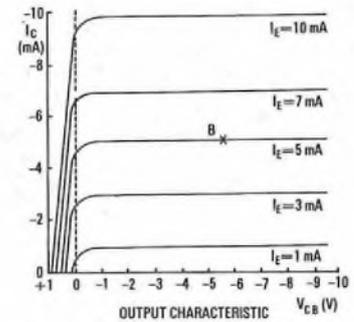
(*c*)
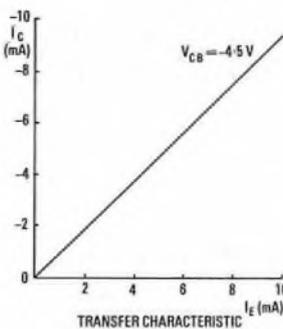
(*d*)
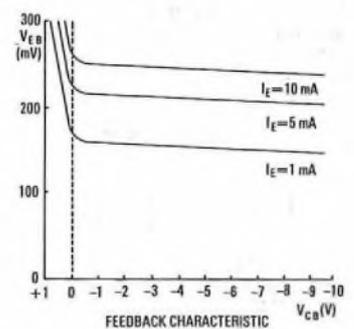
(*e*)

convention, the current flowing into the transistor (I_E) is considered to be positive, and the current flowing out of the transistor (I_C) is considered to be negative.)

The input resistance is given by the reciprocal of the gradient of the nearly-linear region of the input characteristic (about point A). A typical value is approximately $\underline{70 \ \Omega.}$

5

The output resistance is given by the reciprocal of the gradient of the linear region of the output characteristic (about point B). A typical value is approximately 1 MΩ.

Note: Values for *h*-parameters, derived from the static characteristics, are valid provided that the characteristics are within the limits of small-signal operation, even though the open-circuit input and short-circuit output conditions are replaced by fixed-value conditions. For example, considering the output characteristic, small changes in I_C and V_{CB} behave in the same way as alternating components, I_c and V_{cb}, over the linear region of the characteristic. As these changes occur for a fixed value of I_E, then $I_e = 0$, and the open-circuit input condition is satisfied.

Q. **3.** (*a*) *Explain the principle of operation of a d.c. motor. Show how a unidirectional torque is produced, and state the factors that determine the magnitude of this torque.*

(*b*) *Give simple circuit diagrams for d.c. motors with*

(*i*) *shunt-connected field coils, and*
(*ii*) *series-connected field coils.*

Name one application for each type.

A. **3.** (*a*) Sketch (*a*) illustrates the principle of the d.c. motor. The armature is connected to a d.c. supply, via a commutator, and rotates due to the interaction of the magnetic field produced by the magnetic poles with the field set up around the armature by the current flowing in it.

The commutator ensures that the armature rotates continuously in one direction, by reversing the current flowing in the armature twice per revolution. The commutator is a segmented ring, in contact with fixed, spring-mounted carbon brushes, which slide over the surface of the commutator as it rotates. Each side of the armature coil is connected to one of the segments and, hence, is alternately connected to the positive and negative sides of the supply for half of every revolution. The reversals are arranged to occur when the armature is in its vertical, or neutral, position; that is, when the sides of the armature coil are changing their directions of travel across the magnetic field.

Considering the motor at the position shown in sketch (*a*), current is flowing in the armature in the direction indicated by the arrows. By Fleming's left-hand rule, the armature rotates in the clockwise direction. When the armature reaches the vertical position, the commutator reverses the current flowing in it. Hence, as the sides of the armature coil re-enter the magnetic field, the direction of the current is correct for rotation to continue in the clockwise direction. Thus, a unidirectional torque is produced.

In a practical motor, the armature may consist of a number of distributed windings, arranged as coils across a cylindrical drum, giving 2 or more parallel paths in the armature. The coils are wound in lateral slots in the drum, parallel to the axis of rotation, the drum being mounted on a shaft, to which the load is connected. The motor may also have a number of pairs of magnetic poles, usually produced electromagnetically by means of field coils.

The torque, *T*, produced by a d.c. motor, is given by

$$T = \frac{ZP\Phi I_a}{2\pi C} \text{ newton metres,}$$

where *Z* is the total number of conductors on the armature, *P* is the number of poles, Φ is the total flux/pole (Wb), I_a is the armature current (A), and *C* is the number of parallel paths in the armature.

(*b*) (*i*) The circuit diagram for a d.c. motor with a shunt-connected field coil is shown in sketch (*b*). The field coil is connected directly to

(*a*)   (*b*)   (*c*)
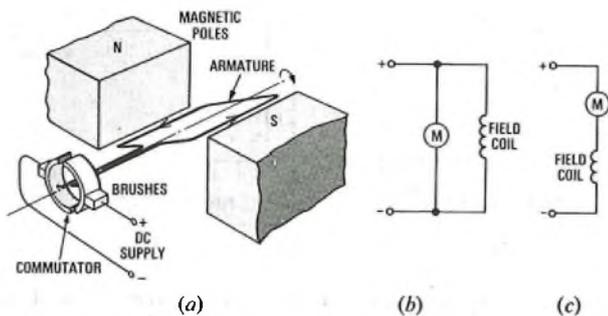
the supply, in parallel with the armature, and the field's magnetic flux is, therefore, independent of the armature current and the load. This gives a low starting torque, but the motor tends to maintain a constant speed, almost independent of the load. This type of motor is, therefore,

useful for driving machines where a constant speed is important; for example, printing machines.

(*ii*) The circuit diagram for a d.c. motor with a series-connected field is shown in sketch (*c*). The field coil is connected in series with the armature, so that the field's magnetic flux is proportional to the armature current. This gives a very high starting torque, since the back e.m.f. is zero when the armature is stationary, and the starting current is, therefore, unopposed. The speed is widely variable, being dependent on the load, and the stalling torque is high. This type of motor is, therefore, useful as a traction motor.

Q. **4.** (*a*) *Describe an experiment to plot the response of each of the following circuits against a range of frequencies through resonance:*

(*i*) *an inductor in series with a capacitor, and*
(*ii*) *an inductor in parallel with a capacitor.*

Show how to find the frequency of resonance in each case.

Set out the answer in the form of an experimental report, and include comments on the difference between the behaviour of the 2 circuits.

(*b*) *In relation to part (a), comment on the importance of the 3 dB points.*

Q. **5.** (*a*) *With the aid of a circuit diagram, explain the operation of an electronic voltmeter for alternating quantities. Comment especially on the input circuit.*

(*b*) *What is the field of use of this type of voltmeter?*

(*c*) *Describe briefly an experiment in which an electronic voltmeter is used to measure current in a circuit.*

A. **5.** See A.4, Telecommunication Principles B, 1972, *Supplement*, Vol. 66, p. 23, Apr. 1973.

Q. **6.** (*a*) *State Faraday's law of electromagnetic induction.*

(*b*) *Use this law to explain*

(*i*) *the property of inductance in an inductor, and*
(*ii*) *the factors in the inductor that will affect its inductance value.*

(*c*) *An air-cored solenoid with 800 turns has a length of 0·1 m and a cross-sectional area of 20 cm². Calculate the inductance.*

A. **6.** (*a*) Faraday's law of electromagnetic induction states that, when the magnetic flux linking a circuit is changing, an e.m.f. is induced in the circuit with a magnitude proportional to the rate of change of flux linkage.

For a magnetic flux, Φ webers, linking a coil of *N* turns (i.e. $N\Phi$ flux linkages), the e.m.f., *e*, induced in the coil when the flux changes at a rate $d\Phi/dt$, where *t* is the time in seconds, is given by

$$e = -N\frac{d\Phi}{dt} \text{ volts.}$$

(*b*) (*i*) When a current flows in a coil, a magnetic flux is set up which must link with the turns of that coil. If the current changes, the flux changes, resulting in a change in flux linkage. Therefore, an e.m.f. is induced in the coil, in such a direction as to tend to oppose the change in the current. The property of the coil which causes the induced e.m.f. is called the *self-inductance* or, simply, the *inductance* of the coil, and is measured in henrys. Whereas the resistance of a circuit tends to oppose the current flowing in it, the inductance of a circuit tends to oppose a change in the current flowing.

(*ii*) The inductance, *L*, of a long solenoid or toroid is given by

$$L = \frac{\mu_r \mu_0 N^2 a}{l}$$ henrys, where μ_0 is the absolute permeability of free space, equal to $4\pi \times 10^{-7}$ H/m.

Hence, the inductance is proportional to the square of the number of turns on the coil, the cross-sectional area, *a* metre², of the coil, and the relative permeability, μ_r, of the core, and inversely proportional to the length, *l* metres, of the coil.

(*c*) For an air-cored solenoid, $\mu_r = 1$. Substituting in the equation in part (*b*) (*ii*) gives

$$L = \frac{1 \times 4\pi \times 10^{-7} \times 800^2 \times 20 \times 10^{-4}}{0·1} \text{ H,}$$

$$= 16·1 \text{ mH.}$$

Q. **7.** (*a*) *Explain the meanings of*

(*i*) *relative magnetic permeability,*
(*ii*) *reluctance, and*
(*iii*) *magnetic leakage factor,*

as applied to a magnetic circuit.

(*b*) *Explain how factors (i), (ii) and (iii) above are needed in the*

calculation of the magnetizing force required to produce a given magnetic flux in an iron ring.

(c) The magnetic circuit of a generator is symmetrical about the centre line, XY, and has the mean dimensions shown in Fig. 1 for its magnetic path. The cross-sectional area of the iron is 1 cm², except for the centre limb which is 2 cm². Each air gap is 1 mm long, with a gap flux of 0·15 mWb. The relative permeability of the iron under these conditions is 600. The magnetic leakage factor is 1·25. Calculate the total magnetomotive force (m.m.f.) produced by the field coils wound on pole pieces A and B.
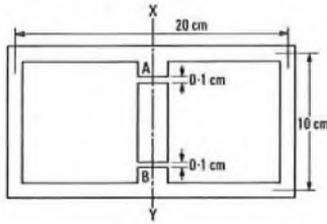
Fig. 1

A. 7. *(a) (i)* Consider a solenoid carrying a given current and having an air core. A certain magnetic flux is set up in the core. If the air core is replaced by an iron core, the magnetic flux in the iron is considerably greater than that previously set up in the air core. The ratio of the magnetic flux in the iron core to that in the air core, for the given current, is called the relative permeability, μ_r, of the iron.

(ii) The reluctance of a magnetic circuit is the ratio of the magneto-motive force acting in the circuit to the resulting magnetic flux, and has the units amperes/weber.

(iii) In a magnetic circuit, part of the magnetic flux follows a path in which it is ineffective for the purpose for which it is intended. This flux is known as the magnetic leakage, and is the difference between the total flux produced and the useful flux. The magnetic leakage factor is the ratio of the total flux to the useful flux.

(b) Consider an iron ring of mean circumference l metres and cross-sectional area a metres², having a relative permeability μ_r. The useful magnetic flux required in the ring is Φ_U webers, and the leakage factor is k. The factors listed in part (a) are used to calculate the magnetizing force in the following manner.

(i) The total flux required, Φ_T webers, is calculated from the leakage factor, using the formula $\Phi_T = \Phi_U k$.

(ii) The reluctance, S amperes/weber, of the iron ring is calculated from the formula $S = (l/\mu_0\mu_r a)$, where μ_0 is the absolute permeability of free space, equal to $4\pi \times 10^{-7}$ H/m.

(iii) The m.m.f., F amperes, required to produce Φ_T is calculated from the formula $F = \Phi_T S$.

(iv) The magnetizing force, H amperes/metre, is calculated from the formula $H = F/l$.

(c) The magnetic circuit shown in Fig. 1 consists of 2 equal paths in parallel, symmetrical about XY. Therefore, the m.m.f. required to produce a given magnetic flux in the whole circuit is the same as that required to give half of that flux in one of the symmetrical parallel paths.

Leakage factor $= 1 \cdot 25 = \Phi_T/\Phi_U = \Phi_T/0 \cdot 15 \times 10^{-3}$.

$$\therefore \quad \Phi_T = 1 \cdot 25 \times 0 \cdot 15 \times 10^{-3} \text{ Wb.}$$

The parallel path to the right of XY consists of an iron path of length $0 \cdot 4$ m (neglecting the air gaps) and cross-sectional area 10^{-4} m², in series with an air gap of the same cross-sectional area, 2 mm long.

Now, $S_{iron} = \dfrac{l}{\mu_0 \mu_r a} = \dfrac{0 \cdot 4}{4\pi \times 10^{-7} \times 600 \times 10^{-4}}$ A/Wb,

and $F_{iron} = \dfrac{\Phi_T}{2} \times S_{iron} = \dfrac{1 \cdot 25 \times 0 \cdot 15 \times 10^{-3} \times 0 \cdot 4}{2 \times 4\pi \times 10^{-7} \times 600 \times 10^{-4}}$,

$$= 497 \text{ A.}$$

Also, $S_{air} = \dfrac{l}{\mu_0 a} = \dfrac{2 \times 10^{-3}}{4\pi \times 10^{-7} \times 10^{-4}}$ A/Wb,

and $F_{air} = \dfrac{\Phi_T}{2} \times S_{air} = \dfrac{1 \cdot 25 \times 0 \cdot 15 \times 10^{-3} \times 2 \times 10^{-3}}{2 \times 4\pi \times 10^{-7} \times 10^{-4}}$,

$$= 1492 \text{ A.}$$

Hence, total m.m.f. = 1989 A.

Q. 8. *(a) Give the meanings of the following terms:*
(i) +20 dB with respect to 1 W,
(ii) −10 dBW,

(iii) +3 dBm, and
(iv) +12 dB with respect to 1 mV.

(b) The voltage gain of an amplifier is 26 when terminated by an output resistance R ohms. This resistance is replaced by a 10 dB attenuator presenting the same input resistance, R ohms, and its output is terminated by another resistance of R ohms. Find

(i) the input voltage to the amplifier to give 50 mV across the output resistance of the attenuator, and
(ii) the power delivered by the amplifier into the attenuator when R is 600 Ω.

A. 8. *(a)* The ratio of 2 powers, P_1 and P_2, expressed in decibels, is given by $10 \log_{10}(P_1/P_2)$ decibels, and between 2 voltages, V_1 and V_2, by $20 \log_{10}(V_1/V_2)$ decibels. In transmission systems, it is often useful to refer to signal levels, in terms of power or voltage, relative to some prescribed power or voltage.

(i) For a power level of $+20$ dB with respect to 1 W, $P_2 = 1$ W in the above equation.

$$\therefore \quad 20 = 10 \log_{10}(P_1/1).$$

$$\therefore \quad 2 = \log_{10} P_1.$$

$$\therefore \quad P_1 = 100 \text{ W.}$$

Hence, $+20$ dB with respect to 1 W is a power level of 100 W.

(ii) A power level of -10 dBW indicates a level of -10 dB with respect to 1 W.

$$\therefore \quad -10 = 10 \log_{10}(P_1/1).$$

$$\therefore \quad 1 = -\log_{10} P_1 = \log_{10}(1/P_1).$$

$$\therefore \quad 1/P_1 = 10.$$

$$\therefore \quad P_1 = 0 \cdot 1 \text{ W.}$$

Hence, -10 dBW is a power level of $0 \cdot 1$ W.

(iii) A power level of $+3$ dBm indicates a level of $+3$ dB with respect to 1 mW.

$$\therefore \quad 3 = 10 \log_{10}(P_1/1), \text{ where } P_1 \text{ is in milliwatts.}$$

$$\therefore \quad 0 \cdot 3 = \log_{10} P_1.$$

$$\therefore \quad P_1 = 2 \text{ mW.}$$

Hence, $+3$ dBm is a power level of 2 mW.

*(iv)* For a voltage level of $+12$ dB with respect to 1 mV,

$$12 = 20 \log_{10}(V_1/1), \text{ where } V_1 \text{ is in millivolts.}$$

$$\therefore \quad 0 \cdot 6 = \log_{10} V_1.$$

$$\therefore \quad V_1 = 4 \text{ mV.}$$

Hence, $+12$ dB with respect to 1 mV is a voltage level of 4 mV.

(b) The sketch shows the amplifier and attenuator configuration, in which $V_3 = 50$ mV and $R = 600$ Ω.
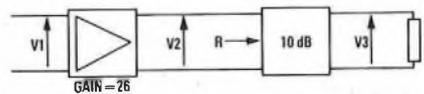
(i) Now, $10 = 20 \log_{10}(V_2/V_3)$.

$$\therefore \quad 0 \cdot 5 = \log_{10}(V_2/50 \times 10^{-3}).$$

$$\therefore \quad V_2 = 3 \cdot 162 \times 50 \times 10^{-3} \text{ V} = 158 \cdot 1 \text{ mV.}$$

Also, $V_2/V_1 = 26$.

$$\therefore \quad V_1 = (158 \cdot 1 \times 10^{-3})/26 \text{ V} = 6 \cdot 08 \text{ mV.}$$

(ii) The power delivered to the attenuator

$$= V_2^2/R,$$

$$= (158 \cdot 1 \times 10^{-3})^2/600 \text{ W} = 41 \cdot 7 \text{ } \mu\text{W.}$$

Q. 9. *(a) What is meant by the impedance ratio of a transformer?*
(b) Describe experiments on a transformer to establish how the ratio of the turns of the windings is related to the ratio of

(*i*) *the primary and secondary voltages, and*
(*ii*) *the primary and secondary currents.*

Set out the answer in the form of a laboratory report. Ignore losses in the transformer.

A. 9. (*a*) The impedance ratio of an ideal transformer is the ratio of the resistance of the primary circuit, R_P, to that of the secondary circuit, R_S, and is equal to the square of the turns ratio, $(N_P/N_S)^2$. Thus, a resistance in the secondary circuit can be referred to the primary circuit in order to simplify calculations. For example, a transformer with a turns ratio of 1 : 2, and having a primary circuit resistance of 50 Ω and a secondary circuit resistance of 400 Ω, can be regarded as having a total primary circuit resistance of $50 + 400 \times (1/2)^2$ Ω.

(*b*) (*i*) and (*ii*) The sketch shows the circuit diagram for an experiment to establish how the turns ratio of a transformer ($N_1 : N_2$) is related to the ratio of primary to secondary voltage ($V_1 : V_2$), and the ratio of the primary to secondary current ($I_1 : I_2$).
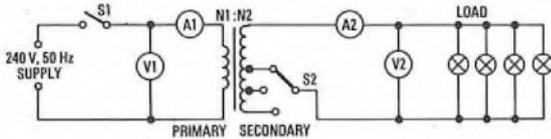
Apparatus

The transformer has a number of tapping points on its secondary winding, selected by switch S2, the number of turns at each tapping point being known. Voltmeters V1 and V2 record the primary and secondary voltages respectively, and ammeters A1 and A2, the primary and secondary currents respectively. Switch S1 isolates the mains supply, and the secondary load consists of a number of lamps in parallel.

Procedure

For each tapping point, readings of the primary and secondary voltages and currents are recorded against the turns ratio operative for that tapping.

Results

Graphs of $(V_1 : V_2)/(N_1 : N_2)$ and $(I_1 : I_2)/(N_2 : N_1)$ are plotted. Both graphs are linear, pass through the origin, and have unity slope. Hence,

$$\frac{V_1}{V_2} = \frac{N_1}{N_2} \text{ and } \frac{I_1}{I_2} = \frac{N_2}{N_1}.$$

Q. **10.** (*a*) *Explain the significance of hysteresis in a magnetic material. State what the area of the hysteresis loop represents and give an expression from which calculations can be made.*
(*b*) *Sketch typical hysteresis loops for*

(*i*) *soft iron, and*
(*ii*) *permanent-magnet steel.*

(*c*) *With reference to your sketches for part* (*b*) *above, explain the meanings of*

(*i*) *coercivity,*
(*ii*) *retentivity, and*
(*iii*) *saturation.*

A. 10. (*a*) Consider a magnetizing force, H amperes/metre, applied to an initially demagnetized ferromagnetic material, and varied by equal positive and negative amounts about zero. The resulting variation in flux density, B teslas, is shown in sketch (*a*). Initially, as H rises from zero to its positive value, B rises to its saturation value according to the curve OA. As H returns to zero, B follows the curve AB, so that when $H = 0$, B has some positive value due to the material retaining some residual magnetism, known as remanence, the value of the remanence being given by OB. As H falls to its negative value, B follows the curve BCD, and when H returns to zero, B has a residual value given by OE. Finally, as H returns to its positive value, B follows the curve EFA. For subsequent cyclic variations in H, B follows the curve ABCDEFA. Hence, once the material has been initially magnetized, the flux density at all times lags the magnetizing force, this property being known as magnetic hysteresis.

The curve ABCDEFA is known as a hysteresis loop, or B/H loop, and its area represents the energy expended/unit volume of material in taking the material through one magnetization cycle.

Hysteresis loss $\propto fB_m^{1\cdot 6}$, where f is the frequency (Hz), and B_m is the maximum value of flux density (T).

(*a*)   (*b*)   (*c*)
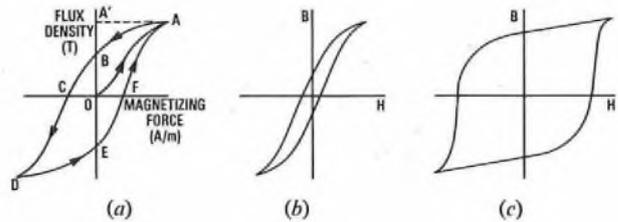
(*b*) (*i*) and (*ii*) Sketches (*b*) and (*c*) show typical B/H loops for soft iron and permanent-magnet steel respectively.
(*c*) (*i*) Coercivity is the value of demagnetizing force required to reduce the magnetic flux density from its remanence value to zero, and is represented by OC in sketch (*a*). From sketches (*b*) and (*c*), it can be seen that soft iron has a low value of coercivity, while permanent-magnet steel has a high value.
(*ii*) Retentivity is a deprecated term, synonymous with remanence, and is the value of flux density remaining after an initial magnetization to saturation, and is the value of flux density remaining after an initial magnetization to saturation. It is represented by OB in sketch (*a*), and from sketches (*b*) and (*c*), it can be seen that soft iron has a low value of remanence, while permanent-magnet steel has a high value.
(*iii*) As H increases, B reaches a maximum value, shown by OA' in sketch (*a*). Further increases in H result in no increase in B, and the material is said to be in saturation.

