N6

Industrial Electronics

Lecturer Guide

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MODULE

1 Transients

By the end of this module, students should be able to:

- learn, understand and define the concepts of a transitional response with regard to DC RL and RC circuits, including time constants, the phenomenon 'ringing' and damping;
- apply Kirchhoff's voltage laws mathematically and the exponential laws to RL and RC circuits; and
- graphically represent the charge and discharge cycles of both RL and RC circuit diagrams.

Exercise 1.1 — SB page 5

1. i) $\tau = R \times C$

$= 8 \times 10^{3} \times 100 \times 10^{-6}$

$= 0{,}8 \text{ s}$

ii) $V_C = V_S \times e^{-t/\tau}$

$= 100 \times e^{-1/0{,}8}$

$= 71{,}35 \text{ V}$

iii) $i = I_M \times e^{-t/\tau}$

$= 12{,}5 \times 10^{-3} \times e^{-1{,}5/0{,}8}$

$= 1{,}917 \text{ mA}$

$I_M = \dfrac{V_S}{R}$

$= \dfrac{100}{8 \times 10^{3}}$

$= 12{,}5 \text{ mA}$

iv) $W = \frac{1}{2} \times C \times V^{2}$

$= \frac{1}{2} \times 100 \times 10^{-6} \times 100^{2}$

$= 0{,}5 \text{ J}$

2. i) **Charge phase**

$I_M = \frac{V_S}{R}$

$= \frac{500}{10 \times 10^3}$

$= 0{,}05$ A

Discharge phase

$I_M = \frac{V_S}{R}$

$= \frac{300}{1 \times 10^3}$

$= 0{,}5$ A

ii) $W = \frac{1}{2} \times C \times V^2$

$= \frac{1}{2} \times 10 \times 10^{-6} \times 500^2$

$= 1{,}25$ J

3. $\tau = R \times C$

$= 2 \times 10^6 \times 2 \times 10^{-6}$

$= 4$ s

$I_M = \frac{V_S}{R}$

$= 100/2 \times 10^6$

$= 50\ \mu A$

$i_1 = I_M \times e^{-t/\tau}$

$= 50 \times 10^{-6} \times e^{-1/4}$

$= 38{,}94\ \mu A$

$i_2 = I_M \times e^{-t/\tau}$

$= 50 \times 10^{-6} \times e^{-2/4}$

$= 30{,}32\ \mu A$

$I_3 = I_M \times e^{-t/\tau}$

$= 50 \times 10^{-6} \times e^{-3/4}$

$= 23{,}62\ \mu A$

$i_4 = I_M \times e^{-t/\tau}$

$= 50 \times 10^{-6} \times e^{-4/4}$

$= 18{,}39\ \mu A$

$I_5 = I_M \times e^{-t/\tau}$

$= 50 \times 10^{-6} \times e^{-5/4}$

$= 14{,}33\ \mu A$

$V_C = V_S\,(1 - e^{-t/\tau})$

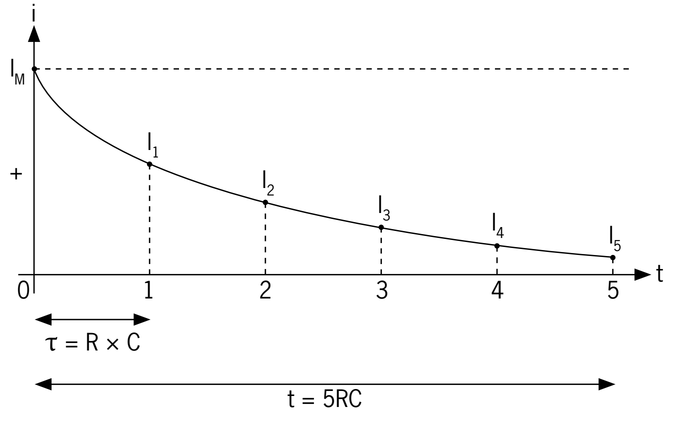
$= 100\,(1 - e^{-4/4})$

$= 63{,}212$ V

$W = \frac{1}{2} \times C \times V^2$

$= \frac{1}{2} \times 2 \times 10^{-6} \times 63{,}212^2$

$= 3{,}996$ mJ



4. $\tau = R \times C$
$= 50 \times 10^{6} \times 50 \times 10^{-6}$
$= 2\,500$ s