RADIO AND LINE TRANSMISSION B, 1973

Students were expected to answer any 6 questions

Q. **1.** (*a*) *Describe the construction of an aerial array, consisting of a horizontal dipole, reflector and director, which is suitable for use at signal frequencies of about 90 MHz.*
(*b*) *State, in metres, suitable dimensions for the elements used.*
(*c*) *Illustrate the actions of the reflector and director by sketching the horizontal pattern of*

(*i*) *the dipole alone,*
(*ii*) *the dipole with reflector, and*
(*iii*) *the complete array.*

A. 1. (*a*) A horizontal aerial array, comprising a dipole, reflector and director, is shown in sketch (*a*). The elements are made from a good metal conductor; for example, aluminium tubing of about 10 mm diameter. They are supported at their centres by a horizontal beam, which is usually, though not necessarily, a conductor, with the dipole insulated from the supporting beam. The feeder is connected to the 2 halves of the dipole, its impedance being about 75 Ω, so that a coaxial-cable feeder may be used. A supporting mast is also shown, to which the coaxial feeder may conveniently be clamped.

Typical dimensions for the elements are also shown in sketch (*a*) in terms of the wavelength, λ, from which it can be seen that the lengths of the elements decrease towards the direction of transmission

(*a*)
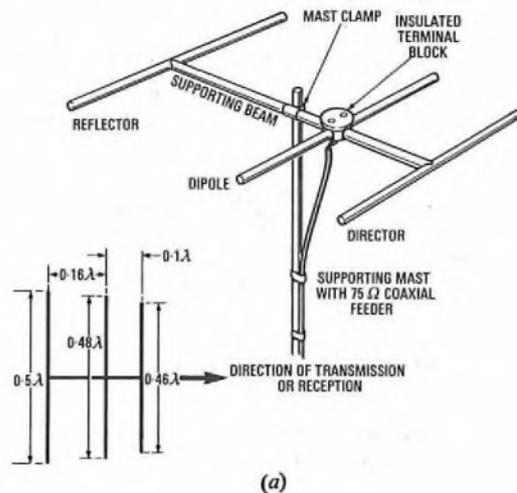
or reception. The dipole is made slightly shorter than half a wavelength to allow for end-effects.

(*b*) The wavelength in metres is given by $\lambda = c/f$, where c is the velocity of light (m/s), and f is the frequency (Hz).

Thus,

$$\lambda = \frac{300 \times 10^6}{90 \times 10^6} = 3 \cdot 33 \text{ m.}$$

Therefore, the dimensions of the elements are as follows.

$$\text{Director: } 0 \cdot 46\lambda = 1 \cdot 53 \text{ m.}$$
$$\text{Dipole: } 0 \cdot 48\lambda = 1 \cdot 60 \text{ m.}$$
$$\text{Reflector: } 0 \cdot 50\lambda = 1 \cdot 67 \text{ m.}$$
$$\text{Director–dipole spacing: } 0 \cdot 1\lambda = 0 \cdot 33 \text{ m.}$$
$$\text{Dipole–reflector spacing: } 0 \cdot 16\lambda = 0 \cdot 53 \text{ m.}$$

(*c*) Sketch (*b*) shows the horizontal radiation pattern of the dipole alone. This is a figure-of-8 response, having maxima at right angles to the dipole, and a minimum end-on to the dipole.

(*b*)

(*c*)

(*d*)
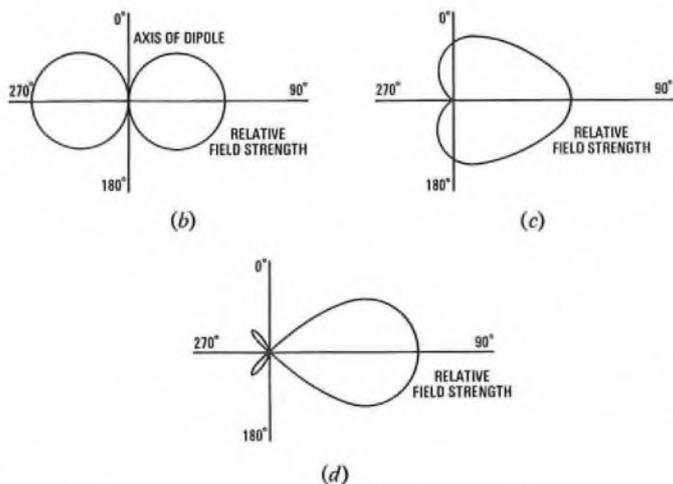
The addition of a reflector modifies the response to that shown in sketch (*c*). The radiation pattern has only one maximum, at right angles to the dipole and away from the reflector. This pattern is known as a cardioid response.

The directional properties of the aerial are further enhanced by the addition of a director. Sketch (*d*) shows the improved forward response achieved from the array, which now has greater gain and a reduced beamwidth. However, the response to the rear of the aerial develops as indicated by the sidelobes in the pattern.

Q. **2.** (*a*) *Define the following terms used in superheterodyne reception:*

(*i*) *image-channel or second-channel rejection,*
(*ii*) *adjacent-channel selectivity, and*
(*iii*) *frequency changer.*

(*b*) *Explain where, in the superheterodyne receiver, image-channel or second-channel rejection and adjacent-channel selectivity are most effectively obtained.*

(*c*) *Why is a frequency-changing process used?*

A. 2. (*a*) (*i*) The image, or second, channel is any frequency falling in a band which is equal to the intermediate-frequency bandwidth and centred on a frequency equal to the local-oscillator frequency plus the intermediate frequency. The image frequency is also equal to the required signal frequency plus twice the intermediate frequency. Image-channel, or second-channel, rejection is the ability of a receiver to reject these frequencies.

(*ii*) Adjacent-channel selectivity is the ability of a receiver to discriminate between the wanted signal and a signal very close to it in the frequency spectrum.

(*iii*) A frequency changer is the stage in a superheterodyne receiver, following the radio-frequency stage, where all incoming signals are converted to a common frequency—the intermediate frequency—at which the gain and selectivity of the receiver are achieved. The frequency changer mixes the incoming signal with a locally-generated oscillator signal in a non-linear element, such as a suitably biased valve or transistor, and the desired frequency is selected by a tuned circuit in the anode or collector.

(*b*) Image-channel, or second-channel, rejection is achieved by the use of a radio-frequency stage with sufficiently selective tuned circuits. It is aided by the choice of a high intermediate frequency, since, as the image channel is spaced from the wanted signal by twice the intermediate frequency, then the higher the intermediate frequency, the

further down the response curve of the radio-frequency stage the image frequency falls.

Adjacent-channel selectivity is achieved in the intermediate-frequency amplifier, by employing a number of tuned circuits with several stages of gain. The choice of a low intermediate frequency assists selectivity, since, for a given Q-factor, the lower the frequency, the lower the bandwidth of the tuned circuit.

(*c*) A frequency-changing process is used so that most of the amplification of the receiver, and all the channel selectivity, may be obtained at a single, common frequency. This simplifies the design of the receiver, by reducing the number of tuned circuits that need to be ganged together, and provides a stable receiver, since the gain is achieved at a frequency lower than that of the wanted signal.

Q. **3.** (*a*) *Sketch, on squared paper, a train of rectangular pulses, given the following data:*

$$\text{pulse amplitude} = 1 \, V,$$
$$\text{pulse width} = 100 \, \mu s,$$
$$\text{pulse-repetition frequency} = 1 \, kHz.$$

(*b*) *Sketch the amplitude/frequency spectrum of the above waveform.*

(*c*) *What bandwidth is required to transmit the pulse train with low distortion?*

A. 3. (*a*) The pulse train is shown in sketch (*a*). As the pulse-repetition frequency is 1 kHz, the period between the leading edges of adjacent pulses is $1/1000 \text{ s} = 1 \text{ ms}$.

(*a*)

(*b*)
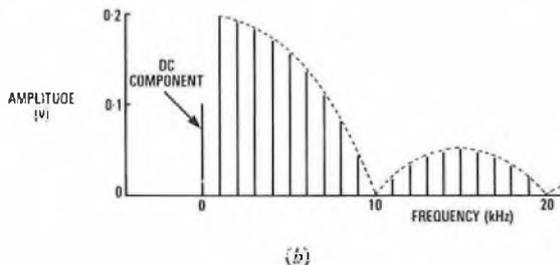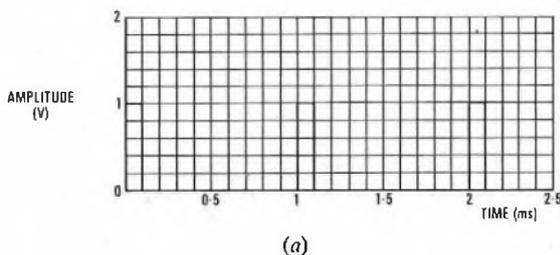
(*b*) Sketch (*b*) shows the amplitude/frequency spectrum of the waveform. The envelope has a shape according to the law (sin x)/x, falling to zero at a frequency of $1/(\text{pulse width}) = 1/(100 \times 10^{-6}) \text{ Hz} = 10 \text{ kHz}$, and comprises a line spectrum spaced at intervals equal to the pulse-repetition frequency.

(*c*) To transmit the pulse train with low distortion, a bandwidth of at least 10 times the reciprocal of the pulse width is required, in order to contain the harmonic frequencies that constitute the waveform. In this case, a bandwidth extending from d.c. to 100 kHz is required.

Since most of the energy of the pulse is contained in a bandwidth equal to the reciprocal of the pulse width, 10 kHz in this case, some systems provide only this, in order to economize on bandwidth.

Q. **4.** (*a*) *Sketch the circuit diagram of a 2-stage, high-frequency, transistor amplifier.*

(*b*) *With reference to the amplifier you have sketched, explain the reasons for using*

(*i*) *any special biasing arrangements, and*
(*ii*) *the transistor configurations shown.*

(*c*) *Where and why would decoupling capacitors be used in such a circuit?*

Q. **5.** (*a*) *What value of inductance resonates with a capacitance of* 100 pF *at a frequency of* 465 kHz?

(*b*) *How could the value of this inductance be checked on a Q-meter?*

(*c*) *Explain the meaning of the Q-value so obtained.*

(*d*) *Explain what is meant by*

(*i*) *the effective series resistance of an inductor, and*
(*ii*) *the dynamic impedance of a parallel resonant circuit.*

A. 5. *(a)* At resonance, the inductive reactance is equal, but opposite in sign, to the capacitive reactance.

Thus, $\omega L = \dfrac{1}{\omega C}$, where L is the inductance (H),
C is the capacitance (F), and
ω is the angular frequency (rad/s), equal to $2\pi \times$ frequency (Hz).

$$\therefore L = \frac{1}{\omega^2 C},$$

$$= \frac{1}{4\pi^2 \times 465^2 \times 10^6 \times 100 \times 10^{-12}} \text{ H},$$

$$= 1 \cdot 17 \text{ mH}.$$

(b) The inductance value is checked on a Q-meter by connecting the component across the INDUCTANCE terminals and setting the frequency of the measuring instrument to 465 kHz. If the inductance value is correct, resonance is obtained, indicated by a maximum reading of the meter, when the variable capacitor is set to 100 pF.

(c) The Q-meter provides information about inductance and capacitance in series resonance. The Q-factor obtained from the instrument is the voltage-magnification factor, and is the ratio of the voltage, V_x, measured across one of the reactances, in this case the capacitor, to the voltage, V_s, applied in series with the circuit.

At resonance, $\omega L = 1/\omega C$, and the current in the circuit, I, $= V_s/r$ amperes, where r is the effective series resistance of the inductor (Ω).

The voltage developed across the capacitor is given by $V_x = I/\omega C = V_s/\omega Cr$.

Since $Q = V_x/V_s$, $Q = 1/\omega Cr = \omega L/r$.

The Q-factor is, therefore, the ratio of capacitive or inductive reactance to the resistance in the circuit.

(d) *(i)* The effective series resistance of an inductor is the equivalent resistance in series with the inductor which, at series resonance, gives a circuit current of $I = v/r$ amperes, where v is the radio-frequency voltage applied to the circuit.

The value of the effective series resistance depends on the d.c. resistance of the wire used in the construction of the inductor and the a.c. skin effect, which is a function of frequency.

(ii) The dynamic impedance, R_D ohms, of a parallel resonant circuit is the resistance obtained across the circuit at resonance, and is given by

$$R_D = Q\omega L = \frac{Q}{\omega C} = \frac{L}{Cr} \text{ ohms.}$$

Q. 6. *(a) With reference to long-distance high-frequency radio transmission, explain, in terms of frequency and mode of propagation, the difference between ground waves and sky waves.*

(b) With the aid of sketches, explain the terms

(i) maximum usable frequency,
(ii) skip distance, and
(iii) refraction.

A. 6. *(a)* The essential difference between a ground wave and a sky wave is that the former is a surface wave, whereas a sky wave depends on the reflecting properties of the ionosphere to reflect the signal around the Earth.

A ground wave must be launched as a vertically-polarized wave. Some of the energy transmitted penetrates the Earth's surface and, thus, the attenuation of the signal depends on the conductivity of the Earth and the surface roughness, so that propagation is more effective at low frequencies and over the sea. Because the transmission is a surface wave, it tends to follow the curvature of the Earth, the limit of transmission being determined by the quality of the surface medium and the frequency of operation. Once communication has been established by the use of ground waves, a permanent link is obtained which is not subject to fluctuation.

When a wave is launched from an aerial, it is possible for both sky waves and ground waves to be generated. At low frequencies, ground waves are predominant, since the majority of the energy is converted into a surface wave. At high frequencies, ground waves are greatly attenuated, but it is possible to transmit signals at an angle above the horizontal and obtain reflection from the ionosphere, thus permitting reception to be obtained at distances considerably beyond the horizon.

The ionosphere consists of layers of ionized gases at heights of 48–400 km above the Earth's surface. These layers cause the incident radio waves to change direction by a process of refraction. If the density of the layers is sufficient, the direction of the waves can be changed so that they return to Earth and, thus, the ionized layers act like a mirror, reflecting the incident waves back to Earth some distance beyond the horizon. The f_1 layer, at 145–241 km, and the f_2 layer, at 322 km above the Earth, have most effect on high-frequency radio propagation. Signals may also reach a receiver after travelling by a series of multiple reflections between the ionosphere and the Earth.

The disadvantage of sky-wave communication between 2 points is that the transmission path is not very stable. The amount of refraction depends on the ionization at any one time and, since this is caused by ultra-violet radiation from the sun, it varies with the time of day and the degree of sunspot activity.

(b) *(i)* The maximum usable frequency (MUF) is the highest frequency at which communication may be established over a given path between 2 fixed points. If a frequency greater than the MUF is transmitted, refraction in the ionsphere is reduced, and the consequent skip distance is increased, thereby placing the receiver in the dead zone. The MUF depends on the time of day and the season, since these affect the state of the ionosphere.

(ii) The skip distance is the distance from the transmitter to the point where the returned sky-wave signal is first received, as illustrated in sketch *(a)*.

(a)
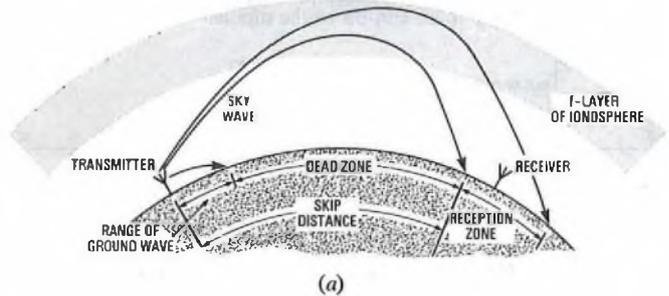
(iii) When a radio wave enters an ionized region of gas from a non-ionized region, or from an ionized region of lesser density, it experiences an increase in phase velocity. As a result, a radio wave passing into the ionosphere, at some angle to the vertical, bends further away from the vertical as it moves into a region of higher ionization. If the rate of change of ionization with height is sufficient, the wave is refracted back towards the Earth.


(b)
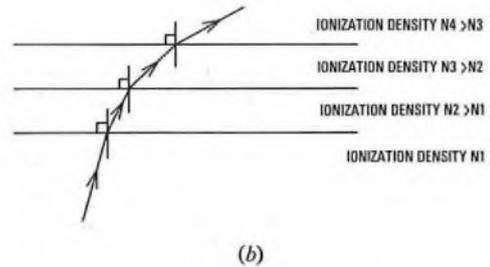
Sketch *(b)* illustrates how a radio wave, incident on successively more-densely ionized layers of the ionosphere, is refracted.

Q. 7. *(a) Describe, with the aid of circuit diagrams, a method of measurement of (i) capacitance, and (ii) inductance, at frequencies of approximately 1 MHz.*

(b) A test circuit, consisting of an inductor in parallel with a calibrated capacitor, resonates at 1 MHz, when the scale reads 200 pF. When an inductor is placed across the circuit, the capacitor value has to be changed to 250 pF to maintain resonance. Calculate the value of the inductor.

A. 7. *(a)* A Q-meter may be used to measure both capacitance and inductance at frequencies of about 1 MHz, and a simplified circuit diagram of such a meter is shown in the sketch.
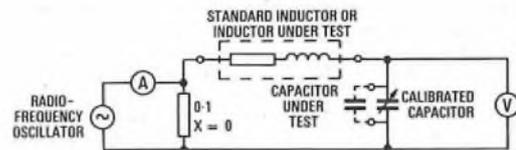
The Q-meter comprises a radio-frequency oscillator, which may be tuned over the frequency range of interest, coupled to a low-value non-inductive resistor of known resistance via a radio-frequency ammeter. This allows the amount of current delivered to the resistor to be set in conjunction with a control on the oscillator. Connected in series with the resistor is the inductor under test and a high-grade calibrated capacitor. A high-impedance radio-frequency voltmeter is connected across the capacitor. At series resonance, the voltage across the capacitor is a maximum, as indicated by the voltmeter, and the voltage magnification,

or *Q*-factor, is given by the ratio of this voltage to the known voltage across the non-inductive resistor.

(*i*) To measure an unknown capacitance at a frequency of 1 MHz, the oscillator frequency is set to this value, and a standard inductor is chosen that resonates at this frequency with the calibrated variable capacitor, when the latter is set near its maximum value, typically 500 pF. Resonance is indicated by a maximum reading on the high-impedance voltmeter. Let the value of the calibrated variable capacitor be C_1 farads at this point. The unknown capacitance is then connected in parallel with the variable capacitor, and the latter is retuned for resonance at the same frequency. If the new value of the variable capacitor is C_2 farads, the unknown capacitance is given by the difference, $C_1 - C_2$ farads.

(*ii*) To measure an unknown inductance, the unknown inductor is connected into the circuit, and resonance is established by tuning the variable capacitor until a maximum reading is observed on the volt-meter. Let the value of the capacitor at this point be *C* farads. Since, at resonance, the values of capacitive and inductive reactance are equal, the value of the inductance is calculated from the formula $\omega L = 1/\omega C$, where *L* is the inductance (H), and $\omega = 2\pi f$, where *f* is the resonant frequency. Hence, $L = 1/\omega^2 C$ henrys.

(*b*) For the given inductor, L_1 henrys, in parallel with the calibrated capacitor, $C = 200$ pF, resonating at a frequency $f = 1$ MHz,

$$\omega L_1 = \frac{1}{\omega C}.$$

$$\therefore \ L_1 = \frac{1}{\omega^2 C},$$

$$= \frac{1}{(2\pi \times 10^6)^2 \times 200 \times 10^{-12}} \text{H},$$

$$= 126 \cdot 7 \ \mu\text{H}.$$

When the second inductor, L_2 henrys, is connected across the circuit, the effective inductance, *L* henrys, is given by

$$L = \frac{1}{(2\pi \times 10^6)^2 \times 250 \times 10^{-12}} \text{H},$$

$$= 101 \cdot 3 \ \mu\text{H}.$$

L represents the total inductance of inductors L1 and L2 in parallel. Therefore,

$$\frac{1}{L} = \frac{1}{L_1} + \frac{1}{L_2}.$$

$$\therefore \ \frac{1}{L_2} = \frac{1}{L} - \frac{1}{L_1},$$

$$= \frac{1}{101 \cdot 3 \times 10^{-6}} - \frac{1}{126 \cdot 7 \times 10^{-6}},$$

$$= 1979.$$

$$\therefore \ \underline{L_2 = 505 \cdot 3 \ \mu\text{H}.}$$

Q. **8.** (*a*) *Draw and describe a fixed-frequency oscillator circuit which provides good frequency stability at a high frequency.*

(*b*) *Give reasons for the choice of circuit.*

(*c*) *Describe any extra steps which may be taken to enhance the frequency stability.*

A. **8.** (*a*) A crystal-controlled oscillator is generally used for good frequency stability, provided that the frequency required is one of constant value. At high frequencies, the circuit shown in sketch (*a*) may be used. The oscillator is, in principle, a tuned-anode/tuned-grid circuit, the tuned grid being provided by the crystal.

The equivalent circuit of the quartz crystal is shown in sketch (*b*). The natural resonance of the crystal is represented by the series circuit formed by resistor R, inductor L and capacitor C. In addition, the terminal capacitance, C_T, provides a parallel-resonance mode. The reactance/frequency diagram of the crystal is, therefore, of the form shown in sketch (*c*).

In the circuit shown in sketch (*a*), the crystal is in parallel resonance at the frequency of oscillation. A voltage in the grid circuit of the valve is amplified and developed across the resonant load in the anode circuit. Feedback to the crystal is obtained through a small capacitance, C_F, between the anode and grid, which may, in the limit, be the inter-electrode capacitance. For correct phasing conditions, the anode load is required to be inductive and, to meet these conditions, the circuit is tuned slightly below the frequency of oscillation. Because of the high *Q*-factor of the crystal, the frequency of oscillation is controlled by the crystal, which is cut for a specific value. However, at high frequencies, precautions are taken to moderate the amplitude of oscillation, which may otherwise damage the crystal.

(*b*) A crystal-controlled oscillator is chosen because it requires parallel resonance in the crystal, and this is more readily obtained at

high frequencies up to about 30 MHz. If higher frequencies are required, the circuit is usually associated with a frequency multiplier.

(*c*) The frequency stability of a crystal-controlled oscillator is largely dependent on the properties of the crystal itself. Although crystals may be cut to be stable over wide ranges of temperature,

(a)

(b)   *(c)*
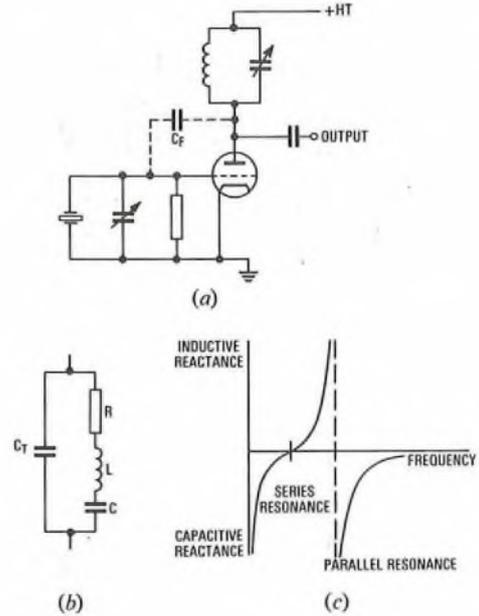
these usually operate at the lower frequencies. At high frequencies, it is possible to increase the frequency stability by enclosing the crystal in a temperature-controlled oven, using stabilized power supplies and a buffer amplifier to reduce the load on the oscillator circuit. A suitable buffer circuit is the cathode or emitter follower, which has a high input impedance and low output impedance.

Q. **9.** (*a*) *Fig.* 1 *shows the circuit diagram of a class-A audio-frequency amplifier. Redraw the circuit with additional components to provide overall voltage negative feedback over the 2 stages.*

(*b*) *The response curve of the amplifier is given in Fig.* 2. *Calculate the resultant values of gain at 100, 1000, 10 000 and 100 000 Hz, when 1% of the output voltage is fed back to the input in antiphase. Sketch the resultant response curve.*
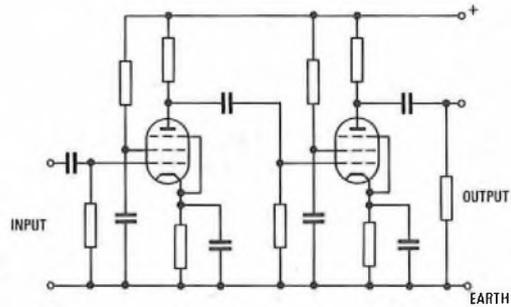

Fig. 1
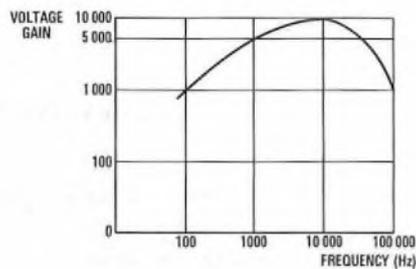
Fig. 2

A. **9.** (*a*) The additional components required to provide overall voltage negative feedback over the 2 stages are shown in sketch (*a*). To

ensure that the sense of the feedback is correct, a suitable fraction of the output from the anode of the second valve is taken to the cathode of the first.

(*a*)
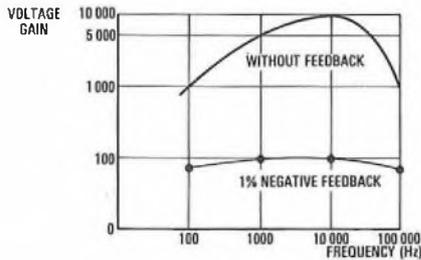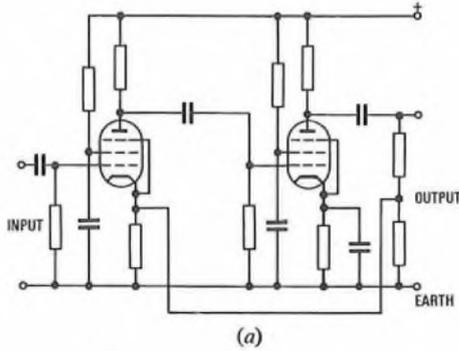
(*b*)

(*b*) When negative feedback is applied to an amplifier, the overall gain is reduced to $A/(1 + A\beta)$, where A is the voltage gain without feedback, and β is the feedback fraction.

For 1 % negative feedback, the resultant values of gain are shown in the table.

Frequency (Hz)	Gain Without Feedback	$\dfrac{A}{1 + A\beta}$ = Gain With Feedback
100	1 000	$\dfrac{1000}{1 + 10} = 90\cdot91$
1 000	5 000	$\dfrac{5000}{1 + 50} = 98\cdot04$
10 000	10 000	$\dfrac{10\ 000}{1 + 100} = 99\cdot01$
100 000	1 000	$\dfrac{1000}{1 + 10} = 90\cdot91$

For high values of gain, $A\beta \geqslant 1$ and, hence, the gain with feedback is given by $1/\beta$. Therefore, as a check, the values for gain with feedback given in the table should lie close to $1/0\cdot01 = 100$, and can be seen to do so.

The resultant response curve is shown in sketch (*b*).

Q. **10.** (*a*) *With the aid of a circuit diagram, describe the action of an emitter-follower circuit, making particular reference to input and output impedances.*

(*b*) *Give 2 uses for the emitter follower.*

(*c*) *Describe suitable biasing arrangements for an emitter follower, giving reasons for your choice of components.*

(*d*) *How are the input and output impedances of the emitter follower affected by the bias circuit chosen in part (c)?*

A. **10.** (*a*) The circuit diagram of an emitter follower is shown in the sketch. For an *n p n* transistor, the collector is connected to the positive supply rail, which is suitably decoupled and common to the input and output. The input is applied to the base, and the output is taken from the emitter resistor, R3.
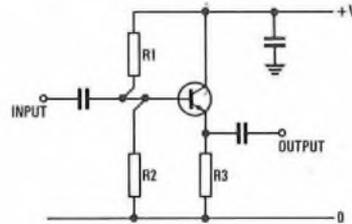
The emitter follower is an example of the extreme use of negative feedback, since the whole of the output, which appears across the emitter resistor, is fed back to the base circuit. Thus, the overall gain is given by $A/(1 + A)$, where A is the gain of the transistor, the overall gain being slightly less than unity.

The effect of the large amount of negative feedback on the input and output impedances is to increase the input impedance to a very high value, while reducing the output impedance to a low value. A typical value for the input impedance is 100 kΩ, or more, and for the output impedance, 50–1000 Ω.

(*b*) An emitter follower may be used as an impedance-matching device between a high and a low impedance; for example, for matching a high-impedance microphone to a low-impedance unbalanced coaxial cable. If the high-impedance microphone were connected directly to a screened cable of any great length, the capacitance of the cable would short circuit the high-frequency components. If it is first matched to a low impedance, the high frequencies are faithfully transmitted.

An emitter follower may also be used as a buffer amplifier for an oscillator. An oscillator should, ideally, be lightly loaded and work into a constant load impedance. The high input impedance of an emitter follower ensures that the load is light, and the low output impedance ensures that changes in the load presented to the emitter follower are not transferred to its input impedance, which, therefore, acts as a constant load for the oscillator.

(*c*) A suitable biasing arrangement is illustrated in the sketch. It comprises a potential divider, formed by resistors R1 and R2, which are of large, equal values, together with an emitter resistor, R3. The potential divider ensures that the base potential is approximately half the supply voltage. The transistor then conducts until the emitter current, I_E, is such that $I_E R_3$ plus the base–emitter potential is equal to half the supply voltage. Therefore, the value of the emitter voltage under no-signal conditions is approximately half the supply voltage, enabling large voltage swings to be handled without clipping occurring. The value of the emitter resistor is chosen to be low by setting the bias current to a suitable value, so as to keep the overall output impedance of the emitter follower low.

(*d*) The input impedance to the emitter follower is reduced by the biasing arrangement, since the bias resistors appear in parallel with the input. The output impedance is determined by the bias applied to the transistor, which determines the slope of the output characteristic. The emitter resistor is in parallel with this impedance.

ELEMENTARY TELECOMMUNICATION PRACTICE, 1974

Students were expected to answer any 6 questions

Q. **1.** (*a*) *Draw a labelled sketch of a moving-iron-diaphragm telephone receiver.*

(*b*) *List the materials of which the main components are constructed, and state the properties for which they are chosen.*

(*c*) *Give reasons for the inclusion of the permanent magnet.*

A. **1.** (*a*) The sketch shows a typical telephone receiver of the type in which a diaphragm of a magnetic material moves under the influence of a varying electromagnetic field, created by currents flowing in the speech coils.

(*b*) (*i*) The permanent magnet is made from an alloy of iron, aluminium, nickel and cobalt, and has a high permeability, high retentivity and high specific resistance. Therefore, it produces a strong and lasting magnetic field, and wastes a comparatively small amount of energy in useless eddy currents.

(*ii*) The pole faces are of nickel-iron, which has a high permeability, low retentivity and low hysteresis and eddy-current losses.

(*iii*) The speech coils are wound with copper wire to reduce resistive losses.

(*iv*) The diaphragm is made of an alloy of iron and cobalt, which

has a high permeability and low retentivity, so that it responds readily to the changing fields set up by currents in the speech coils. It is mechanically stiff, so that its thickness can be kept small and its weight low.

SILK CLOTH COVERING HOLE
SMALL HOLES
NICKEL-IRON POLE PIECES
SPEECH COILS
NICKEL-COBALT-ALUMINIUM-IRON MAGNET
COBALT-IRON (PERMENDUR) DIAPHRAGM
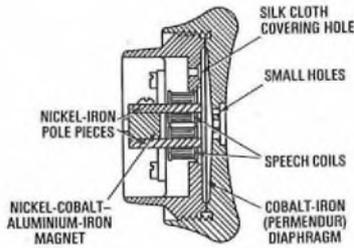
(c) The pull exerted by an electromagnet is proportional to the square of the flux density. If the flux density due to the speech coil is b, then, without a permanent magnet, the pull is proportional to b^2. If the current is reversed, so that the flux becomes $-b$, the pull is similarly proportional to b^2.

Thus, the effect of applying an alternating signal to such a device is to attract the diaphragm twice in each cycle, and this produces a sound wave of twice the frequency of the applied electrical signal. This is equivalent to raising the reproduced sound by one octave.

If a permanent magnet is included, producing a flux density of B, where $B > b$, the flux density in the magnetic circuit now varies between $(B + b)$ and $(B - b)$. Hence, the pull on the diaphragm varies between values proportional to $(B + b)^2 = (B^2 + 2Bb + b^2)$ and $(B - b)^2 = (B^2 - 2Bb + b^2)$, giving a variation in pull between values proportional to $+2Bb$ and $-2Bb$. Therefore,

(i) the frequency-doubling phenomenon is avoided, and

(ii) the sensitivity changes from a value proportional to b^2 to a value proportional to $2Bb$. Since B can readily be made much larger than b, there is a considerable improvement in sensitivity.

Q. 2. (a) Describe, with the help of a sketch, a jack of the type used in a telephone switchboard.

(b) Explain what happens when a suitable plug is inserted in the jack.

(c) Sketch the type of contact used in the jack, and explain why it is shaped as it is.

A. 2. See A.2, Elementary Telecommunication Practice, 1971, *Supplement*, Vol. 64, p. 82, Jan. 1972.

Q. 3. (a) Describe, preferably with the aid of a sketch, the Murray (5-unit) code.

(b) Explain why it is suitable for teleprinter transmission.

(c) Explain how it is arranged that more characters can be transmitted than the number of code combinations available.

A. 3. (a) Each character in the 5-unit code consists of a discrete combination of 5 elements, each of the same duration (20 ms), derived from 2 basic signal conditions. These are the *mark* and *space* signals, corresponding to negative and positive potentials respectively for double-current signalling.

As all characters are of the same duration, no time interval is required between the elements, but it is necessary to distinguish between characters. The introduction of a *start* signal before each character, and a *stop* signal after it, gives the necessary separation and, at the same time, provides signals for the synchronization of the transmitter and receiver for each character. The complete signal for one character is shown in the sketch, assuming double-current working.

START SIGNAL
+80 V
SPACE 0
MARK
−80 V
0 20 40 60 80 100 120 140
150
TIME (ms)
STOP SIGNAL
5 ELEMENTS
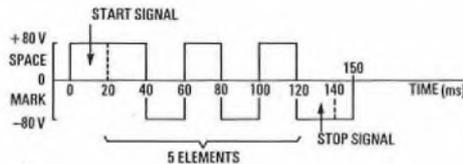
(b) The constant transmission time of each character makes the code suitable for transmitting between teleprinters.

(c) With a 5-unit code using 2 elements, the number of code combinations possible is $2^5 = 32$. This is an insufficient number to permit the allocation of a unique code to each of 26 letters, 10

numerals and the various punctuation marks which are required. Therefore, 2 meanings are allocated to most codes, and 2 codes are reserved for shift signals. After the receipt of the *letter-shift* signal, all subsequent codes are recognized as letters. After receipt of the *figure-shift* signal, the same codes are recognized as having their alternative meanings of figures and punctuation marks.

Q. 4. (a) Explain, quoting a code with which you are familiar, how individual wires of an internal telephone cable can be identified.

(b) Describe how the cable is terminated using the wire-wrapping method.

(c) Briefly state the precautions which must be observed in doing the work.

A. 4. (a) The wires of an internal cable are coded by combinations of 2 colour sequences. These are

(i) white, red, black and yellow, and

(ii) blue, orange, green, brown and grey.

The first 5 pairs are identified by the A-wires taking the first colour, white, of the first sequence, while the B-wires follow the second (blue-orange-green-brown-grey) sequence. Hence, pair 1 is white/blue, pair 2 is white/orange, pair 3 is white/green, pair 4 is white/brown, and pair 5 is white/grey.

Pairs 6–10 are identified by the A-wire taking the second colour, red, of the first sequence, while the B-wires again follow the second sequence. Similarly, pairs 11–15 have black A-wires, and pairs 16–20 have yellow A-wires.

For pairs 21–40, this structure is repeated, with the addition of a blue marking superimposed on the A-wire colour. Pairs 41–60 have an orange marking superimposed on the A-wire colour, and the pattern is continued to pair 120.

In cables made up of triples, the C-wire of each triple is coloured turquoise, and in those made up of quads, the D-wire is coloured violet.

(b) Wire-wrapping is a technique which is superseding soldered joints for making permanent connexions. Wire-wrapped joints have a more uniform resistance, and their resistance remains constant with time.

Wire-wrapping can be carried out only on tag blocks designed for the purpose; these have tags of rectangular cross-section. The wire is wrapped round the tag about 7 times, with enough tension to cause the corners of the tag to bite into the wire. To terminate a cable, the wires are fanned out in the usual way. A wrapping tool, which may be either manually operated or motor driven, is placed over the tag, and the appropriate wire is laid in guides on the tool. In operation, the tool crops the wire to the correct length, strips the insulation, and wraps the wire round the tag at the correct tension.

(c) The important features of a correctly made joint are that

(i) the wire is not scored or nicked during the stripping process,

(ii) the correct number of turns are made to ensure there are sufficient points of contact between the wire and the tag,

(iii) the tension is correct, so that the wire and tag are in close contact without damage to the wire, and

(iv) the turns of wire do not overlap.

The presence of these features is ensured by regularly making test wraps, checking them visually and measuring the pull required to remove the wrapping.

*Q. 5. (a) Explain why the specific gravity of a lead-acid cell varies during its discharge.*

(b) A cell, having a capacity of 100 A h, is discharged at a steady rate of 10 A. Sketch labelled graphs showing

(i) specific gravity/time, and

*(ii) potential difference across the cell/time.*

(c) What is the likely effect of persistently over-charging a cell?

A. 5. (a) The electrolyte contained in a lead-acid cell is a mixture of sulphuric acid and water. Because sulphuric acid has a specific gravity higher than that of water, the specific gravity of the mixture is greater than unity, the actual value depending on the proportions of the 2 constituents. In a typical case, the specific gravity of the electrolyte of a fully charged cell is 1·215. The electrochemical processes, which take place during discharge, result in the conversion of sulphuric acid to water, and this change in the relative amounts of the 2 constituents produces a fall in the specific gravity of the mixture. Because the amount of acid converted is directly related to the energy released by the cell, the specific gravity is a direct measure of the state of charge of the cell.

13

(*b*) Sketches (*a*) and (*b*) show the graphs of specific gravity/time and potential difference across the cell/time, respectively.
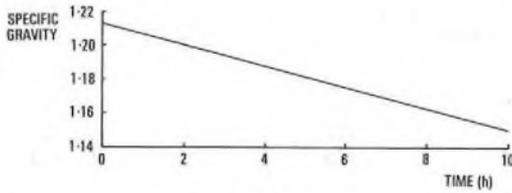

(*a*)
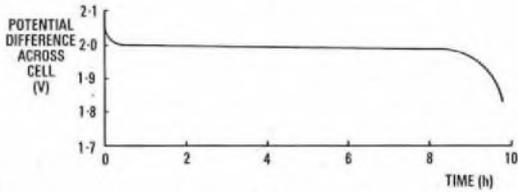
(*b*)

(*c*) When a charging current continues to flow after a cell is fully charged, the water in the electrolyte decomposes, giving off hydrogen and oxygen. This evolution of gases has the effect of loosening particles of active material from the plates, with any or all of the following consequences.

(*i*) The active material available is reduced, so reducing the capacity of the cell.

(*ii*) Sediment may accumulate at the bottom of the cell until the plates are short-circuited.

(*iii*) Flakes of active material may fall across the gap between plates, thus causing short-circuits.

(*iv*) The mechanical strains set up in the plates by the loss of material, and by overheating, may give rise to buckling of the plates.

Q. 6. (*a*) *With the help of a sketch, describe an alarm-type fuse suitable for protecting equipment in a telephone exchange.*

(*b*) *Distinguish between the rating and fusing currents of a fuse.*

(*c*) *Explain why it is essential for a fuse to have electrical resistance.*

A. 6. (*a*) Sketch (*a*) shows an alarm-type fuse of the sort currently used in telephone exchanges. It consists, essentially, of 2 flat springs mounted in a plastic body. The springs are tensioned apart, but are held towards each other by a link of fusible wire. Current is fed to the exchange equipment via the lower spring, the fuse wire and the upper spring. If the current drawn by the equipment rises above its proper value, the fuse wire melts, disconnecting the circuit and allowing the flat springs to fly apart, as shown in sketch (*b*). The lower spring makes contact with a metallic bar which is connected to an alarm circuit, so providing audible or visible alarm conditions, as required, and the upper spring enables the blown fuse to be visually located.
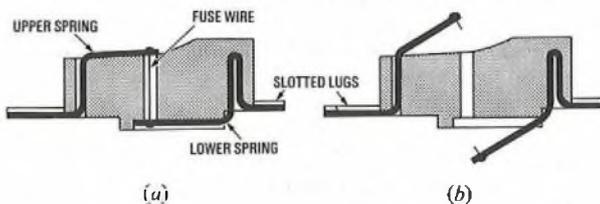
(*a*)   (*b*)

(*b*) The rating current of a fuse is the maximum current which the fuse will carry indefinitely without blowing. The fusing current is the minimum current which causes the fuse to blow almost instantaneously. The rating current is usually about half the fusing current.

(*c*) For its action, a fuse relies on the generation of heat in the fuse wire, to raise the temperature of the wire to the point where it melts, or yields under tension. The heat produced in a wire is proportional to I^2Rt, where I is the current flowing, R is the resistance of the wire, and t is the time for which the current flows. If R were zero, then I^2Rt would be zero, and no heat would be produced in the wire.

Q. 7. (*a*) *Sketch a printed-circuit board, and describe how components are connected to it.*

(*b*) *From both economic and operational viewpoints, list the advantages and disadvantages of this method of construction, compared with the use of wiring forms.*

(*c*) *Briefly describe how the printed circuit is coupled to power supplies and other associated circuits.*

A. 7. See A.3, Elementary Telecommunication Practice, 1972, *Supplement*, Vol. 66, p. 6, Apr. 1973.

Q. 8. (*a*) *State 3 types of insulation which are likely to be found in capacitors and, by reference to their properties, say what are their fields of application.*

(*b*) *State the factors on which the choice of a capacitor depends in a particular installation.*

A. 8. (*a*) Three insulators used in capacitors are air, impregnated paper and mica.

(*i*) Air has a low permittivity, low dielectric losses, and good insulating properties and electric strength. Air-dielectric capacitors are, therefore, used in situations where capacitors of a small and stable value are required. In general, dielectric losses increase with frequency, so that air-dielectric capacitors can be used at high frequencies. The design of variable capacitors is simplified by the use of air, rather than a solid, as the dielectric.