$V_C = V_S \times e^{-t/\tau}$
$500 = 60 \times e^{-t/2\,500}$
$8{,}333 = e^{-t/2\,500}$
$\frac{-t}{2\,500} = \ln 8{,}333$
$-t = 2\,500 \times \ln 8{,}333$
$= 5\,300{,}55$ s
$= 1$ h 28 m 20,55 s

5. i) $\tau = R \times C$
$= 25 \times 10^{3} \times 5 \times 10^{-6}$
$= 0{,}125$ s

$V_C = V_S\,(1 - e^{-t/\tau})$
$\frac{d\,V_C}{di} = \frac{V_S \times e^{-t/\tau}}{\tau}$
$= \frac{V_S}{\tau}$
$= \frac{50}{0{,}125}$
$= 400\ V_S^{-1}$

ii) $V_R = V_S \times e^{-t/\tau}$
$0{,}25 = e^{-t/\tau}$
$-t = 0{,}25 \ln 0{,}125$
$t = 519{,}86$ ms

iii) $i = I_M \times e^{-t/\tau}$
$= 2 \times 10^{-3} \times e^{-0{,}1/0{,}125}$
$= 898{,}657\ \mu A$

$I_M = \frac{V_S}{R}$
$= \frac{50}{25} \times 10^{3}$
$= 2$ mA

6. $\tau = R \times C$
$= 680 \times 6{,}8 \times 10^{-6}$
$= 0{,}0046$ s

$t = \frac{1}{f}$
$= \frac{1}{100}$
$= 0{,}01$ s

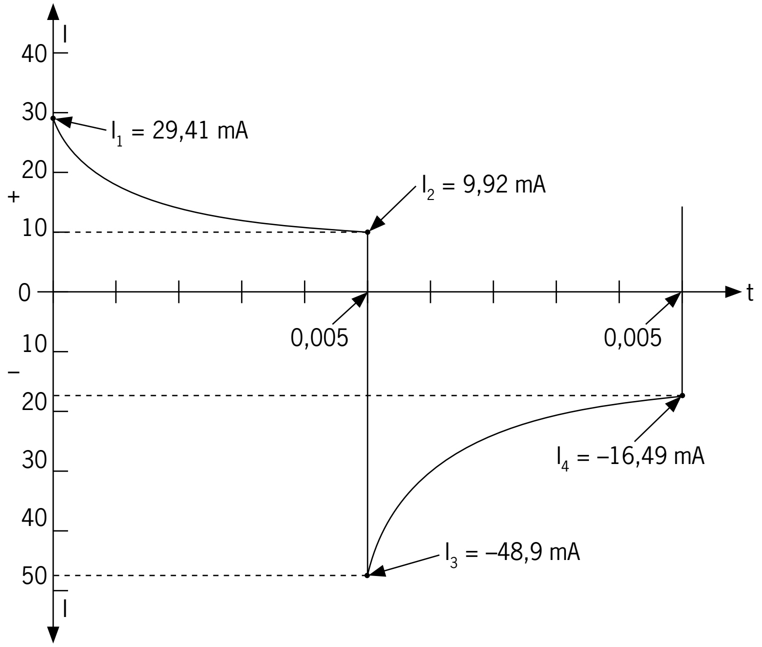
$i_1 = \frac{V_S}{R}$
$= \frac{20}{680}$
$= 29{,}41$ mA

$i_2 = i_1 \times e^{-t/\tau}$
$= 29{,}41 \times 10^{-3} \times e^{-0{,}005/0{,}0046}$
$= 9{,}92$ mA

$V_C = V_S\,(1 - e^{-t/\tau})$
$= 20\,(1 - e^{-0{,}005/0{,}0046})$
$= 13{,}255$ V

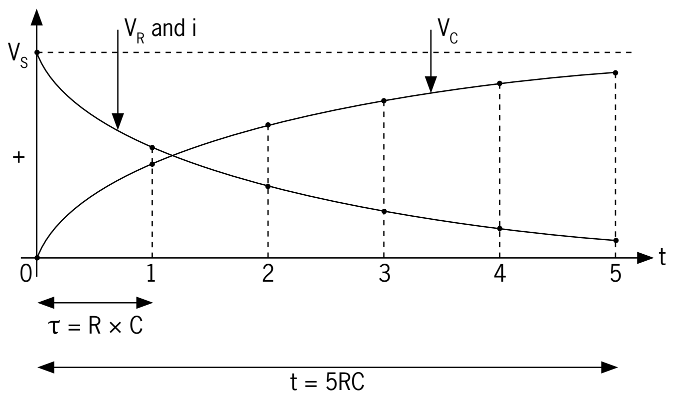
$I_3 = -(\frac{V_2 + V_C}{R})$
$= -(\frac{20 + 13{,}255}{680})$
$= -48{,}9$ mA

$i_4 = -(I_3 \times e^{-t/\tau})$
$= -(48{,}9 \times 10^{-3} \times e^{-0{,}005/0{,}0046})$
$= -16{,}49$ mA



7. A transient is a period during which known voltage and/or current values will change from existing values to new values caused by a change in network parameters.