(*ii*) Impregnated paper is used to insulate metallic films, which are tightly rolled together to provide capacitors of medium value, typically, in the range 1–10 μF. The paper dielectric has fairly large electrical losses and, so, cannot be used at high frequencies. Although its electric strength is good, the area of paper used in a typical capacitor is large, and weak spots limit the safe voltage which can be applied. Capacitors of this type are, therefore, used in low-voltage, low-frequency circuits.

(*iii*) Mica is a rigid, stable and homogeneous material. Its dielectric losses are low and it has a good electric strength. Its physical structure does not permit bending and, so, it is used as the insulator between stacked metallic plates. For these reasons, mica-dielectric capacitors are used where small fixed values of capacitance are required to work at high voltages in high-frequency circuits.

(*b*) The factors to be considered in choosing a capacitor for a particular purpose include

(*i*) the capacitance required and the acceptable tolerances,
(*ii*) the need for variability or adjustment,
(*iii*) the voltage rating,
(*iv*) the permissible loss factor,
(*v*) the working frequency,
(*vi*) the available space, and
(*vii*) the ambient temperature and humidity.

Q. 9. (*a*) *A local automatic telephone network consists of a central exchange, providing 2000 lines, and 3 satellite exchanges, each providing 600 lines. Develop a numbering scheme to enable all local traffic to be routed automatically.*

(*b*) *If the central exchange is located in a town with typical business activity, sketch a graph of the likely distribution of traffic through a mid-week 24 h period.*

(*c*) *By reference to the graph, show what is meant by the busy hour.*

A. 9. See A.4, Elementary Telecommunication Practice, 1973, *Supplement*, Vol. 67, p. 7, Apr. 1974.

Q. 10. (*a*) *List the descriptions of the numbered components in Fig. 1.*

(*b*) *Describe the purpose and operation of the circuit.*
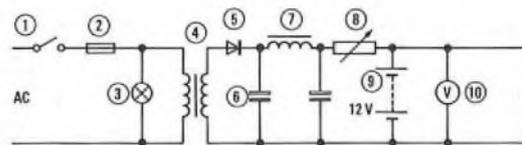
Fig. 1

A. 10. (*a*) The components are listed in the following table, together with their descriptions.

Component No.	Description
1	Switch (*make* contact)
2	Fuse
3	Signal lamp
4	Iron-cored transformer
5	Semiconductor rectifying diode
6	Electrolytic capacitor
7	Iron-cored inductor
8	Variable resistor
9	12 V battery
10	Voltmeter

(*b*) The diagram represents a simple power supply of the floated-battery type. The a.c. input is applied to the primary winding of the transformer through the switch and protective fuse. The secondary voltage is rectified, and the resulting unidirectional supply is smoothed by the filter formed by the 2 capacitors and the inductor. When the load current is heavy, the battery assists the rectified supply, and when the load current is light, the battery is recharged. These conditions can be controlled by the variable resistor. The lamp indicates that the switch is closed, and the voltmeter monitors the voltage applied to the load.

CORRECTION

ELEMENTARY TELECOMMUNICATION PRACTICE, 1973

(*Supplement*, Vol. 67, Apr. 1974)

A. 7. (*a*) (*i*) In the Morse code, letters are separated by an interval equal in length to 3 dots, and not 5 as stated.

TELECOMMUNICATION PRINCIPLES A, 1974

Students were expected to answer 6 questions, including at least one from questions 9–10

Q. 1. (*a*) *When resistance is measured by the ammeter-voltmeter method, 2 different circuit connexions are possible.*

(*i*) *Explain the principle of the resistance measurement.*

(*ii*) *Sketch each circuit, and carefully explain why one is more suitable for measuring high resistance, and the other for measuring low resistance.*

(*b*) (*i*) *Explain how the value of an unknown resistor may be determined using the substitution method. Include a circuit diagram in your answer.*

(*ii*) *Give one possible source of error.*

A. 1. (*a*) (*i*) Using the ammeter-voltmeter method, an unknown resistance value is determined by measuring the voltage across the resistance, due to a measured current flowing through it, and calculating the value from Ohm's law:

resistance (Ω) = voltage (V)/current (A).

(*ii*) The 2 circuit connexions possible are shown in sketches (*a*) and (*b*).

(*a*)

(*b*)

(*c*)
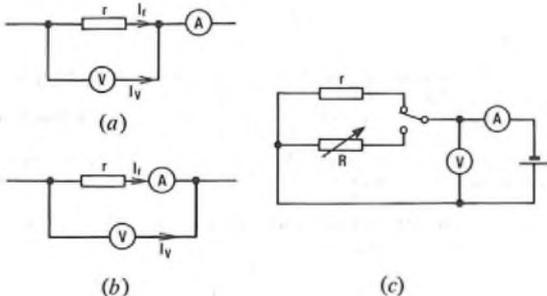
In the short-shunt circuit (sketch (*a*)), the voltmeter indicates the voltage, V_r, across the unknown resistance, r, but the ammeter indicates the sum of the currents flowing in the resistance and the voltmeter, $I_r + I_V$. Hence, the value of the resistance is given by $r = V_r/(I_r+I_V)$.

Provided that I_V is small compared with I_r, the error due to I_V is negligible. The short-shunt circuit is, therefore, more suitable for low values of r, as I_r is then relatively high compared with I_V.

In the long-shunt circuit (sketch (*b*)), the ammeter indicates the current flowing in the resistance, but the voltmeter indicates the sum of the voltage drops across the resistance and the ammeter, $V_r + V_A$. Hence, the value of the resistance is given by $r = (V_r + V_A)/I_r$.

Provided that V_A is small compared with V_r, the error due to V_A is negligible. The long-shunt circuit is, therefore, more suitable for high values of r, as V_r is then relatively high compared with V_A.

(*b*) (*i*) Sketch (*c*) shows a circuit diagram for the substitution method of determining the value of an unknown resistance, r. The variable resistor, which must be calibrated and stable, with a negligible temperature coefficient of resistance, is adjusted until the ammeter gives the same reading for both positions of the change-over switch. Then, $r = R$. The value of the battery voltage is not important, but must remain constant during a set of measurements, so that the basis of comparison for the 2 switch positions is the same.

(*ii*) A possible source of error is a variation in the battery voltage during a set of measurements.

Q. 2. (*a*) *What is meant by the terms*

(*i*) *magnetic saturation, and*

(*ii*) *relative permeability?*

(*b*) *An iron ring has a mean diameter of* 0·4 *m and a cross-sectional area of* 0·0005 m^2. *It is wound with a coil of* 2000 *turns. Calculate the current required in the coil to produce a flux of* 0·5 *mWb in the ring. The relative permeability of the iron under these conditions is* 1000.

(*c*) *Explain the change in magnetic effect when the ring is made in 2 halves (as in a C-core). The magnetizing current remains constant.*

A. 2. (*a*) (*i*) Magnetic saturation is the condition prevailing when the magnetic flux density in a magnetic material no longer increases as the magnetizing force rises. The magnetic domains are then all aligned, to give the maximum magnetic flux the material can sustain.

(*ii*) The relative permeability of a material is the factor by which the flux density within a solenoid carrying a current increases due to the presence of the material, compared with the flux density obtained with air as the core material.

(*b*) Now, $B = \dfrac{\Phi}{a}$, where B is the magnetic flux density (T),

Φ is the magnetic flux (Wb), and

a is the cross-sectional area of the ring (m^2),

$$= \frac{0·5 \times 10^{-3}}{0·0005} = 1 \text{ T}.$$

Also, $H = \dfrac{B}{\mu_0 \mu_r} = \dfrac{NI}{l}$, where H is the magnetizing force (A/m),

μ_0 is the absolute permeability of free space, equal to $4\pi \times 10^{-7}$ H/m,

μ_r is the relative permeability of the material,

N is the number of turns on the coil,

I is the current flowing in the coil (A), and

l is the mean circumference of the ring (m).

Hence, $\dfrac{1}{4\pi \times 10^{-7} \times 1000} = \dfrac{2000 \times I}{\pi \times 0·4}$.

$$\therefore I = 0·5 \text{ A}.$$

(*c*) If the ring is made of 2 C-cores butted together, it has an increased reluctance due to imperfections at the joints. Therefore, the magnetic flux set up in the iron, for a given value of magnetomotive force, decreases compared with that for the same ring without joints.

Q. 3. *Two currents,* $i_1 = 3 \sin \omega t$ *amperes and* $i_2 = 5 \sin (\omega t + \pi/2)$ *amperes, are flowing through the same* 10 Ω *resistor.*

(*a*) *Plot one cycle of each current waveform on the same axes and, hence, obtain graphically the waveform of the resultant current.*

(*b*) *From the waveforms, find*

(*i*) *the maximum value of the resultant current,*

(*ii*) *the phase angle between the resultant current and* i_1, *and*

(*iii*) *the maximum voltage across the* 10 Ω *resistor.*

(*c*) *What should be the preferred power rating of the resistor?*

A. 3. (*a*) The current waveforms are shown in the sketch. Both are sine waves of the same angular frequency, ω radians/second, and hence, the resultant must be a sine wave of the same frequency. Waveforms i_1 and i_2 have peak values of 3 A and 5 A respectively, and i_2 is advanced in phase by $\pi/2$ rad (90°) over i_1. The resultant is obtained graphically by addition of the ordinates of i_1 and i_2 at, say, $\pi/6$ rad (30°) intervals.

15

(*b*) (*i*) From the sketch, the maximum value of the resultant current, $i_1 + i_2$, is 5·83 A.
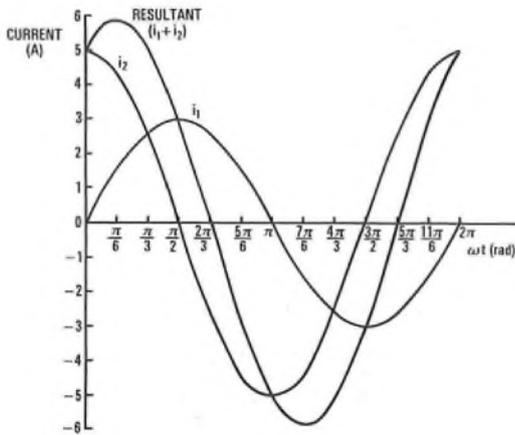
(*ii*) The phase angle between the resultant and i_1 can be taken as the angle between points of zero amplitude of these currents. From the sketch, the phase angle is 1 rad (57·3°).

(*iii*) The maximum voltage developed across the 10 Ω resistor, by Ohm's law, is $5·83 \times 10 = 58·3$ V.

(*c*) The total power dissipated in the resistor is given by the product of the square of the r.m.s. value of the resultant current and the value of the resistor.

Thus, power dissipated $= \left(\dfrac{5·83}{\sqrt{2}}\right)^2 \times 10 = 170$ W.

Therefore, the preferred power rating of the resistor is some value in excess of 170 W; for example, 200 W.

Q. 4. (*a*) *In Fig. 1, derive an expression for the voltage, V_1, across resistor R1 in terms of the other symbols on the diagram.*
(*b*) *In Fig. 2, derive an expression for the current I_2 in terms of the other symbols on the diagram.*
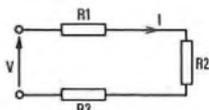
Fig. 1
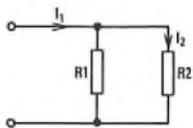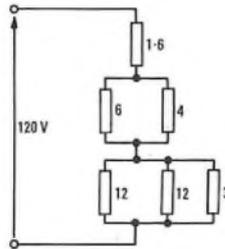
Fig. 2   *Fig. 3*

(*c*) *The circuit shown in Fig. 3 is connected across a 120 V d.c. supply. Calculate*
(*i*) *the voltage across the 1·6 Ω resistor,*
(*ii*) *the current flowing through the 4 Ω resistor, and*
(*iii*) *the power dissipated in the 6 Ω resistor.*

A. 4. (*a*) From Ohm's law, the current, I amperes, flowing in the circuit shown in Fig. 1 is given by $I = V/(R_1 + R_2 + R_3)$, and the voltage across resistor R1 is given by $V_1 = IR_1$.

$$\therefore \ V_1 = \frac{VR_1}{R_1 + R_2 + R_3} \text{ volts.}$$

(*b*) The voltage across resistor R1, V volts, is the same as the voltage across resistor R2.

$$\therefore \ V = I_2R_2 = (I_1 - I_2)R_1.$$

$$\therefore \ I_2(R_2 + R_1) = I_1R_1.$$

$$\therefore \ I_2 = \frac{I_1R_1}{R_1 + R_2} \text{ amperes.}$$

(*c*) Let R_x be the resistance of the 6 Ω and 4 Ω resistors in parallel, and R_y be the resistance of the two 12 Ω and the 3 Ω resistors in parallel.

Then,
$$R_x = \frac{1}{\frac{1}{6} + \frac{1}{4}} = 2·4 \ \Omega,$$

and
$$R_y = \frac{1}{\frac{1}{12} + \frac{1}{12} + \frac{1}{3}} \text{ F} = 2 \ \Omega.$$

Hence, the total resistance of the circuit $= 1·6 + 2·4 + 2 = 6$ Ω. Therefore, the total current flowing $= 120/6 = 20$ A.

(*i*) The voltage across the 1·6 Ω resistor $= 20 \times 1·6 = 32$ V.
(*ii*) The voltage across the 6 Ω and 4 Ω resistors in parallel
$$= 20 \times 2·4 = 48 \text{ V.}$$
Therefore, the current flowing through the 4 Ω resistor
$$= 48/4 = 12 \text{ A.}$$
(*iii*) The power dissipated in the 6 Ω resistor $= 48^2/6 = 384$ W.

Q. 5. (*a*) *Define the unit of capacitance.*
(*b*) *On what factors does the capacitance of a capacitor depend?*
(*c*) *For the circuit shown in Fig. 4, Calculate*
(*i*) *the equivalent capacitance,*
(*ii*) *the total energy stored, and*
(*iii*) *the voltage across the 1 μF capacitor.*
(*d*) *What should be the working voltage rating of the 1 μF capacitor?*
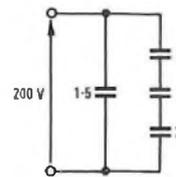
Fig. 4

A. 5. (*a*) The unit of capacitance is the farad, and is that capacitance which requires a charge of 1 C to raise its potential by 1 V.

(*b*) The capacitance, C farads, of a capacitor is proportional to the area of the conducting surfaces forming the capacitor, A metres^2, and the absolute permittivity of the dielectric, ϵ farads/metre, and inversely proportional to the distance, d metres, between the conducting surfaces. Hence, $C = \epsilon A/d$ farads.

(*c*) Let C_s be the total capacitance of the two 2 μF capacitors and the 1 μF capacitor in series.

Then,
$$C_s = \frac{1 \times 10^{-6}}{\frac{1}{1} + \frac{1}{2} + \frac{1}{2}} \text{ F} = 0·5 \ \mu\text{F.}$$

(*i*) The total equivalent capacitance of the circuit
$$= C_s + 1·5 \ \mu\text{F,}$$
$$= 2 \ \mu\text{F.}$$

(*ii*) The total energy stored $= \frac{1}{2}CV^2$ joules, where V is the applied voltage (V).
$$= \frac{2 \times 10^{-6} \times 200^2}{2},$$
$$= 0·04 \text{ J.}$$

(*iii*) The total charge stored by the two 2 μF capacitors and the 1 μF capacitor in series, Q_s coulombs, is given by
$$Q_s = C_sV = 0·5 \times 10^{-6} \times 200 = 100 \times 10^{-6} \text{ C.}$$

As this is also the charge on each of the capacitors in the series circuit, then the voltage across the 1 μF capacitor
$$= \frac{Q_s}{1 \times 10^{-6}} = \frac{100 \times 10^{-6}}{1 \times 10^{-6}} = 100 \text{ V.}$$

(*d*) The working voltage rating of the 1 μF capacitor should be at least twice the value of the voltage across it; that is, greater than 200 V.

Q. 6. *Describe one of the following experiments:*
(*a*) *the determination of the forward and reverse characteristics of a p n junction, or*

(*b*) *the determination of the variation of the reactance and impedance of an inductor with frequency.*

Set out your answer in the form of a laboratory report, which should include a circuit diagram, details of the equipment and procedure, typical results and conclusions.

Q. 7. *The generator shown in Fig. 5 supplies a voltage, V_s, at a frequency of 1000 Hz.*

(*a*) *Sketch and explain the construction of a phasor diagram showing the current and voltages V_1, V_2 and V_s.*

(*b*) *Find*

(*i*) *the voltages V_1, V_2, and V_s,*
(*ii*) *the phase angle of the circuit, and*
(*iii*) *the power dissipated by the circuit.*
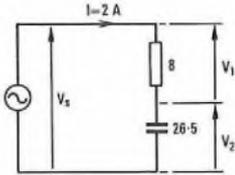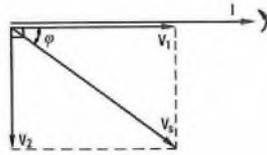
Fig. 5   (*a*)

A. 7. (*a*) The phasor diagram is shown in sketch (*a*). Phasor I represents the current, and is the reference phasor. Phasor V_1 is drawn in phase with the current, its magnitude being given by the product of the current and the resistance. Phasor V_2 is drawn lagging the current by 90°, its magnitude being given by the product of the current and the reactance of the capacitor. Phasor V_s is the resultant, and represents the supply voltage.

(*b*) (*i*) $V_1 = 2 \times 8 = \underline{16\ V.}$

$$V_2 = \frac{2}{2 \times \pi \times 1000 \times 26 \cdot 5 \times 10^{-6}} = \underline{12\ V.}$$

$$V_s = \sqrt{(16^2 + 12^2)} = \underline{20\ V.}$$

(*ii*) The phase angle, $\phi = \tan^{-1}\left(\dfrac{12}{16}\right) = \underline{36°\ 52'.}$

(*iii*) Power is dissipated only in the resistor, and is equal to the square of the current times the resistance.

$$\therefore \text{power dissipated} = 2^2 \times 8 = \underline{32\ W.}$$

Q. 8. (*a*) *Explain what is meant by the term self-inductance, and give one factor on which it depends.*

(*b*) *An inductor, having a resistance of $1\ \Omega$ and an inductance of $2\ H$, is connected across a d.c. supply. The current increases from 0 to 20 A in $0 \cdot 5$ s. Calculate*

(*i*) *the average value of the induced e.m.f.,*
(*ii*) *the energy stored after $0 \cdot 5$ s, and*
(*iii*) *the power needed to sustain the current at 20 A.*

(*c*) *State one practical application of such an inductor.*
(*d*) *Where is the energy stored in the inductor?*

A. 8. (*a*) A circuit possesses self-inductance if a change in current, accompanied by a change in magnetic flux, produces an induced e.m.f. in the circuit.

The self-inductance, L henrys, of a coil is given by

$L = \dfrac{\mu_0 \mu_r N^2 a}{l}$, where μ_0 is the absolute permeability of free space, equal to $4\pi \times 10^{-7}$ H/m,

μ_r is the relative permeability of the core material,

N is the number of turns on the coil,

a is the cross-sectional area of the coil (m²),

and

l is the length of the coil (m).

From this, it can be seen that the nature of the core material affects the self-inductance. If the core is soft iron, rather than air, a very much larger rate of change of magnetic flux is produced for a given rate of change of current, since the permeability of soft iron is much greater than that of air. Hence, a higher value of induced e.m.f. is produced.

(*b*) (*i*) The value of the induced e.m.f.

$= L\dfrac{di}{dt}$ volts, where di/dt is the rate of change of current (A/s),

$= 2 \times \dfrac{20}{0 \cdot 5} = \underline{80\ V.}$

(*ii*) The energy stored $= \frac{1}{2}LI^2$ joules, where I is the current (A),

$$= \frac{2 \times 20^2}{2} = \underline{400\ J.}$$

(*iii*) Power is dissipated only in the resistance of the inductor. Hence, the power needed

$= I^2R$ watts, where R is the resistance of the inductor (Ω),

$= 20^2 \times 1 = \underline{400\ W.}$

(*c*) A practical application of such an inductor is as a series element in the smoothing circuit following a rectifier, where it tends to oppose the variations in current caused by the ripple voltage from the rectifier.

(*d*) When an inductor carries a current, energy is stored in the magnetic field around the inductor, since energy is initially required to set up the field. When the current ceases, the field collapses, returning its energy to the circuit of the inductor.

Q. 9. (*a*) *Draw a simple circuit, with typical component values, of a single-stage audio-frequency triode-valve amplifier with a resistive load. Explain the action of this amplifier.*

(*b*) *The table shows the relationship between the anode current, I_a milliamperes, and the anode–cathode voltage, V_a volts, for various values of grid-bias voltage, V_g volts.*

(*i*) *Plot these anode characteristics.*
(*ii*) *Draw the load line for a resistive anode load of $10\ k\Omega$ with an h.t. supply voltage of $250\ V$.*

V_g (V)	I_a (mA) for values of $V_a =$					
	50 V	100 V	150 V	200 V	250 V	275 V
0	3	8	14·5	24	—	—
−1	1·5	4	8	13	19	—
−1·5	—	2·5	5·5	9	14	17·5
−2	—	1·5	3	6	10	12·5
−3	—	—	1	2·5	5	7

(*c*) *Hence, find*

(*i*) *the fixed negative bias that should be used, and*
(*ii*) *the output voltage across the anode load that will result from an r.m.s. input signal of $0 \cdot 707\ V$.*

A. 9. (*a*) Sketch (*a*) shows the circuit diagram of a single-stage audio-frequency triode amplifier, with typical component values
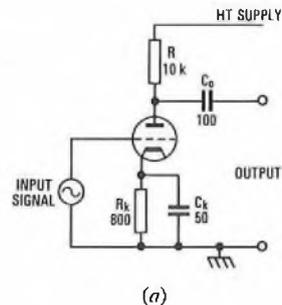

(*a*)

A thermionic triode valve contains 3 electrodes: the anode, cathode and grid, all carefully positioned within an evacuated glass envelope. The cathode, which is connected to the negative side of the h.t. supply, is heated until its specially-coated surface emits electrons. These, being negatively charged, are attracted to the anode, which is connected to the positive side of the h.t. supply. The control grid, consisting of a

wire mesh, is mounted in the path of the electron stream from the cathode to the anode. The grid is biased slightly negatively with respect to the cathode by means of resistor R_k, so that it tends to repel the negatively-charged electrons. The number of electrons reaching the anode is, therefore, reduced as the grid becomes more negative, or increased as it becomes less negative. Hence, the intensity of the electron stream, and thus, the anode current, depends on the grid potential.