8.



The time constant of an energy-storing circuit may be defined as the time taken for the circuit to reach steady-state conditions after a disturbance, assuming a constant rate of change in the charging or discharging current.

A time constant is the product of the capacitor value and the resistor value, i.e. $\tau = R \times C$ seconds, and is the time the network will take to undergo a change of ± 63% of its initial value. No difference is made between charging and discharging.

During every time constant, the capacitor voltage will change by 63,2% and will increase towards maximum.

During every time constant, the charging current through the resistor will change and decrease by 62,3% and will fall to minimum.

During every time constant, the resistor voltage will decrease by 63,2% and will fall to minimum.

9. Kirchhoff's laws

Charging cycle

$\tau = R \times C$

$V_R = V_S \times e^{-t/\tau}$

$I = I_M \times e^{-t/\tau}$

$V_C = V_S\,(1 - e^{-t/\tau})$

$I_M = \frac{V_S}{R}$

Discharge cycle

$\tau = R \times C$

$V_R = -(V_S \times e^{-t/\tau})$

$I = -(I_M \times e^{-t/\tau})$

$V_C = V_S \times e^{-t/\tau}$

$I_M = \frac{V_S}{R}$

Exercise 1.2 — SB page 10

1. i) $I_M = \frac{V_S}{R} = \frac{200}{5} = 40\ A$

ii) $\tau = \frac{L}{R} = \frac{10 \times 10^{-3}}{5} = 2\ ms$

iii) $i = I_M \times (1 - e^{-t/\tau})$
$= 40 \times (1 - e^{-25 \times 10-3/2 \times 10-3})$
$= 39{,}999\ mA$

iv) $W = \frac{1}{2} \times L \times I^2$
$= \frac{1}{2} \times 10 \times 10^{-3} \times 40^2$
$= 8\ J$

2. $\tau = \frac{L}{R} = \frac{0{,}4}{10} = 40\ ms$

$I_M = \frac{V_S}{R} = \frac{100}{10} = 10\ A$

i) $i = I_M \times (1 - e^{-t/\tau})$
$5 = 10 \times (1 - e^{-t/\tau})$
$0{,}5 = 1 - e^{-t/\tau}$
$-t = 40 \times 10^{-3} \ln 0{,}5$
$= 27{,}72\ ms$

ii) $i_{(t=0)} = \frac{I_M}{\tau} = \frac{10}{40 \times 10^{-3}} = 250\ A/s$

iii) $W = \frac{1}{2} \times L \times I^2$

$= \frac{1}{2} \times 0{,}4 \times 10^2$

$= 20$ J

3. i) $I_M = \frac{V_S}{R}$

$R = \frac{V_S}{I_M}$

$= \frac{200}{2}$

$= 100\ \Omega$

$\tau = \frac{L}{R}$

$= \frac{10}{100}$

$= 0{,}1$ s

$I_{(t=0)} = \frac{I_M}{\tau}$

$= \frac{2}{0{,}1}$

$= 20\ \text{As}^{-1}$

ii) $i = I_M (1 - \tau/I_M \times di/dt)$

$i = 2(1 - 0{,}1/2 \times 5)$

$i = 1{,}5$ A

4. $i = I_M \times (1 - e^{-t/\tau})$

$I_M = \frac{i}{1 - e^{-t/\tau}}$

$= \frac{0{,}2}{1 - e^{-0{,}003/0{,}003}}$

$= 0{,}316$ A

$\tau = \frac{L}{R}$

$L = \tau \times R$

$= 3 \times 10^{-3} \times 632{,}9$

$= 1{,}898$ H

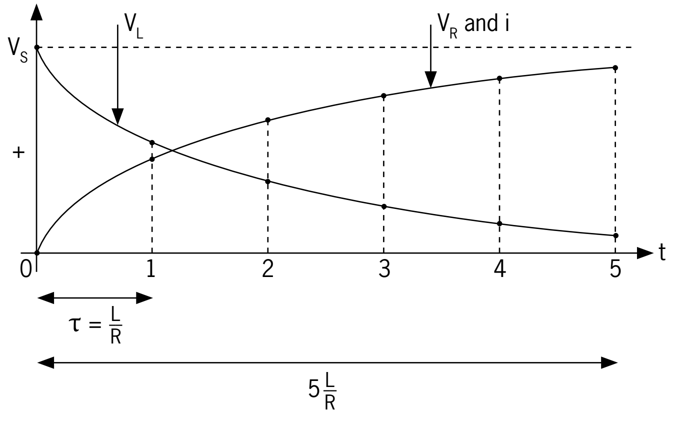
$I_M = \frac{V_S}{R}$

$R = \frac{V_S}{I_M}$

$= \frac{200}{0{,}316}$

$= 632{,}9\ \Omega$

5.



The time constant of an RL circuit is the time taken for the circuit to reach steady-state conditions after a disturbance, assuming a constant rate of change in charging or discharging current.

A time constant is given by $\tau = L/R$ seconds and is the time the network will take to undergo a change of ± 63% of its initial value. No difference is made between charging and discharging.

During every time constant, the inductor voltage will change by 63,2% and will decrease towards zero.

During every time constant, the charging current through the resistor will change and increase by 62,3% and will rise to maximum.

During every time constant, the resistor voltage will increase by 63,2% and will rise to maximum.

Exercise 1.3 — SB page 13

a) $$\frac{1}{LC} = \frac{R^2}{4L^2}$$

$$\therefore R = \frac{\sqrt{4L}}{C} = \frac{\sqrt{4 \times 20 \times 10^{-3}}}{10 \times 10^{-6}}$$

$$= 89{,}14\ \Omega$$

Therefore we will use a value of 89 Ω for oscillations to take place. This is done so that oscillations may take place, which is the desired action we require.

b) $$f_N = \frac{1}{2\pi}\sqrt{\frac{1}{LC} - \frac{R^2}{4L^2}}$$

$$= \frac{1}{2\pi}\sqrt{\frac{1}{20 \times 10^{-3} \times 10 \times 10^{-6}} - \frac{89^2}{4 \times (20 \times 10^{-3})^2}}$$

$$= 35{,}37\ \text{Hz}$$

$$\int = \frac{\pi R}{\omega L}$$

$$= \frac{\pi \times 89}{2 \times 3{,}142 \times 21{,}2 \times 20 \times 10^{-3}}$$