The input signal to be amplified is connected across the grid and cathode, so that the grid voltage varies in accordance with the signal. As a result, the anode current through resistor R, and hence, the voltage across resistor R, varies in sympathy with the input signal, voltage amplification being obtained according to the values of resistor R and the anode current. The amplified signal voltage across resistor R is delivered to the output via capacitor C_o, which isolates the output from the quiescent direct voltage on the anode, but which presents a low impedance at the signal frequency.

Capacitor C_k also has a low impedance at the signal frequency and, thus, tends to short circuit resistor R_k to the alternating signal currents, hence reducing their attenuation due to resistance in the anode–cathode path.

(*b*) (*i*) The anode characteristics are shown in sketch (*b*).
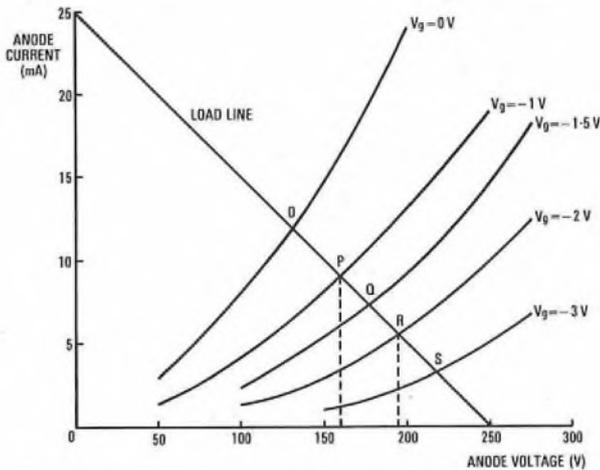
(*b*)

(*ii*) Now, $E = V_a + I_a R$, where E is the h.t. supply voltage.

Therefore, when $I_a = 0$, $V_a = E = 250$ V,

and when $V_a = 0$, $I_a = \dfrac{E}{R} = \dfrac{250}{10 \times 10^3}$ A $= 25$ mA.

The load line is plotted from these 2 points (see sketch (*b*)).

(*c*) (*i*) For minimum distortion of the signal, a value of grid-bias voltage is chosen which lies at the centre of that section of the load line which is divided into portions of equal length by the intersecting characteristics. From sketch (*b*), it can be seen that $V_g = -1 \cdot 5$ V is the required value of grid-bias voltage (point Q).

(*ii*) The voltage amplification, M, of the valve is given by the change in anode voltage divided by the change in grid voltage. From sketch (*b*), if the grid voltage changes from -1 V to -2 V, the anode voltage changes from 160 V to 195 V.

$$\therefore M = \frac{195 - 160}{2 - 1} = 35.$$

Hence, for an r.m.s. input signal of $0 \cdot 707$ V, the output voltage

$$= 35 \times 0 \cdot 707 = \underline{24 \cdot 75 \text{ V r.m.s.}}$$

Note: The output resulting from an input signal of $0 \cdot 707$ V r.m.s. is likely to be distorted, since the peak-to-peak value of the input signal is 2 V. From sketch (*b*), it is apparent that the maximum value of input signal for distortionless operation is between $V_g = -1$ V and $V_g = -2$ V (that is, a peak-to-peak input signal of 1 V), since points P and R are equidistant from point Q, whereas points O and S are not.

Q. **10.** (*a*) *Explain, with the aid of sketches, the principle of a moving-coil loudspeaker.*
(*b*) *On what factors does the deflecting force on the coil depend?*
(*c*) *State 2 factors that affect the power-handling capacity.*

A. **10.** (*a*) The moving-coil loudspeaker operates on the principle that a force acts on a conductor carrying a current in a magnetic field. The sketch shows a cross-section of a typical loudspeaker.
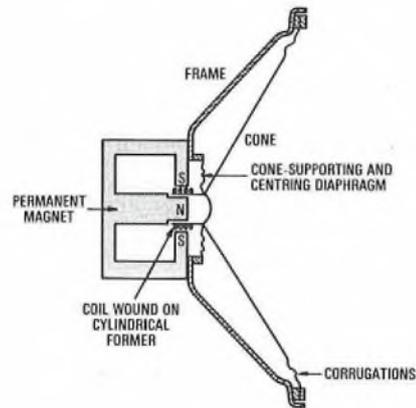
A coil is wound on a light cylindrical former, constrained by a diaphragm to move axially within the annular air-gap of a permanent-magnet system. Connexions to the coil are made via flexible leads. Signal currents, flowing in the coil, set up a varying magnetic field around the coil which interacts with the permanent magnetic field, causing a varying force to be exerted on the coil. This force causes the coil to move axially in sympathy with the amplitude and frequency of the signal currents and, since the coil is attached to a cone, this movement is transmitted to it. The movement of the cone creates sympathetic pressure variations in the air adjacent to it, thus giving rise to sound waves which reproduce the intelligence contained in the signal currents.

The cone is light and firmly constructed, so that it acts as a piston without distorting during its action. It is attached to a frame via flexible corrugations around its major circumference.

(*b*) The deflecting force on the coil is proportional to the magnetic flux density in the air gap, the current flowing in the coil, and the length of the conductor forming the coil.

(*c*) The power-handling capacity of a loudspeaker is the input power at which distortion just becomes audible. It is affected by the force which can be exerted on the coil (see part (*b*)), and the free axial movement of the coil and cone. Flexing of the cone gives rise to the production of harmonics, so that the rigidity of the cone also limits its power-handling capacity.

TELEPHONY AND TELEGRAPHY A, 1974

Students were expected to answer any 6 questions

Q. **1.** (*a*) *Explain why an elementary telephone instrument produces sidetone.*
(*b*) *With the aid of a simplified circuit diagram, explain how sidetone can be reduced.*
(*c*) *What are the effects of local sidetone on*

(*i*) *the transmission of speech, and*
(*ii*) *the reception of speech?*

A. **1.** (*a*) The circuit diagram of an elementary telephone instrument is shown in sketch (*a*). Since both the transmitter and receiver are connected to the same pair of wires, the user's speech and local room noises are reproduced in the local receiver, as well as being

(*a*)   (*b*)
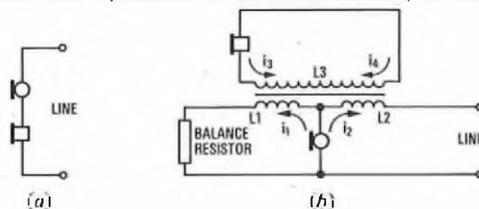
transmitted via the line to the distant instrument. This effect is called *sidetone*.

(*b*) Sketch (*b*) illustrates the principle of a telephone equipped with an anti-sidetone induction coil. The 3-winding coil is arranged as a balanced bridge, the balance impedance being matched to that of an average subscriber's line. Speech current from the transmitter splits into 2 components, i_1 and i_2, flowing in the balance resistor and line respectively. The current i_1, in winding L1, induces an e.m.f. in winding L3 to produce a current, i_3. Similarly, current i_2 flowing in winding L2 produces a current, i_4, in winding L3. The windings are arranged so that currents i_3 and i_4 are in opposite directions and, if the bridge is perfectly balanced, cancel each other. In practice, perfect balance is not achieved, since balance networks are designed for an average line, and very few lines are precisely average. However, a small amount of sidetone is desirable, and the basic circuit is modified to ensure that some sidetone is always present.

(*c*) (*i*) Local sidetone reduces the level of speech transmitted to line in 2 ways. The effect of local sidetone on a user is to cause him to lower his voice and, thus, to reduce the power input to the transmitter. In addition, sidetone is produced by power dissipated in the local receiver, and this power is, therefore, ineffective with respect to the distant receiver.

(*ii*) During the reception of speech, local sidetone results in the reproduction of local room noise in the local receiver, and this noise may mask incoming speech.

Q. **2.** (*a*) *Sketch the construction of a reed-relay insert and identify the materials used.*

(*b*) *Explain how operation of the reed insert can be controlled electrically.*

(*c*) *Explain why a reed relay is fast to operate and release.*

A. 2. (*a*) Sketch (*a*) shows a typical reed-relay insert and identifies the materials used in its construction.

(*a*)   (*b*)
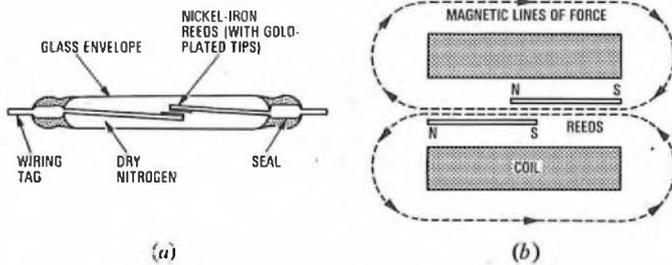
(*b*) Reed-relay inserts are controlled electromagnetically. In the presence of a magnetic field of such a direction that the magnetic lines of force run along the length of the reeds, each reed acquires a magnetic polarity with opposite poles appearing where they overlap, as shown in sketch (*b*). Hence, the reeds are attracted together, thus closing an external electrical circuit. When the magnetic field is removed, the reeds return to the open-circuit position, due to an outward tension imparted to them during construction.

In practice, a number of reed-relay inserts, typically 4, are inserted into a hollow bobbin on which the electromagnetic operating coil is wound. Control of the current in the coil, therefore, operates and releases a group of reed contacts.

(*c*) The following features contribute to the fast operate and release characteristics of the reed relay.

(*i*) The low mass of the moving parts means they have small inertia and require only a small tractive force to accelerate them.

(*ii*) The small contact gap means that the distance which the reeds have to travel is small and, hence, transit time is low.

(*iii*) The absence of a heavy metal yoke and core, in which eddy currents can circulate, helps to ensure that the controlling magnetic flux rises and falls rapidly.

Typically, the operate and release times of a reed relay are of the order of 1 ms.

Q. **3.** (*a*) *Explain the electrical action of a heel-end slug on a telephone relay.*

(*b*) *Describe 2 other methods of achieving a similar effect on the performance of a relay.*

A. 3. (*a*) A slug on the core of a telephone relay is, in effect, a single short-circuited turn of extremely low resistance. It acts as the secondary winding of a transformer with the relay's operating coil as the primary. A rise or fall in current in the operating coil causes a change in core

flux which links with the slug. This change in flux linkage produces an e.m.f. in the slug, which results in a current being set up within the slug. Because of the very low resistance of the slug, the value of this current is large. The current produces a flux in such a direction as to oppose the change which gave rise to it. Thus, when the relay operates, the slug flux is in opposition to the coil flux, and when the relay releases, the slug flux is in the same direction as the decaying coil flux.

By fitting the slug at the heel end of the relay, the operation of the armature is not delayed significantly, and the relay's operating time is normal; that is, fast. This is because the flux paths are such that the full coil flux flows through the armature air-gap, as shown in sketch (*a*). At the point where the opposing fluxes meet, both flux paths cross the core-yoke air-gap and follow separate paths, as shown.

(*a*)   (*b*)
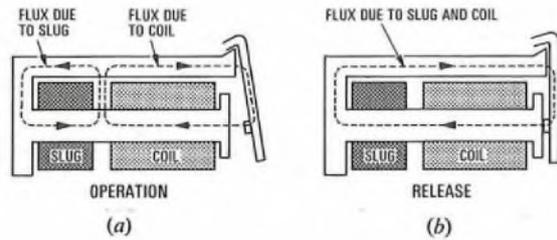
However, on release of the relay, the slug flux aids the decaying coil flux, and both fluxes follow the same path, as shown in sketch (*b*). Thus, the slug flux flows through the armature, and maintains the relay in the operated position for some time after the coil current has ceased, thus causing the relay to be slow to release.

(*b*) Sketches (*c*) and (*d*) show 2 other methods of achieving a fast-to-operate, slow-to-release feature.

(*c*)   (*d*)
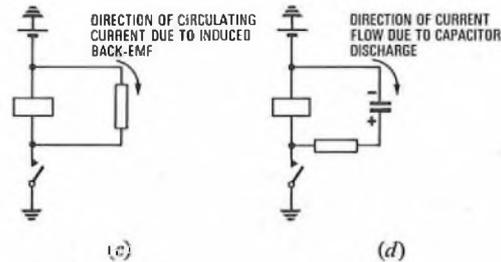
Sketch (*c*) shows a non-inductive resistor in parallel with the relay coil. Provided that the normal 50 V potential is maintained across the parallel combination, the rise of current and flux and, hence, the operating time of the relay, are independent of the value of the resistor. On release, however, the resistive shunt forms a closed path, and the relay's self-induced back-e.m.f. causes a circulating current to flow in such a direction as to maintain the main flux. Thus, the relay is slow to release.

Sketch (*d*) shows a capacitor connected in parallel with the relay coil. The capacitor becomes charged when the relay is operated. When the operating circuit is disconnected, the energy stored in the capacitor is discharged via the relay winding and, thus, maintains the operating current for some time, making the relay slow to release. A small resistor is included in the circuit to limit the charging current of the capacitor.

Q. **4.** (*a*) *With the aid of sketches of the circuit elements concerned, describe the pulse-sending and pulse-receiving arrangements for a loop-disconnect pulsing circuit between 2 exchanges.*

(*b*) *What limitations apply to the loop-disconnect method of pulsing?*

(*c*) *Outline one other method of pulsing which can be used where loop-disconnect pulsing is unsuitable.*

A. 4. (*a*) Loop-disconnect pulsing between 2 exchanges is similar in principle to that between a subscriber and the local exchange, except that the transmitting element is a contact of a pulse-repeating relay, instead of dial contacts. The circuit elements concerned are shown in sketch (*a*).

(*b*) Loop-disconnect signalling is restricted to circuits with a physical signalling path; it cannot be used over carrier, coaxial or radio circuits.

On physical circuits, the use of loop-disconnect signalling is subject to the limitations imposed by the following conditions.

(*i*) The loop resistance of the circuit must be low enough to permit the operation of the distant pulse-receiving relay. Too high a value of resistance limits the current to a value lower than that required to operate the relay. This condition limits both the physical length of the circuit and the use of small-gauge conductors.

(*ii*) The leakance of the circuit must be low enough to ensure the release of the distant pulse-receiving relay during pulsing.

(*iii*) The capacitance of the circuit must be low enough to ensure the release of the distant pulse-receiving relay during pulsing.


(*a*)   (*b*)
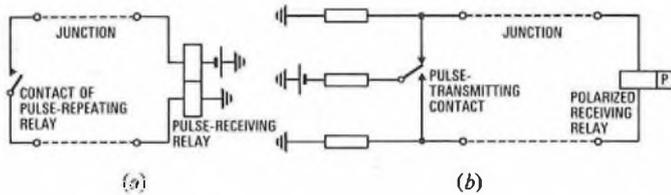
(*c*) Sketch (*b*) shows an arrangement known as single-commutation d.c. signalling, using a polarized receiving relay. This has the following advantages over loop-disconnect pulsing.

(*i*) The polarized relay is very sensitive and, therefore, higher line-resistance limits are possible.

(*ii*) The use of double-current signalling (that is, *make* and *break* signals are represented by opposite directions of current flow) ensures that line leakance cannot suppress *break* signals. Also, the use of double-current signalling minimizes the effect of line capacitance, since the line is discharged more quickly by the application of the opposite potential.

Q. 5. (*a*) *For what purpose are signalling lamps provided at a manual switchboard?*

(*b*) *Briefly explain how the lamps are electrically controlled according to their purpose. Illustrate your answer with sketches of the circuit elements concerned.*

(*c*) *What advantage does a series multiple offer over a parallel multiple for the control of lamp signals?*

A. 5. (*a*) At a manual switchboard, signalling lamps are typically used to indicate

(*i*) a calling line in the answering multiple,

(*ii*) the progress of a call, by means of a supervisory lamp in the cord circuit,

(*iii*) an engaged outgoing circuit, and

(*iv*) the next free line in an outgoing group (free-line signalling).

(*b*) (*i*) Sketch (*a*) illustrates how the calling loop from a subscriber operates the line relay, relay L, which lights the CALL lamp.

(*ii*) Sketch (*b*) shows the principle of operation of the SUPERVISORY lamp circuit. During conversation, relay P is operated from the battery-feeding relay, the ring wire, the telephone loop, the tip wire, relay P, and earth. On cleardown, relay P releases and contact P1 lights the SUPERVISORY lamp.

(*a*)   (*b*)

(*c*)   (*d*)
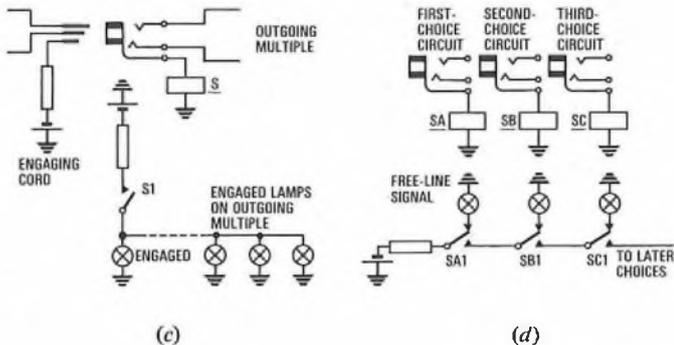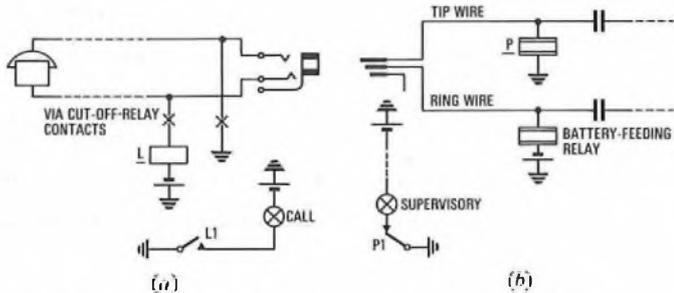
(*iii*) Sketch (*c*) shows how an engaging cord operates relay S to its sleeve battery for an outgoing circuit. Contact S1 lights the ENGAGED lamps associated with each appearance of that circuit in the outgoing multiple.

(*iv*) Sketch (*d*) shows the principle of free-line signalling, which provides an indication of the lowest-numbered free outgoing circuit in a group. On engaged circuits, relay SA, SB, SC etc., as appropriate, operates to the battery on the sleeve, hence transferring the free-line indication to the next free circuit.

(*c*) A series multiple consists of a number of break jacks in series, with the line relay connected via the inner springs. When a plug is inserted at any position, it breaks the contact between the inner and outer springs, releases relay L and, thus, extinguishes the CALL lamp.

A parallel multiple consists of a number of plain jacks in parallel. No inner break springs are fitted, and the line relay is connected via cut-off-relay contacts as shown in sketch (*a*). When an answering plug is inserted, the cut-off relay operates and disconnects relay L.

The main advantage of the series multiple is that it dispenses with the need for a separate cut-off relay.

Q. 6. (*a*) *For a uniselector, sketch the stepping mechanism and describe its operation.*

(*b*) *Draw a circuit element suitable for hunting and testing, and describe its operation.*

A. 6. See A.6 and A.10, Telephony and Telegraphy A, 1972, *Supplement*, Vol. 66, p. 94 and 95, Jan. 1974.

Q. 7. (*a*) *Describe the 5-unit telegraph code.*

(*b*) *Why is it preferred for machine telegraphy?*

(*c*) *Explain how start-stop working permits successful operation whilst the sending and receiving machines are running at slightly different speeds.*

A. 7. (*a*) See A.6, Telephony and Telegraphy A, 1970, *Supplement*, Vol. 64, p. 61, Oct. 1971.

(*b*) The 5-unit code is preferred for machine telegraphy because all of the code elements and characters are of the same duration. In addition, the code, whilst having sufficient capacity for transmitting the required intelligence, is economical in that it contains very few redundant combinations.

(*c*) Start-stop working permits successful operation whilst the sending and receiving machines are running at slightly different speeds, because the *start* and *stop* signals are used to synchronize the receiving machine by means of a clutch. Therefore, the machines are synchronized at the start of each character and, hence, any difference in speed does not give rise to a cumulative time difference over a number of characters.

Q. 8. (*a*) *Draw a trunking diagram for a 4000-line Strowger telephone exchange. Include the various distribution frames through which calls are routed.*

(*b*) *Name the supervisory tones a caller might hear, and indicate the points at which they could be applied.*

(*c*) *Consider the following alternative arrangements.*

(*i*) *100-outlet group selectors with 200-outlet final selectors.*

(*ii*) *200-outlet group selectors with 100-outlet final selectors.*

State which arrangement would tend to increase, and which would tend to decrease, the total number of gradings. Give reasons for your answers.

A. 8. (*a*) The sketch shows a trunking diagram for a 4000-line Strowger exchange.

Note: The symbol used for the 2-motion selectors is that recommended in British Standard 3939.

(*b*) The supervisory tones, and the points at which they are applied, are shown in the following table.

Tone	Point of Application
Dial tone	First selector
Ring tone	Final selector
Busy tone	Final selector
Equipment-engaged tone	The eleventh-step bank contact of a first-selector or second-selector level
Number-unobtainable tone	A spare level of a first or second selector, or a spare line on a final selector
Pay tone	Coin-and-fee-checking relay-set

(*c*) (*i*) The arrangement of 100-outlet group selectors with 200-outlet final selectors tends to increase the number of gradings because, as the group-selector availability is only 10, there are more final selectors per pair of group-selector levels than for the alternative arrangement.

MDF: MAIN DISTRIBUTION FRAME
IDF: INTERMEDIATE DISTRIBUTION FRAME
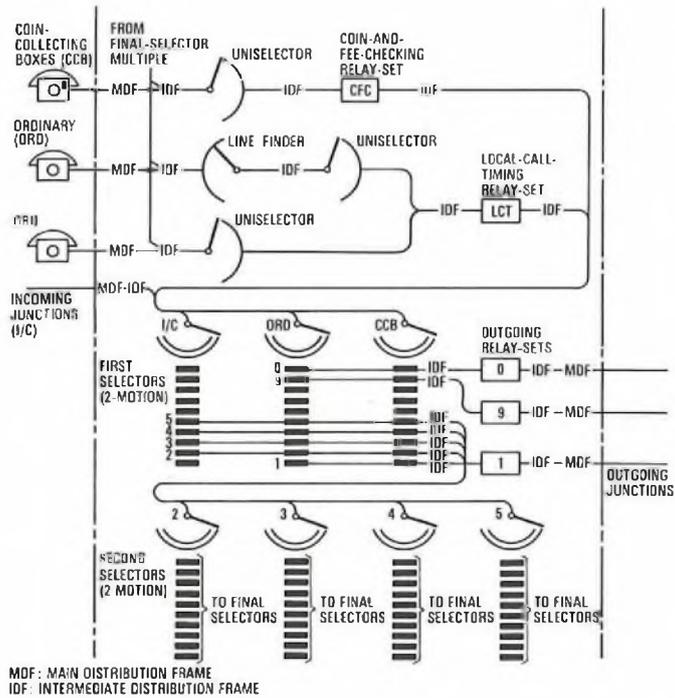
(*ii*) The arrangement of 200-outlet group selectors with 100-outlet final selectors tends to decrease the number of gradings because, as the group-selector availability is 20, there are less final selectors per pair of group-selector levels than for the alternative arrangement.

Q. 9. (*a*) *Draw simplified circuit diagrams to show the interconnexion of the principal components of a mains-driven battery-float power plant using*

(*i*) *a rotary machine, and*
(*ii*) *a rectifier set.*

(*b*) *Briefly describe the functions of the principal components of each system.*

(*c*) *List the advantages of rectifier sets over rotary machines.*

Q. 10. (*a*) *Design a grading suitable for giving access from a level of 100-outlet first group-selectors to 20 second group-selectors.*

(*b*) *Assuming 16 shelves of first selectors, draw the grading and show the formation of grading groups from the first-selector level.*

A. 10. (*a*) For 100-outlet first selectors, the availability, A, is 10 outlets/level. The level has access to 20 second selectors and, hence, the number of trunks, n, is 20. The number of grading groups, g, is given by the equation

$$g = \frac{2n}{A} = \frac{2 \times 20}{10} = 4.$$

The factors of 4 are 1, 2 and 4, so that there will be singles, pairs and full commons. Let a, b and c be the numbers of singles, pairs and full commons, respectively.

Then, $\qquad\qquad 4a + 2b + c = 20,$ $\qquad$ ..... (1)

and $\qquad\qquad a + b + c = 10.$ $\qquad$ ..... (2)

Subtracting equation (2) from equation (1) gives

$$3a + b = 10. \qquad\qquad ...... (3)$$

Assigning values to a, the value of b is calculated from equation (3) and the value of c from equation (2). The following table shows these values with their respective sums of successive differences.

a	1	2	3	4
b	7	4	1	Equations not satisfied
c	2	4	6	
Sum of successive differences	11	2	7	—

The smoothest grading is that having the least sum of successive differences. Therefore the grading is 2 singles, 4 pairs and 4 full commons.
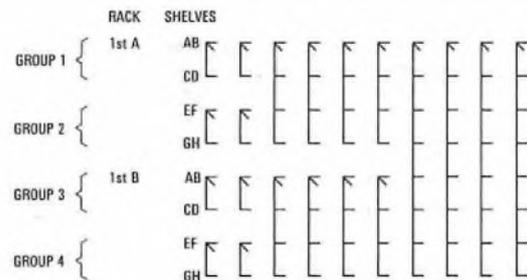
(*b*) The grading is shown in the sketch.
Since there are 16 shelves of first selectors, 4 shelves are allocated to each group, as illustrated in the sketch, there being 2 racks of selectors. The full commons involve wiring between the 2 racks, via tie cables.

LINE PLANT PRACTICE A, 1974

Students were expected to answer any 6 questions

Q. 1. *Give a detailed account of how a dangerous pole is replaced in an existing aerial cable route.*

A. 1. The replacement for a dangerous pole is erected as closely as possible to the existing pole, the excavation for the replacement being made immediately adjacent to it. If the decay is close to ground level, the existing pole is reinforced by stakes, driven into the ground and lashed to it. Alternatively, the pole may be reinforced by 2 ladders, placed against it and lashed together to form a braced structure.

After the stability of the existing pole has been secured, the new pole is erected, using the method most suited to the conditions. A pole-erection unit is used above the top of the jack. After erection, the new pole is made firm, and stayed if the existing pole was stayed.

The bolt supporting the D-bracket is removed to allow the aerial cable to be placed over the open end of the bracket, which is then firmly refixed to the pole. The aerial cable is secured to the D-bracket by a 1·8 m long piece of PVC-insulated binding wire, wrapped 3 times around the cable immediately adjacent to the bracket, and taken to the other side of the bracket to be made off in the opposite direction.

This operation is repeated on the opposite side of the bracket. When the cable is firmly secured to the new pole, the bindings on the D-bracket of the existing pole are removed.

The dangerous pole is then removed, either by means of the pole-erection unit or a pole-jack.

If a pole-jack is used, its base is supported on a stout piece of timber, so as to distribute the load. A chain is placed around the pole, near its base, and over the top of the jack. Operation of the jack causes the pole to be gradually lifted vertically out of the ground. The pole is temporarily stayed during the operation and, when out of the ground, is either lowered or permitted to fall, if the latter can be done safely.

An extensively decayed pole, in danger of collapse, may be lowered simply by pulling on temporary stays, or by sawing through the pole below ground level.

Q. 2. (*a*) *Describe how service can be provided to a subscriber's premises from a self-supporting aerial cable used for continuous distribution.*

(*b*) *Describe the work of making the cable joint.*
Illustrate your answers with sketches.

Q. 3. (a) Briefly describe a common type of self-supporting steel mast suitable for carrying a 35 m high aerial.

(b) With the aid of sketches, describe fully the foundation block for such a mast.

Q. 4. (a) With the aid of a sketch, describe the main features of an earthenware duct with a plastic-sleeve joint.

(b) What are the advantages of this type of duct?

(c) This duct is used in multiway formations. What formations are used for 3-way, 4-way and 5-way routes?

A. 4. (a) Sketch (a) shows general and cross-sectional views of an earthenware duct with a plastic sleeve. The duct, known as *uni-diameter*, is 1 m in length with an internal diameter of 92 mm, the bore being finished with a ceramic glaze. One end of the duct is fitted with a polyethylene sleeve during manufacture, using a bitumen compound; the other is a spigot, of reduced external diameter, made to fit into the polyethylene sleeve of an adjoining duct. The sleeve is 47 mm wide, and has an inner flange to provide a flush-fitting joint.
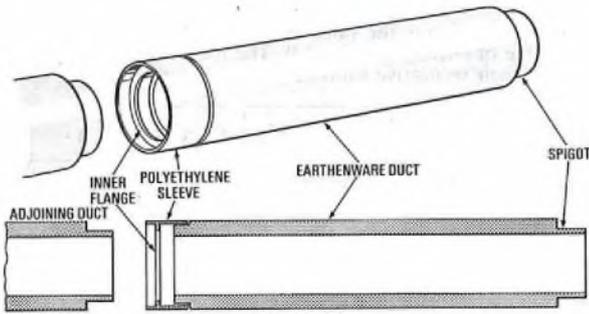
(a)

(b) The advantages of this type of duct are that

(i) cable laying is easier, due to the low friction offered by the flush joints and internal glazing,

(ii) it can be laid in short sections, so that the length of open trench is minimal,

(iii) it is not necessary to use jointing compound,

(iv) bends of a reasonably small radius can be negotiated,

(v) it can be moved if small changes in position are required during its life,

(vi) it is durable in the face of chemical attack and, therefore, has a long life,

(vii) its cost is low, and

(viii) it can be laid in multiway formations up to 18 ways and, consequently, there is no need to stock multiway ducts.
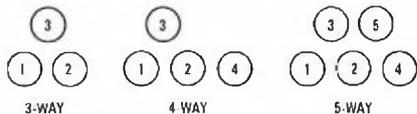
STAGGERED DUCT FORMATION
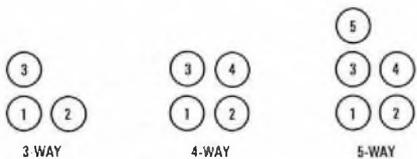
RECTANGULAR DUCT FORMATION

(b)

(c) For the main part of a duct route of 3, 4 or 5 ways, a staggered duct formation is normally used, as illustrated in sketch (b). This changes to a rectangular formation, also shown in sketch (b), before the route enters a jointing chamber, the change taking place over a distance of 5 m.

Q. 5. Describe the work involved in providing an underground 2-way duct route along the footway and across the carriageway of a suburban road.

Q. 6. (a) Describe the make-up of a 50-pair telephone distribution cable, suitable for burying directly in the ground.

(b) Describe, with the aid of a sketch, how a joint is made between the 50-pair cable and two 20-pair cables of similar make-up.

A. 6. (a) The distribution cable consists of 50 pairs of annealed-copper wire, 0·5 mm in diameter, each wire being insulated by an extruded cellular polyethylene covering, which is coloured for identification purposes. Two insulated conductors are uniformly twisted together to form a pair. The pairs are stranded in layers around a central core of 3 pairs, with 9 pairs in the first layer, 16 pairs in the second layer and 22 pairs in the outer layer, alternate layers having clockwise and counter-clockwise directions of lay. An open helical lapping of blue tape is applied to the central core and second layer, and an orange tape to the first layer. Two layers of insulating paper are lapped over, or applied longitudinally to, the outer layer, and the whole is sheathed with an extrusion of polyethylene, which contains carbon black to minimize deterioration of the polyethylene when exposed to sunlight. The air space in the core of the cable is completely filled with petroleum jelly.

The cable is also available with an armouring of galvanized mild-steel wires, stranded over the polyethylene sheath, with a further layer of natural polyethylene over the armouring.

(b) The following describes the method of making a joint between the 50-pair cable and two 20-pair cables, using the in-line taped-sleeve method.

The ends of the cables are cleaned with a dry rag, and arranged in such a position as to allow a 300 mm jointing gap. The tapered nozzles of 2 polyethylene collars are cut to accept the cables, and passed over the ends of the cables, ensuring a tight fit. A polyethylene sleeve is passed over the 50-pair cable and set away from the jointing position. The sheaths are removed from the cables as far as the jointing-gap mark, and the free ends of the cores are tied so as to preserve the cable formation while jointing is in progress.

The first pair of the appropriate 20-pair cable, and the first pair of the 50-pair cable, are selected from each side of the joint. Assuming that the hand-crimped method of jointing is to be used, the 4 wires are twisted together for one complete turn at a point about 30 mm from one end of the jointing gap. Succeeding pairs are similarly treated, and laid at 30 mm intervals across the jointing gap, as shown in sketch (a), so as to form regular banks of individual joints. The conductors are cropped about 25 mm from the point at which they are twisted. Each group of 4 wires is separated into A-wires and B-wires, and an insulated connector is passed over them until the ends of the wire butt against the end of the connector's inner sleeve, it being ensured that the wires are parallel. When all connectors have been fitted, they are crimped with hand-crimping pliers and folded down to form banks of connectors, which are tied in position, as shown in sketch (b). The 10 pairs left dead in the joint are neatly tied to the core of the joint, and the whole is wrapped with insulating paper.
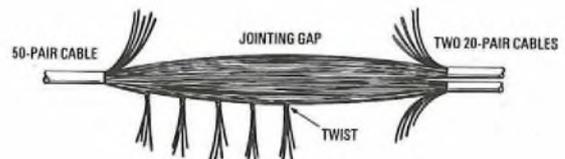
(a)
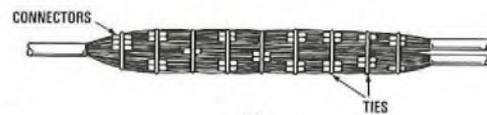
(b)
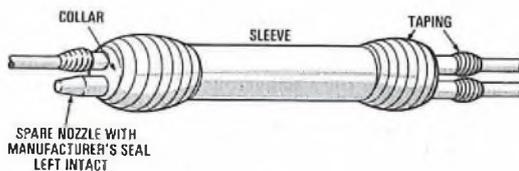
(c)

The sleeve is placed centrally over the completed joint, and the collars are moved along the cables and inserted into the sleeve so that the shoulder of each collar fits closely into the end of the sleeve. To ensure that there are no longitudinal score-marks on the cable and jointing parts, which may allow water to seep into the joint, the surfaces of the cables and sleeve are smoothed with glass paper for a distance of 50 mm each side of the collar before being taped. Self-amalgamating tape is then wrapped around each end of the joint, the tape being applied under a light tension and with an overlap on adjacent lappings of 50%. Over this, an adhesive plastic tape is also applied, again with an overlap of 50%, to compress the self-

amalgamating tape to ensure a good seal. The completed joint is illustrated in sketch (*c*).

Q. 7. (*a*) *With the aid of sketches, describe the various types of support used to carry cables and joints in jointing chambers.*
(*b*) *Explain how a large plastic-sheathed cable is secured in position.*

A. 7. (*a*) The various types of support are described below.

Cable-Bearers and Brackets
There are 2 types of cable-bearer: one that is fixed to the wall of a jointing chamber, and the other to the floor. The wall-mounted bearer, shown in sketch (*a*), is made from light pressed-steel channel section, and has a galvanized finish. The floor-mounted bearer, shown in sketch (*b*), is cut from H-section rolled-steel joist, and is fixed to the floor (and roof, if required) by rolled-steel angle-brackets. The brackets which carry the cables are made of grey cast iron, and are secured in position by pins of high-tensile steel wire, bent to form self-locking pivots.

(*a*)

(*b*)

(*c*)
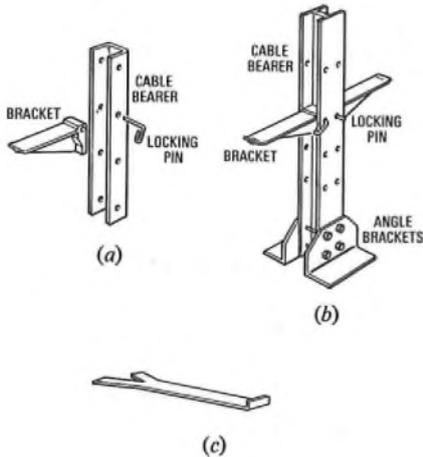
Joint-Box Brackets
Supporting brackets for cables and joints in small footway joint boxes, constructed of brick or concrete, are made of wrought iron or mild steel. Sketch (*c*) shows such a bracket. The forked end is built into the side wall of the joint box.

Cable Supports
A cable support, illustrated in sketch (*d*), is a galvanized mild-steel bar, designed to give support to cables and joints. The ends of the support rest on cable-bearer brackets or in the mouths of ducts.

(*d*)

(*e*)

Asbestos-Cement Troughing
Square-section asbestos-cement troughing, shown in sketch (*e*), is used for the lengthwise support of small cables and joints. It rests on cable-bearer brackets.

(*b*) Restrainers, made of lengths of slotted 25 mm angle-section mild steel, provided with stops and fixing holes, are used in conjunction with binding strips and clips for securing large plastic-sheathed cables. Left-hand and right-hand restrainers are located each side of a cable-bearer bracket, and are fixed in position by means of the locking pin.

(*f*)
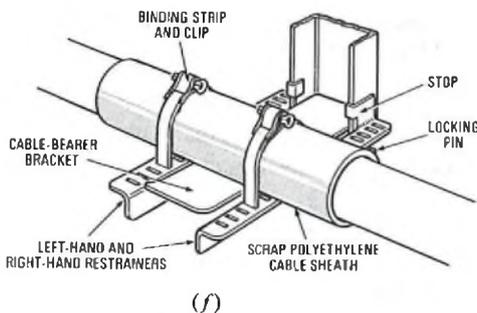
The restrainers enable the cable to be bound on each side of the bracket. The clip and strip assemblies are threaded through suitable slots in the restrainer, a piece of scrap polyethylene cable-sheath being placed over the cable before the bindings are applied. After tightening the bindings, the clips and strips are painted with black paint. The completed securing arrangements are shown in sketch (*f*).

Q. 8. (*a*) *In what circumstances would a duct be laid by means of a thrust-borer?*
(*b*) *With the aid of a sketch, show how the carriage of a thrust-borer is thrust forward and moved back.*

A. 8. (*a*) A duct may be laid by means of a thrust-borer when the soil is of such a nature that it can be compressed; for example, loam, clay or soft chalk—in fact, any soil which can be dug with a spade. Soils containing boulders, stones and tree roots make the operation uncertain and hazardous. Special difficulties may arise at road crossings and similar situations, due to the presence of underground pipes and cables.
Thrust-boring is usually employed when the surface paving is very expensive to reinstate; for example, for

(*i*) a road crossing,
(*ii*) a short length in the footway under a driveway,
(*iii*) a level-crossing or railway embankment (but not for an electrified railway), or
(*iv*) a point in the footway or carriageway where material is stacked which would be very expensive to remove.
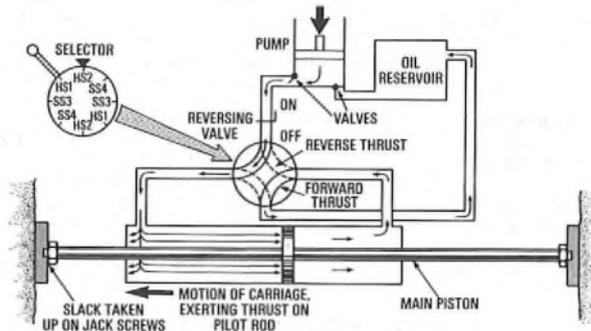
(*b*) The sketch shows how the direction of thrust in a thrust-borer is reversed by means of a reversing valve in the hydraulic system. Appropriately setting the valve causes the hydraulic pressure to be applied firstly to one side of the main piston and then to the other.

Q. 9. *Describe the precautions taken to prevent damage to overhead telephone lines which cross or approach high-voltage power lines.*

A. 9. (*a*) The normal minimum horizontal separation between telephone wires and the conductors of a high-voltage power line is a distance equal to the height of the power conductors at the point of attachment to their supports, or 1·5 times the height of the telephone wires, whichever is the greater. Heights on uneven ground are measured from the lower ground level. In circumstances where it is difficult to achieve these distances, a reduced separation is allowed in accordance with the following 2 conditions.

(*i*) Should the power line overturn in the direction of the telephone line, no part of an unearthed power conductor must pass within 1·8 m of the telephone line.
(*ii*) Should the telephone line overturn in the direction of the power line, no part of the telephone line must pass within 3·65 m of any unearthed power conductor.

For these reduced separations to apply, the difference in ground level between the 2 sets of plant must be less than 1·8 m, and both sets of plant are considered to pivot at the ground line or, if the pole is in soft ground, at the butt of the pole. These clearances also apply to the 25 kV overhead contact wire of electrified railways.

(*b*) If the above clearances cannot be achieved, the precautions taken vary with the voltage of the power line, as detailed below.

Insulated Telephone Line
Blue PVC-insulated cadmium-copper wire may be used for the telephone wires if the power-line voltage does not exceed 33 kV. Polyethlene-insulated aerial cable may be used if the power-line voltage does not exceed 11 kV.
At crossings, the power conductors must be above the telephone wires, with a minimum separation of 1·83 m at 50°C. The separation between the lowest power conductor and the nearest telephone pole,

23

when viewed in plan, should not be less than 1·82 m or, alternatively, the vertical distance between the lowest power conductor and the top of the telephone pole should not be less than 2·7 m at 50°C. The insulated telephone wires should extend on each side of the power line for the normal minimum horizontal separating distance. Insulated wire should have no exposed joints.

Where telephone wires, insulated as described above, approach high-voltage power conductors up to 11 kV, the horizontal separating distance may be reduced to not less than 3·66 m. For power conductors working at a voltage of up to 33 kV, the horizontal separating distance may be reduced to not less than 3·66 m only where blue PVC-insulated wire is used.

Insulated Power Lines

PVC-insulated power conductors may be used, provided that they are above the telephone wires, and the power-line voltage does not exceed 11 kV.

The insulated power conductors may be supported by

(*i*) a section pole, which is a support having conductors terminating on each side of it, connected through by bow connexions, the section pole being at least 4·6 m from the nearest part of the telephone line,

(*ii*) an intermediate pole, which is a pole where the conductors are only bound to the insulators, the minimum clearance from the nearest part of the telephone line being 1·22 m,

(*iii*) a terminal pole, which is a pole at the end of an overhead route, the normal minimum horizontal separation being required, or

(*iv*) a transformer pole, the minimum separation being the same as that required for a terminal pole.

The power conductors should be insulated for a distance, on each side of the telephone line, equal to the minimum horizontal separation, except that, where they are attached to a section pole, and the angle of the crossing is greater than 60°, the insulation need not extend beyond the section pole.