$$= 62{,}906$$

c) $$n = \frac{1 + 4{,}605}{\int}$$

$$= \frac{1 + 4{,}605}{62{,}906}$$

$$= 1{,}0732$$

$$t = \frac{1}{f_N \times n}$$

$$= \frac{1}{35{,}37 \times 1{,}0732}$$

$$= 30{,}34\ \text{ms}$$

Exercise 1.4 SB page 14

1. $\tau = \frac{L}{R}$

$= \frac{20 \times 10^{-3}}{200}$

$= 0{,}1$ ms

$I_M = \frac{V_S}{R}$

$= \frac{50}{200}$

$= 0{,}25$ A

i) $i_{(t=0)} = \frac{I_M}{\tau}$

$= \frac{0{,}25}{0{,}1 \times 10^{-3}}$

$= 2{,}500$ As^{-1}

ii) $V_L = V_R$

$V_S \times e^{-t/\tau} = V_S\,(1 - e^{-t/\tau})$

$1 - e^{-t/\tau} = e^{-t/\tau}$

$-t = \tau \ln 0{,}5$

$= 0{,}1 \times 10^{-3} \ln 0{,}5$

$= 69{,}31$ ms

iii) $i = I_M \times (1 - e^{-t/\tau})$

$0{,}8 = 1 - e^{-t/\tau}$

$-t = 0{,}1 \ln 0{,}2$

$= 160{,}94$ ms

iv) $W = \frac{1}{2} \times L \times I_M^{\;2}$

$= \frac{1}{2} \times 20 \times 10^{-3} \times 0{,}25^2$

$= 0{,}625$ mJs

2. i) $i = I_M \times (1 - e^{-t/\tau})$

$I_M = \frac{i}{1 - e^{-t/\tau}}$

$= \frac{0{,}250 \times 10^{-3}}{1 - e^{-0{,}002/0{,}002}}$

$= 395{,}5$ mA

$\tau = \frac{L}{R}$

$L = \tau \times R$

$= 0{,}002 \times 252{,}84$

$= 505{,}68$ mH

ii) $I_M = \frac{V_S}{R}$

$R = \frac{V_S}{I_M}$

$= \frac{100}{395{,}5 \times 10^{-3}}$

$= 252{,}84\ \Omega$

iii) $i = I_M \times (1 - e^{-t/\tau})$

$0{,}6 = 1 - e^{-t/\tau}$

$-t = 0{,}002 \ln 0{,}4$

$= 1{,}833$ ms

3. i) $\tau = R \times C$

$= 20 \times 10^3 \times 1 \times 10^{-6}$

$= 20\ \text{ms}$

Charge phase

$I_M = \frac{V_S}{R}$

$= \frac{100}{20 \times 10^3}$

$= 5\ \text{mA}$

$\frac{di}{dt} = \frac{I_M}{\tau}$

$= \frac{5 \times 10^{-3}}{20 \times 10^{-3}}$

$= 0{,}25\ \text{As}^{-1}$

ii) $\tau = R \times C$

$\tau = 1 \times 10^6 \times 1 \times 10^{-6}$

$= 1\ \text{s}$

Discharge phase

$I_M = \frac{V_S}{R}$

$= \frac{100}{1 \times 10^6}$

$= 100\ \mu\text{A}$

$\frac{di}{dt} = \frac{I_M}{\tau}$

$= \frac{100 \times 10^{-6}}{1}$

$= -100\ \mu\text{As}^{-1}$

iii) $i = I_M \times e^{-t/\tau}$

$= 5 \times 10^{-3} \times e^{-0{,}001/0{,}02}$

$= 4{,}756\ \mu\text{A}$

$V_C = V_S(1 - e^{-t/\tau})$

$= 100(1 - e^{-0{,}001/0{,}02})$

$= 4{,}877\ \text{V}$

iv) $W = \frac{1}{2} \times C \times V^2$

$= \frac{1}{2} \times 1 \times 10^{-6} \times 100^2$

$= 5\ \text{mJ}$

4. $\tau = R \times C$

$= 1 \times 10^3 \times 10 \times 10^{-6}$

$= 0{,}01\ \text{s}$

$t = \frac{1}{f}$

$= \frac{1}{50}$

$= 0{,}02\ \text{s}$

$i_1 = \frac{V_S}{R}$

$= \frac{5}{1 \times 10^3}$

$= 5\ \text{mA}$

$i_2 = i_1 \times e^{-t/\tau}$

$= 5 \times 10^{-3} \times e^{-0{,}01/0{,}01}$

$= 1{,}84\ \text{mA}$

$V_C = V_S(1 - e^{-t/\tau})$

$= 5(1 - e^{-0{,}01/0{,}01})$

$= 3{,}16\ \text{V}$

$I_3 = \frac{-(V_2 + V_C)}{R}$

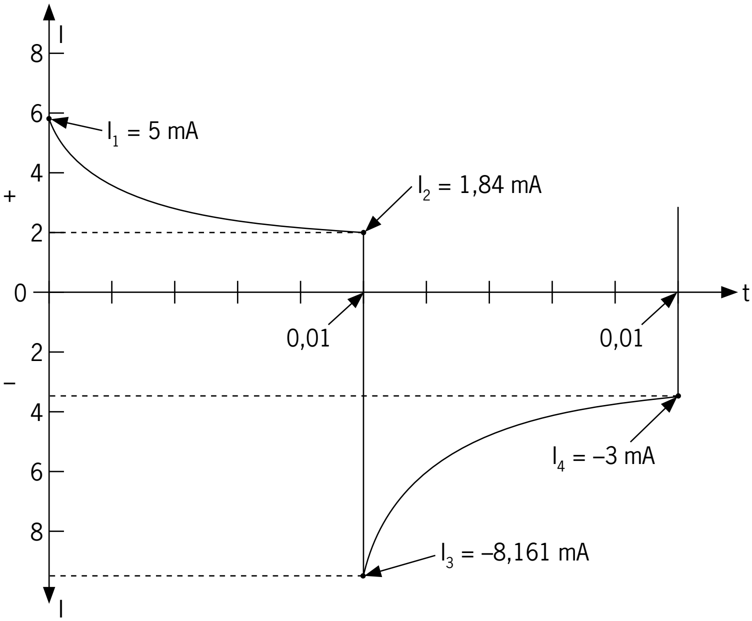
$= \frac{-(5 + 3{,}16)}{1 \times 10^3}$

$= -8{,}16\ \text{mA}$

$i_4 = -(I_3 \times e^{-t/\tau})$

$= -(8{,}16 \times 10^{-3} \times e^{-0{,}01/0{,}01})$

$= -3\ \text{mA}$