(*c*) If none of the precautions detailed in parts (*a*) and (*b*) can be achieved, or are not applicable, such as in the case of a power-line voltage exceeding 33 kV, one set of plant should be placed underground. Because of the expense involved, this would normally be the telephone line. To provide adequate separation, the underground portion of the telephone line should extend on each side of the crossing so that the conditions of separation given in part (*a*) are fulfilled. However, the relaxations detailed in parts (*a*) (*i*) and (*ii*) do not apply.

(*d*) A cradle-guard is a permissible form of guarding, and consists of wires slung beneath the power conductors, with cross-lacings forming a large mesh. With the advent of insulated overhead power conductors and polyethylene-sheathed aerial telephone cables, a cradle-guard is rarely necessary, and its use is avoided if possible. A cradle-guard is not used if the power-line voltage exceeds 33 kV.

Q. **10.** (*a*) *What is a Varley test used for on telephone cables?*
(*b*) *Draw a circuit diagram showing how the connexions for Varley-test apparatus are made.*

(*c*) *A Varley test on a cable, 10 km long, uses the same gauge conductor for the good wire as for the faulty wire, and equal ratio arms. If the loop reading is 550 Ω and the Varley reading is 220 Ω, calculate the distance to the fault.*

A. 10. (*a*) The Varley test is used at an exchange or repeater station to determine the distance to an earth or contact fault on a wire in a cable. The Varley test can locate the fault to the cable length containing it. The more accurate Murray test is then used to locate the exact position of the fault.

(*b*) The sketch shows how the Varley-test apparatus is connected.
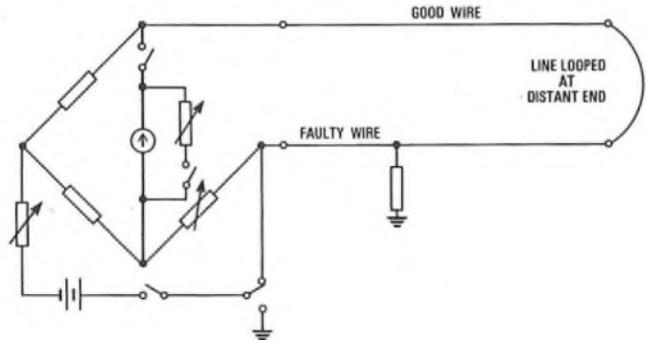
(*c*) The resistance to the fault

$$= \frac{\text{(loop reading)} - \text{(Varley reading)}}{2},$$

$$= \frac{550 - 220}{2} = 165\ \Omega.$$

The resistance of the conductor per kilometre

$$= \frac{\text{loop resistance}}{2 \times \text{(cable length)}},$$

$$= \frac{550}{2 \times 10} = 27 \cdot 5\ \Omega/\text{km}.$$

Hence, the distance to the fault

$$= \frac{\text{resistance to the fault}}{\text{resistance of conductor/km}},$$

$$= \frac{165}{27 \cdot 5} = \underline{6\ \text{km}.}$$

MATHEMATICS B, 1974

Students were expected to answer any 6 questions

Q. **1.** (*a*) *An insurance company annually secured new business, y, in millions of pounds, as shown for 3 years in the table below.*

Year	1962	1967	1972
y	28	51	87

Assuming the formula $y = a + bx + cx^2$, *where x represents the number of years after 1962,*

(*i*) *calculate a, b and c, and*
(*ii*) *estimate the firm's new business in 1974.*

(*b*) *Solve the simultaneous equations:*

$$p^2 - q^2 = 3 \cdot 77,$$
$$2p - 3q = 1 \cdot 8.$$

A. 1. (*a*) Since x represents the number of years after 1962, the variables x and y are shown in the following table.

Year	1962	1967	1972
x	0	5	10
y	28	51	87

(*i*) On substituting the data into the formula $y = a + bx + cx^2$, the following equations are obtained:

$$28 = a, \qquad \dots\dots (1)$$
$$51 = a + 5b + 25c, \qquad \dots\dots (2)$$
$$87 = a + 10b + 100c. \qquad \dots\dots (3)$$

Substituting for $a = 28$ in equations (2) and (3) gives

$$51 = 28 + 5b + 25c,$$
$$\text{or } 5b + 25c = 23, \qquad \dots\dots (4)$$
$$\text{and } 87 = 28 + 10b + 100c,$$
$$\text{or } 10b + 100c = 59. \qquad \dots\dots (5)$$

Multiplying equation (4) by 2 gives

$$10b + 50c = 46. \qquad \ldots\ldots (6)$$

Subtracting equation (6) from equation (5) gives

$$50c = 13.$$

$$\therefore c = \frac{13}{50} = 0\cdot 26.$$

Substituting for c in equation (5) gives

$$10b + 26 = 59.$$

$$\therefore 10b = 33.$$

$$\therefore b = 3\cdot 3.$$

Hence, $a = 28$, $b = 3\cdot 3$ and $c = 0\cdot 26$.

(*ii*) The year 1974 is 12 years after 1962 and, hence, substituting for $x = 12$ in the formula $y = a + bx + cx^2$ gives

$$y = 28 + 3\cdot 3 \times 12 + 0\cdot 26 \times 12^2,$$

$$= 28 + 39\cdot 6 + 37\cdot 44,$$

$$= 105\cdot 04.$$

Therefore, the firm's new business in 1974 = £105 040 000.

(*b*) Now, $p^2 - q^2 = 3\cdot 77$, $\qquad \ldots\ldots (1)$

and $2p - 3q = 1\cdot 8$. $\qquad \ldots\ldots (2)$

From equation (2),

$$3q = 2p - 1\cdot 8,$$

$$\text{or } q = \frac{2}{3}p - 0\cdot 6. \qquad \ldots\ldots (3)$$

Substituting for q in equation (1) gives

$$p^2 - \left(\frac{2}{3}p - 0\cdot 6\right)^2 = 3\cdot 77.$$

$$\therefore p^2 - \frac{4}{9}p^2 + 0\cdot 8p - 0\cdot 36 = 3\cdot 77.$$

$$\therefore \frac{5}{9}p^2 + 0\cdot 8p - 4\cdot 13 = 0.$$

$$\therefore 5p^2 + 7\cdot 2p - 37\cdot 17 = 0.$$

$$\therefore p = \frac{-7\cdot 2 \pm \sqrt{\{7\cdot 2^2 - 4\times 5 \times(-37\cdot 17)\}}}{2\times 5},$$

$$= \frac{-7\cdot 2 \pm 28\cdot 2}{10},$$

$$= 2\cdot 1 \text{ or } -3\cdot 54.$$

Substituting for p in equation (3) gives

$$q = \frac{2}{3} \times 2\cdot 1 - 0\cdot 6 = 0\cdot 8,$$

or $\qquad q = \frac{2}{3} \times (-3\cdot 54) - 0\cdot 6 = -2\cdot 96.$

Hence, either $p = 2\cdot 1$ and $q = 0\cdot 8$, or $p = -3\cdot 54$ and $q = -2\cdot 96$.

Note: The answers should be checked in each of the original equations.

Q. 2. *The formula* $v = 80(1 - e^{-t/CR})$ *gives the voltage across a capacitor, C farads, charged through a resistance, R ohms, at an instant, t seconds.*
   Making use of tables,
   (*a*) *calculate v when* $t = \frac{1}{2}CR$,
   (*b*) *sketch the graph of v/t for positive values of t, carefully marking the axes and indicating the significance of the number 80, and*
   (*c*) *express t in terms of v, C, R and e and, hence, calculate the time taken for the voltage to rise from zero to 8 V, given that* $C = 80\ \mu F$ *and* $R = 50\ k\Omega$.

A. 2. $\qquad\qquad v = 80(1 - e^{-t/CR}).$

(*a*) Substituting for $t = \frac{1}{2}CR$ gives

$$v = 80(1 - e^{-0\cdot 5}),$$

$$= 80(1 - 0\cdot 6065) \text{ from a table of} \atop \qquad\qquad\qquad \text{exponential functions,}$$

$$= 31\cdot 48\text{ V}.$$

(*b*) The graph of v/t, shown in the sketch, is constructed from the following table of values, derived from a table of exponential functions for values of t up to $t = 4CR$, in a manner similar to that used in part (*a*).

t (s)	0	$0\cdot 5CR$	CR	$2CR$	$3CR$	$4CR$
v (V)	0	$31\cdot 48$	$50\cdot 57$	$69\cdot 17$	$76\cdot 02$	$78\cdot 53$

As t assumes values greater than $4CR$, $e^{-t/CR}$ becomes correspondingly smaller, and eventually approximates to zero. Thus, v approaches 80 V, as shown in the sketch. At $t = 5CR$, the difference between the value of v and its final value of 80 V is less than 1%.
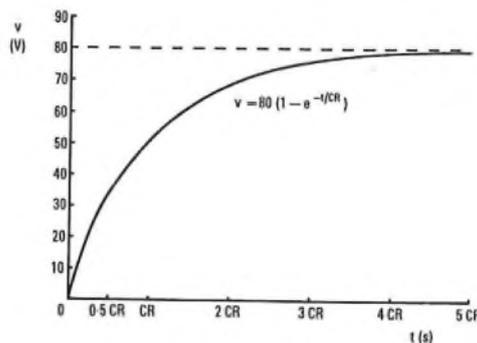
Note: As the question asks only for a sketch of the graph of v/t, it is not necessary to plot the curve accurately. A rough sketch of the general exponential charging curve would be sufficient, provided that the axes were labelled and the curve was shown to be asymptotic to the value $v = 80$ V.

(*c*) Now, $v = 80(1 - e^{-t/CR})$.

$$\therefore \frac{v}{80} = 1 - e^{-t/CR}.$$

$$\therefore e^{-t/CR} = 1 - \frac{v}{80},$$

$$= \frac{80 - v}{80}.$$

Taking logarithms to base e,

$$-\frac{t}{CR} = \log_e\left(\frac{80 - v}{80}\right).$$

$$\therefore t = -CR\log_e\left(\frac{80 - v}{80}\right).$$

Substituting the given values for $v = 8$ V,

$$t = -80 \times 10^{-6} \times 50 \times 10^3 \times \log_e\left(\frac{80 - 8}{80}\right),$$

$$= -4\log_e 0\cdot 9,$$

$$= -4 \times (\bar{1}\cdot 8946) \text{ from a table of Napierian logarithms,}$$

$$= -4 \times (-0\cdot 1054)\text{ s},$$

$$= 421\cdot 6\text{ ms}.$$

Q. 3. *The formula* $P = \frac{a}{v} + bv^2$ *is thought to give the daily operating cost, P pounds, of a ship travelling at v knots. Values of P and v are given in the following table.*

P	2300	1800	1600	2100
v	2	4	5	6

Plot suitable variables to obtain a straight-line graph to test this formula. From it, obtain values for the constants a and b.

A. 3. Now, $\qquad\qquad P = \frac{a}{v} + bv^2.$

Multiplying throughout by v gives

$$Pv = a + bv^3. \qquad \ldots\ldots (1)$$

25

Since P and v are variables and a and b are constants, equation (1) is of the linear type, $y = mx + c$. Hence, if Pv is plotted against v^3, a straight-line graph should be obtained. The required values of Pv and v^3 are given by the following table.

P	2300	1800	1600	2100
v	2	4	5	6
Pv	4600	7200	8000	12 600
v^3	8	64	125	216

The graph of Pv/v^3 is shown in the sketch, from which it can be seen that the 4 plotted points lie reasonably close to a straight line. The approximation to the linear law is not very good, but the worst departures from it occur in the middle of the range of speed values, where the operating costs are lowest. Consequently, these departures are of small significance.
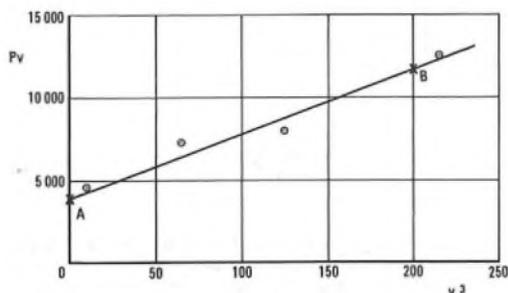
When $v^3 = 0$, $Pv = a$. Hence, the value of a is given by the intercept of the graph and the Pv axis.

From the graph, $a = 4000$.

The gradient of the line gives the value of b, and is obtained from points A and B, which are widely separated and actually lie on the curve. From the co-ordinates of points A and B,

$$b = \frac{11\ 750 - 4000}{200},$$

$$\simeq 39.$$

Q. 4. (*a*) *The first 2 terms of a series are 16 and 12·8. Calculate the sum of 10 terms of the series if it is*

(*i*) *an arithmetical progression, and*
(*ii*) *a geometrical progression.*

(*b*) *The population of a new town is planned to increase by 12% each year (compound-interest law) over the first 10 years of its life. If the population after 3 years is 2400, calculate the expected size of the town's population at the end of the 10-year period.*

A. 4. (*a*) (*i*) If a and l respectively denote the first and last terms of an arithmetical progression, d is the difference between successive terms, and n is the number of terms, then

$$l = a + (n - 1)d,$$

and $\qquad S_n = \dfrac{n}{2}(a + l)$, where S_n is the sum of n terms.

Since 16 and 12·8 are the first 2 terms of the series,

$$d = 12 \cdot 8 - 16 = -3 \cdot 2.$$

When $n = 10$, $l = 16 + (10 - 1) \times (-3 \cdot 2)$,

$$= -12 \cdot 8.$$

$$\therefore S_{10} = \frac{10}{2}(16 - 12 \cdot 8),$$

$$= 16.$$

(*ii*) For a geometrical progression, if r denotes the common ratio, S_n is given by the formula

$$S_n = \frac{a(1 - r^n)}{1 - r}.$$

Now, $\quad r = \dfrac{12 \cdot 8}{16} = 0 \cdot 8.$

$$\therefore S_{10} = \frac{16(1 - 0 \cdot 8^{10})}{1 - 0 \cdot 8},$$

$$= 80(1 - 0 \cdot 1074),$$

$$= 71 \cdot 41.$$

(*b*) Let P be the population at the beginning of the first year. At the end of the first year,

$$P_1 = P + 0 \cdot 12P = 1 \cdot 12P,$$

and, at the end of the second year,

$$P_2 = 1 \cdot 12P + 0 \cdot 12 \times 1 \cdot 12P = 1 \cdot 12^2P.$$

Similarly, at the end of n years,

$$P_n = 1 \cdot 12^nP.$$

Hence, after 3 years,

$$2400 = 1 \cdot 12^3P.$$

$$\therefore P = \frac{2400}{1 \cdot 12^3}.$$

Therefore, at the end of 10 years, the expected size of the town's population is given by

$$P_{10} = 1 \cdot 12^{10} \times \frac{2400}{1 \cdot 12^3},$$

$$= 1 \cdot 12^7 \times 2400,$$

$$= 5306.$$

Q. 5. (*a*) *Show that $\log_e N = k \log_{10} N$, and evaluate the constant k using logarithms to base 10 only.*

(*b*) *A voltage gain may be expressed either as $20 \log_{10}(V_2/V_1)$ decibels, or as $\log_e(V_2/V_1)$ nepers. Express a voltage gain of 5·4 nepers in decibels.*

(*c*) *Using a table of Napierian (natural) logarithms, plot the graph of $y = \log_e x$ from $x = 0 \cdot 5$ to $x = 3$. Use the graph to verify that, for the value $x = 2$, $\dfrac{dy}{dx} = \dfrac{1}{x}$ for this function.*

A. 5. (*a*) Let $\log_e N = p$.

Then, $N = e^p$.

$$\therefore \log_{10} N = \log_{10}(e^p),$$

$$= p \log_{10} e.$$

$$\therefore p = \frac{\log_{10} N}{\log_{10} e}.$$

$$\therefore \log_e N = k \log_{10} N, \qquad\qquad \text{QED,}$$

where k is a constant equal to $\dfrac{1}{\log_{10} e}$,

$$= \frac{1}{\log_{10} 2 \cdot 7183},$$

$$= \frac{1}{0 \cdot 4343} = 2 \cdot 303.$$

(*b*) From part (*a*),

$$\log_{10} N = \frac{\log_e N}{k} = 0 \cdot 4343 \log_e N. \qquad \ldots\ldots (1)$$

Voltage gain in decibels

$$= 20 \log_{10}\left(\frac{V_2}{V_1}\right),$$

$$= 20 \times 0 \cdot 4343 \times \log_e\left(\frac{V_2}{V_1}\right), \text{ from equation (1),}$$

$$= 8 \cdot 686 \log_e\left(\frac{V_2}{V_1}\right),$$

$$= 8 \cdot 686 \times (\text{voltage gain in nepers}),$$

$$= 8 \cdot 686 \times 5 \cdot 4 = 46 \cdot 9 \text{ dB}.$$

(*c*) The sketch shows the graph of $y = \log_e x$, plotted from the

following table of values, which are derived from a table of Napierian logarithms.

x	0·5	1·0	1·5	2·0	2·5	3
y	−0·6931	0	0·4055	0·6931	0·9163	1·0986
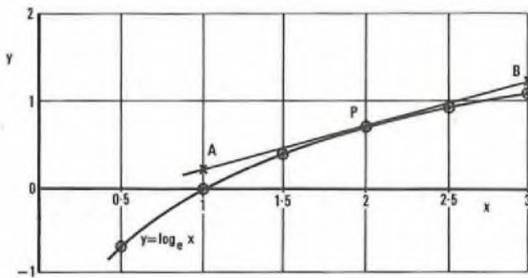
P is the point on the graph at $x = 2$. The gradient of the tangent, AB, to the curve, drawn through P, gives the value of dy/dx at $x = 2$. From the co-ordinates of points A and B, the gradient of the tangent at P

$$= \frac{1·2 - 0·2}{3 - 1},$$

$$= \frac{1}{2} = \frac{1}{x}. \qquad \text{QED.}$$

Q. 6. (a) *Sketch the graph of the current* $i = 8 \cos (\omega t - \pi/6)$ *mA, where* ω *is in radians/second, from* $t = 0$ *to* $t = 4\pi/\omega$ *seconds.*
(b) *If* $\omega = 5000$ *rad/s, calculate this current when* $t = 24 \times 10^{-6}$ *s.*
(c) *Calculate the time interval after* $t = 0$ *when the current first reaches it peak value. Express the answer in milliseconds to 3 significant figures.*

A. 6. (a) The graph of $i = 8 \cos (\omega t - \pi/6)$ is shown in the sketch. The range $t = 0$ to $t = 4\pi/\omega$ seconds is an angular range of $0-720°$, or 4π rad. There are 2 complete cycles of the cosine function, as shown. The peak value, of $i = 8$ mA, is displaced by $\pi/6\omega$ seconds, or $30°$, from $t = 0$ s.
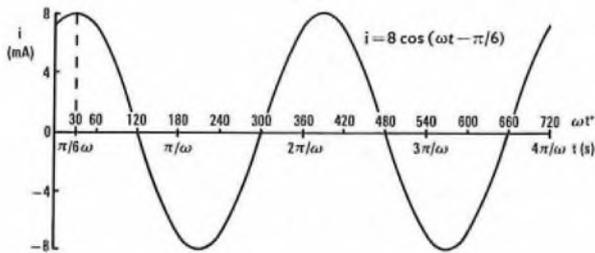
(b) Substituting the given values,

$$i = 8 \cos (5000 \times 24 \times 10^{-6} - \pi/6),$$

$$= 8 \cos (-0·4036 \text{ rad}),$$

$$= 8 \cos (-23° \, 7'),$$

$$= 8 \times 0·9197 = 7·357 \text{ mA}.$$

(c) The current first reaches the peak value when $\cos (\omega t - \pi/6) = 1$; that is, when $\omega t - \pi/6 = 0$.

Hence, $\omega t = \dfrac{\pi}{6}$.

$$\therefore \ t = \frac{\pi}{5000 \times 6},$$

$$= 1·047 \times 10^{-4} \text{ s},$$

$$= 0·105 \text{ ms, correct to 3 significant figures.}$$

Q. 7. (a) *State the expansions for* $\sin (A + B)$ *and* $\cos (A + B)$ *and, hence,*

(i) *evaluate* $\cos 105°$ *to 3 significant figures without using trigonometrical tables, and*
(ii) *find, to the nearest degree, the 2 acute angles which differ by* $10°$, *and whose sines differ by* $0·1$.
Assume the larger angle to be $(\theta + 5)°$.

(b) *Calculate the smallest angle of the triangle whose sides are 27, 20 and 16 mm.*

A. 7. (a) $\sin (A + B) = \underline{\sin A \cos B + \cos A \sin B}$.

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

(i) $\cos 105° = \cos (60° + 45°)$,

$$= \cos 60° \times \cos 45° - \sin 60° \times \sin 45°,$$

$$= \frac{1}{2} \times \frac{1}{\sqrt{2}} - \frac{\sqrt{3}}{2} \times \frac{1}{\sqrt{2}},$$

$$= -0·259, \text{ to 3 significant figures.}$$

(ii) If the larger angle is $(\theta + 5)°$, then the smaller angle is $(\theta - 5)°$. Hence, $\sin (\theta + 5)° - \sin (\theta - 5)° = 0·1$, since the larger angle has the larger sine.

$$\therefore \ \sin \theta \cos 5° + \cos \theta \sin 5° - \sin \theta \cos 5° + \cos \theta \sin 5° = 0·1.$$

$$\therefore \ 2 \cos \theta \sin 5° = 0·1,$$

$$\therefore \ \cos \theta = \frac{0·1}{2 \sin 5°},$$

$$= \frac{0·1}{2 \times 0·0872},$$

$$= 0·5734.$$

$$\therefore \ \theta = 55°.$$

Hence, the required angles are $\underline{50°}$ and $\underline{60°}$.

(b) In a triangle, the smallest angle is opposite the smallest side. Hence, the required angle is opposite the 16 mm side.
From the cosine rule,

$$\cos C = \frac{a^2 + b^2 - c^2}{2ab},$$

where a, b and c represent the sides of the triangle, and C the angle opposite side c.

$$\therefore \ \cos C = \frac{27^2 + 20^2 - 16^2}{2 \times 27 \times 20},$$

$$= 0·8083.$$

$$\therefore \ \underline{\angle C = 36° \, 4'}.$$

Q. 8. (a) *If* $y = 6x^2 - x^3$, *derive from first principles an expression for* dy/dx.
(b) *Using part (a), find the values of* y *for which* $dy/dx = 0$.
(c) *Sketch the graph of* $y = 6x^2 - x^3$, *displaying all its significant features.*

A. 8. (a) $\qquad\qquad y = 6x^2 - x^3.$

Let x increase by a small amount, δx, and let the corresponding change in y be δy.

$$\therefore \ y + \delta y = 6(x + \delta x)^2 - (x + \delta x)^3.$$

$$\therefore \ \delta y = 6(x + \delta x)^2 - (x + \delta x)^3 - 6x^2 + x^3,$$
$$\text{since } y = 6x^2 - x^3.$$

$$\therefore \ \delta y = 6x^2 + 12x\delta x + 6(\delta x)^2 - x^3 - 3x^2\delta x$$
$$-3x(\delta x)^2 - (\delta x)^3 - 6x^2 + x^3,$$

$$= 12x\delta x + 6(\delta x)^2 - 3x^2\delta x - 3x(\delta x)^2 - (\delta x)^3.$$

$$\therefore \ \frac{\delta y}{\delta x} = 12x - 3x^2 + 6\delta x - 3x\delta x - (\delta x)^2.$$

$$\therefore \ \frac{dy}{dx} = \underset{\delta x \to 0}{\text{limit}} \frac{\delta y}{\delta x},$$

$$= \underline{12x - 3x^2}.$$

(b) From part (a), when $dy/dx = 0$,

$$12x = 3x^2.$$

$$\therefore \ x = 0 \text{ or } 4.$$

27

When $x = 0$, $y = 0$, and when $x = 4$, $y = 32$. Thus, the values of y for which $dy/dx = 0$ are 0 and 32.

(*c*) The graph is sketched from the following table of values.

x	-4	-2	0	2	4	6	8
x^2	16	4	0	4	16	36	64
$6x^2$	96	24	0	24	96	216	384
x^3	-64	-8	0	8	64	216	512
$y = 6x^2 - x^3$	160	32	0	16	32	0	-128
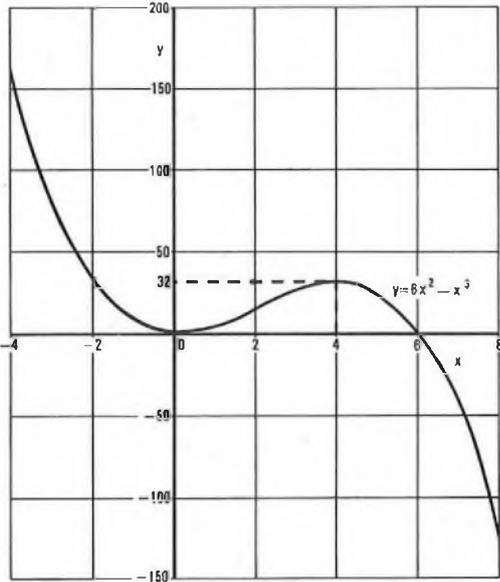
The graph is shown in the sketch, from which it is seen that tangents to the curve at $x = 0$ and $x = 4$ are parallel to the x-axis; that is, $dy/dx = 0$ at these points. A maximum value of $y = 32$ occurs at $x = 4$, and a minimum of $y = 0$ at $x = 0$. Below $x = -4$, the curve ascends rapidly, and above $x = 8$, it descends rapidly.

Q. 9. (*a*) *Find the equation of the curve which passes through the point $x = 2$, $y = 4$, such that $dy/dx = 2x - 7$.*

(*b*) *Evaluate $\int_{-1}^{3}(4x - x^2)\,dx$, and deduce the mean value of the function $4x - x^2$ from $x = -1$ to $x = +3$.*

A. 9. (*a*) If $\dfrac{dy}{dx} = 2x - 7$,

then $y = \int (2x - 7)\,dx$,

$= x^2 - 7x + c$, where c is a constant.

At $x = 2$, $y = 4$. Substituting these co-ordinates in the above equation gives

$$4 = 4 - 14 + c.$$

$$\therefore \ c = 14.$$

Hence, the equation of the curve is $y = x^2 - 7x + 14$.

(*b*) $\displaystyle\int_{-1}^{3}(4x - x^2)\,dx = \left[2x^2 - \frac{x^3}{3}\right]_{-1}^{3}$;

$= (18 - 9) - (2 + \tfrac{1}{3})$,

$= 6\tfrac{2}{3}$.

The mean value of $y = 4x - x^2$ between $x = -1$ and $x = 3$

$$= \frac{\displaystyle\int_{-1}^{3}(4x - x^2)\,dx}{3 - (-1)},$$

$$= \frac{6\tfrac{2}{3}}{4},$$

$$= 1\tfrac{2}{3}.$$

Q. 10. (*a*) *If $z = 3 - 5j$, express, in the form $a + jb$,*

(*i*) *$(z + 3)(4 - z)$, and*

(*ii*) *$\dfrac{1 - z}{z + 4j}$.*

(*b*) *When $\omega = 5000$, $R = 40$, and $C = 2\cdot4 \times 10^{-6}$, calculate, in the form $a + jb$, the value of $Z = 1/\{(1/R) + j\omega C\}$. Show that the magnitude (modulus) of Z is approximately 36.*

A. 10. (*a*) (*i*) If $z = 3 - 5j$, then

$$z + 3 = 3 - 5j + 3 = 6 - 5j,$$

$$\text{and } 4 - z = 4 - (3 - 5j) = 1 + 5j.$$

$$\therefore \ (z + 3)(4 - z) = (6 - 5j)(1 + 5j),$$

$$= 6 + 30j - 5j + 25,$$

$$= 31 + j25.$$

(*ii*) Similarly,

$$1 - z = 1 - (3 - 5j) = -2 + 5j,$$

$$\text{and } z + 4j = 3 - 5j + 4j = 3 - j.$$

$$\therefore \ \frac{1 - z}{z + 4j} = \frac{-2 + 5j}{3 - j},$$

$$= \frac{(-2 + 5j)(3 + j)}{(3 - j)(3 + j)},$$

$$= \frac{-6 - 2j + 15j - 5}{9 + 1},$$

$$= \frac{-11 + 13j}{10},$$

$$= -1\cdot1 + j1\cdot3.$$

(*b*) $$Z = \frac{1}{\dfrac{1}{R} + j\omega C},$$

$$= \frac{R}{1 + j\omega CR}.$$

Substituting the given values,

$$Z = \frac{40}{1 + j \times 5000 \times 2\cdot4 \times 10^{-6} \times 40},$$

$$= \frac{40}{1 + j0\cdot48},$$

$$= \frac{40(1 - j0\cdot48)}{1\cdot2304},$$

$$= 32\cdot51 - j15\cdot61.$$

Hence, $$Z \simeq 32 - j16.$$

$$\therefore \ |Z| \simeq \sqrt{(32^2 + 16^2)},$$

$$= 16 \times \sqrt{(2^2 + 1^2)},$$

$$= 16 \times \sqrt{5},$$

$$\simeq 36. \qquad\qquad \text{QED}$$

Students were expected to answer any 6 questions

Q. 1. (a) *By means of a table, compare and contrast the operation of modern digital and analog computers.*

(b) *Write down, and briefly discuss, 2 typical applications that would be run on each of the following:*

(i) *a digital computer, and*

(ii) *an analog computer.*

A. 1. (a) The operation of digital and analog computers is compared and contrasted in the following table.

Analog Computer	Digital Computer
Number Representation Numbers are represented by the dimension of a continuous physical variable, usually voltage. Inputs and outputs are time-varying waveforms.	Numbers are represented by a series of discrete voltages, each having one of 2 values. Thus, numbers are stored in binary form.
Arithmetic The basic unit is an operational amplifier with a high gain, of the order of 10^6. All arithmetical operations are performed by using this amplifier with feedback and embedded networks which specify the operation to be performed.	A number of bistable circuits make up a register in which a number is held. Arithmetic is performed in an arithmetic unit, by manipulation of the numbers in the registers, using logical processes.
Storage There is no storage of data.	Large amounts of data can be stored in a magnetic-core store. There is also a backing store, which gives the computer access to very large quantities of data.
Operation A mathematical model of the task to be performed is worked out, so that the elements of the computer can be patched together to implement the model. The computing units are able to respond immediately to changes in the input variables, and the connexion of the various units in a particular fashion enables the computer to carry out rapidly very complex arithmetical functions.	The task to be performed is represented by a sequence of instructions called the program. The program is stored in the computer's memory in the same form as the data. Each instruction is obeyed sequentially, and specifies the function to be performed. The instruction is decoded, and the corresponding control signals, required by the arithmetic unit, are generated. The processes performed by the computer are, essentially, simple but, because they are performed at a very high speed, and with a high degree of co-ordination between the functional units, the computer can carry out highly complex tasks.
Accuracy The accuracy is dependent upon component tolerances and, although results are reasonably accurate, the cost of the hardware required to provide a high degree of accuracy is often prohibitive.	The accuracy is dependent only upon the length of the binary word used and, thus, a very high degree of accuracy is obtained.

(b) (i) Digital computers have the ability to store large quantities of data and, as a result, are often used for the calculation of a company's payroll. The computer is, then, being used as a labour-saving device, as it can calculate the gross pay, tax deduction, etc. of each employee much more quickly and accurately than by any manual method.

Another application of digital computers is for on-line process control. For example, a digital computer can be used to control an airline's seat-reservation system. Such a system may be operated from terminals at reservation desks. The numbers of empty seats on a particular flight are immediately displayed at a terminal before a booking is accepted. When a booking is made, all the details are typed into the computer, from the terminal, for the preparation of passenger lists, etc. by the computer.

(ii) Analog computers perform complex functions at a high speed whilst the actual process under study is in operation. This ability to operate in the real-time mode means that analog computers have many applications in the scientific and industrial fields, in simulating various physical systems or automatically controlling industrial processes.

The simulation of physical systems often involves the solution of differential equations of a high degree which, in most cases, cannot be solved by analytical means. One such system is that of an aeroplane in flight, and an aircraft simulator is an analog computer dedicated to recreating the conditions of an aircraft's cockpit.

Analog computers are extensively used in automatically controlling industrial processes. This could involve the control of a chemical reaction or the mixing of ingredients in the production of foodstuffs. One such application is the control of the whisky-making process. An analog computer controls the mixing of the correct quantities of malt, barley, water, etc., and monitors the fermentation and distillation processes.

Q. 2. (a) *Draw a simple block diagram of a typical digital computer, showing clearly the interconnexions and, hence, the flow of data and control information between the blocks.*

Briefly explain the function of each of the blocks.

(b) *Using the block diagram drawn for part (a), describe the flow of data if the computer were to input 2 integers, add them together and output the sum.*

A. 2. See A.1, Computers A, 1969, *Supplement*, Vol. 63, p. 7, Apr. 1970.

Q. 3. (a) *Convert the following denary numbers into binary, showing all working:*

(i) 3247,

(ii) 631.

(b) *Convert the following octal numbers into denary, showing all working:*

(i) 777,

(ii) 46.

(c) *Convert the following binary numbers into octal:*

(i) 1 011 101,

(ii) 11 010 001 111.

Q. 4. (a) *What is a Venn diagram?*

(b) *Draw the outline of a 2-element Venn diagram.*

(c) *Using the outline drawn for part (b), shade the appropriate areas which represent the following logic statements, where . and + represent the logic functions* AND *and* OR *respectively, and* ‾ *represents the logic function* NOT:

(i) $A . B$,

(ii) $A + B$,

(iii) $\bar{B}$,

(iv) $\bar{A} + \bar{B}$.

A. 4. (a) Simple relationships between logical sets can be illustrated by Venn diagrams. A convenient shape is drawn to represent the contents of each set, and the shapes are positioned in relation to each other according to any properties they may have. For instance, if A and B are disjoint sets, the shapes representing them are drawn such that they do not intersect. If the universal set, E, is shown, it is drawn such that it surrounds all the other sets.

Venn diagrams may be used only to illustrate relationships between sets or logic statements; they cannot be used to prove them.

(b) Sketch (a) shows a 2-element Venn diagram. It consists of 2 intersecting circular shapes, representing the sets A and B, contained within a rectangular shape which represents the universal set, E.

(a)

(b) What are the circumstances in which A is stored in location 50. Give 2 examples.
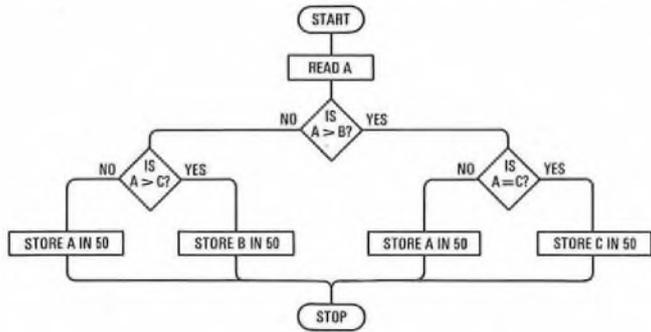
Fig. 1

A. 7. (a) Table 1 gives details of machine-code instructions which could be used to write a program to perform the flow chart's function.

Table 1

Instruction		Function
LD N	(N) → A	Loads content of location N into accumulator. Overwrites any information already in accumulator
STR N	(A) → N	Stores content of accumulator in location N. Does not affect content of accumulator
SUB N	(A) − (N) → A	Subtracts content of location N from content of accumulator and leaves result in accumulator
JPOS N	If (A) > 0 then N → PC	Jumps to address N if content of accumulator is positive
JZER N	If (A) = 0 then N → PC	Jumps to address N if content of accumulator is equal to zero
JUMP N	N → PC	Unconditional jump to address N
STOP		Stops program

A = Accumulator
(A) = Content of accumulator
N = Location N
(N) = Content of location N
PC = Program counter

Using these instructions, a program to perform the flow chart's function is shown in Table 2.

Table 2

Label	Program		Comment
	LD	64	Loads A into accumulator
	SUB	66	Subtracts C from A
	STR	67	Stores (A − C) in location 67
	LD	64	Loads A into accumulator
	SUB	65	Subtracts B from A
	JPOS	L3	Jumps to L3 if (A − B) > 0
	LD	67	Loads (A − C) into accumulator
	JPOS	L1	Jumps to L1 if (A − C) > 0
	LD	64	Loads A into accumulator
	JUMP	L2	Jumps to L2
L1	LD	65	Loads B into accumulator
L2	STR	50	Stores A or B in location 50
	JUMP	L6	Jumps to L6
L3	LD	67	Loads (A − C) into accumulator
	JZER	L4	Jumps to L4 if A = C
	LD	64	Loads A into accumulator
	JUMP	L5	Jumps to L5
L4	LD	66	Loads C into accumulator
L5	STR	50	Stores A or C in location 50
L6	STOP		

(b) A is stored in location 50 only if

(i) $A > B$ and $A \neq C$, or

(ii) $A < B$ and $A \leqslant C$.

Q. 8. (a) With the aid of a circuit diagram, explain the operation of both positive-logic and negative-logic semiconductor-diode AND logic elements.

(b) If a transistor stage is connected to the output of the positive-logic AND gate drawn for part (a), sketch the arrangement and explain its electrical operation. What logic function is now being performed?

A. 8. (a) *Positive Logic*

Considering the circuit diagram shown in sketch (a), when both inputs are at zero volts, diodes D1 and D2 are forward biased and, hence, current flows through the resistance. Since there is only a very small voltage drop across the diodes, the output voltage is approximately zero.

When one input is at −V volts and the other is at zero volts, both diodes are initially forward biased, and conduct until the output voltage becomes zero. The diode with an input of −V volts is still forward biased and, therefore, the output falls towards −V volts. Thus, the diode with an input of zero volts is now reverse biased. The output voltage is, therefore, approximately −V volts.

When both inputs are at −V volts, both diodes are forward biased. The output voltage is, therefore, approximately −V volts.

Using positive-logic convention, in which the more positive voltage level is defined as logic 1 and the less positive voltage level as logic 0, the truth table for the circuit is as follows.

Input 1	Input 2	Output
1	1	1
0	1	0
1	0	0
0	0	0

It can be seen that the above is the truth table for an AND gate.

(a)　　　　(b)
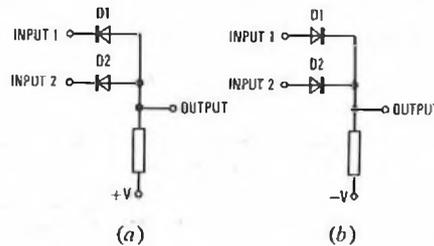
Negative Logic

Considering the circuit diagram shown in sketch (b), when both inputs are at zero volts, diodes D1 and D2 are forward biased and, hence, conduct. There is a very small voltage drop across each diode and, therefore, the output voltage is approximately zero.

When one input is at +V volts and the other is at zero volts, both diodes are initially forward biased, and conduct until the output voltage becomes zero. The diode with an input of +V volts is still forward biased and, therefore, the output voltage rises towards +V volts. Thus, the diode with an input of zero volts is now reverse biased. Therefore, the output voltage is approximately +V volts.

When both inputs are at +V volts, both diodes are forward biased. Thus, the output voltage is approximately +V volts.

Using negative-logic convention, in which the more negative voltage level is defined as logic 1 and the more positive voltage level as logic 0, the truth table for the circuit is as follows.

Input 1	Input 2	Output
1	1	1
0	1	0
1	0	0
0	0	0

It can be seen that the above is the truth table for an AND gate.

(b) Sketch (c) shows a transistor stage connected to the output of the positive-logic AND gate.

When the output of the AND gate is zero volts, the base current of the transistor is zero, and the transistor is, therefore, in its OFF state. The collector current consists only of a very small leakage current and, hence, the potential difference set up across the load resistance is also small. Therefore, the output voltage is approximately −V volts.

31

When the output of the AND gate is $-V$ volts, there is sufficient base current to turn the transistor on. The potential difference between the collector and the emitter falls almost to zero and, therefore, the steady output voltage is approximately zero.

(c)
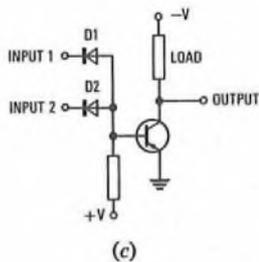
Using positive-logic convention, the truth table for the inputs to the diodes and the output from the transistor is as follows.

Input 1	Input 2	Output
1	1	0
0	1	1
1	0	1
0	0	1

The above is the truth table for a NAND gate and, hence, a logical NAND operation is being performed.

Q. 9. (*a*) *What is meant by the term backing store, and how does it compare with a digital computer's main store?*
(*b*) *Briefly compare the advantages and disadvantages of fixed-head and moving-head discs.*
(*c*) *Explain one major disadvantage of magnetic-tape backing stores.*

A. 9. See A.10, Computers A, 1972, *Supplement*, Vol. 66, p. 43, July 1973.

Q. 10. (*a*) *With the aid of a circuit diagram, illustrate and explain how 3 analog variables can be added together using a passive resistor network.*
(*b*) *From the explanation formed in part* (*a*)*, devise a general formula that will express the output in terms of any number of inputs.*

A. 10. (*a*) The sketch shows a passive resistor network suitable for adding 3 analog variables, V_1, V_2 and V_3.
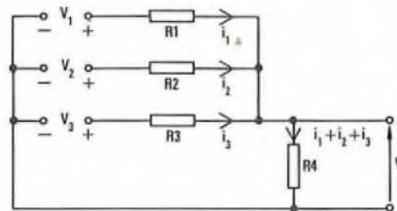
The output voltage, V_4, is given by

$$V_4 = (i_1 + i_2 + i_3)R_4,$$

$$= \left(\frac{V_1 - V_4}{R_1} + \frac{V_2 - V_4}{R_2} + \frac{V_3 - V_4}{R_3} \right) R_4,$$

$$= \left\{ \left(\frac{V_1}{R_1} + \frac{V_2}{R_2} + \frac{V_3}{R_3} \right) - V_4 \left(\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} \right) \right\} R_4.$$

$$\therefore \frac{V_4}{R_4} + V_4 \left(\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} \right) = \left(\frac{V_1}{R_1} + \frac{V_2}{R_2} + \frac{V_3}{R_3} \right).$$

$$\therefore V_4 \left(\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} + \frac{1}{R_4} \right) = \left(\frac{V_1}{R_1} + \frac{V_2}{R_2} + \frac{V_3}{R_3} \right).$$

If $R_1 = R_2 = R_3 = R_4 = R$, then

$$V_4 = \frac{\frac{1}{R}(V_1 + V_2 + V_3)}{\frac{4}{R}} = \frac{V_1 + V_2 + V_3}{4}.$$

(*b*) From part (*a*), a general formula for adding N inputs can be expressed as

$$V_{N+1} = \frac{\sum\limits_{n=1}^{N} V_n}{N + 1}.$$

CORRECTION

COMPUTERS A, 1973 (*Supplement*, Vol. 67, July 1974)

A. 6. (*a*) The failure rate of a component should have been defined as the ratio of the rate of decrease of the number of components surviving to the number of components surviving at any time.
Hence, the expression should have been

$$\lambda = -\frac{1}{N_s} \times \frac{dN_s}{dt}.$$

Similarly, in the example given, the expression should have been

$$\lambda = \frac{-(n_2 - n_1)}{(t_2 - t_1)} \times \frac{1}{n_s}.$$