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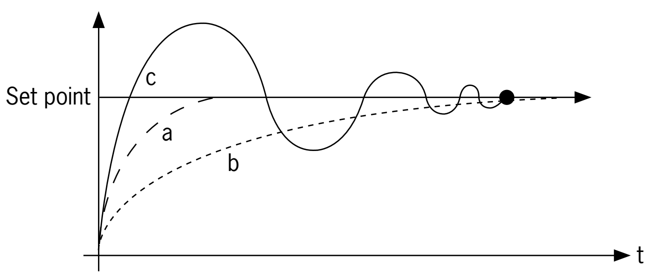


a = critical damping

The network is critically damped when $\frac{1}{LC} = \frac{R^2}{4L^2}$ and all energy stored is dissipated by the resistance before the current even has a chance to reverse. The network is actually on the verge of becoming oscillatory.

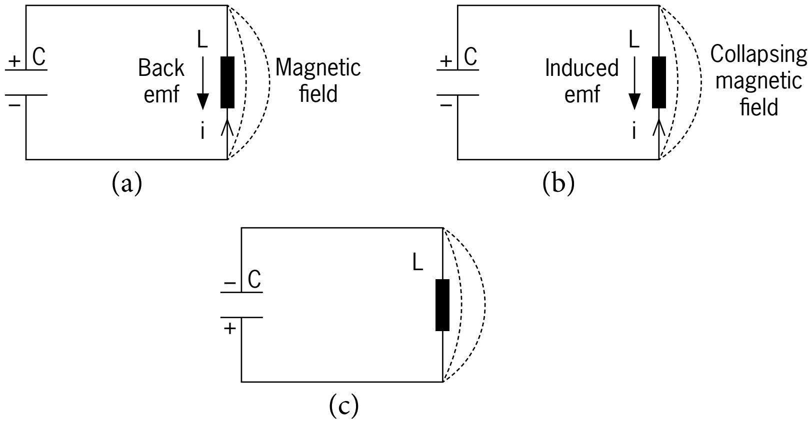
b = overdamping

The network is overdamped when $\frac{R^2}{4L^2}$ is greater than $\frac{1}{LC}$ and in this instance the resistance is so high that it causes the current to build up to a maximum from zero and dies away to zero over a relatively long period.

c = critical damping

The network is underdamped when $\frac{R^2}{4L^2}$ is less than $\frac{1}{LC}$ > and in this instance the energy is transferred a number of times between the capacitor and the inductor before the energy is dissipated through the resistance.

6. 6.1



Assuming the capacitor C is to be charged to the polarities indicated in (a) and is then connected to the inductor, the energy stored in capacitor C will discharge trough inductor L and will set up a magnetic field around the inductor, as illustrated in (a).

At the point where capacitor C has discharged completely, the current flow will cease, inducing an emf in the inductor, but in the opposite direction to the voltage that was applied by capacitor C, as in (b). This induced emf now causes a current to flow in the opposite direction than that which caused the induced emf and will charge the capacitor in the opposite direction than in (a), as illustrated in (c).

This process will continue indefinitely, provided there is no resistance losses. If we assume the above conditions to be valid, then the network will resonate at its resonant frequency.

$$f_r = \frac{1}{2 \times \pi \times \sqrt{(L \times C)}}$$

The above conditions are not practical, as the inductor will always have a resistance that will dissipate power during each cycle. This will result in the amplitude of the oscillation to diminish during each interchange and will eventually die away completely. We therefore need to determine the natural frequency of oscillation that will always be less than the resonant frequency.

$$f_N = \frac{1}{2 \times \pi \sqrt{(\frac{1}{LC} - \frac{R^2}{4L^2})}}$$

6.2 **Desirable**

Circuits containing a capacitor and inductance may resonate when suitably supplied with AC power and are very useful in communication networks when resonance is required.

Undesirable

It causes extra current to flow, thereby wasting energy and causing extra heating of the components.

It can cause unwanted electromagnetic radiation to be emitted.

It can delay arrival at a desired final state (increase settling time).

7. i) $\int \frac{\pi.R}{\omega.L} = \frac{\pi \times 73}{2.\pi.f \times 6}$

$$= \frac{\pi \times 73}{2.\pi \times 29{,}72 \times 6}$$

$$= 40{,}935$$

ii) $n = \frac{1 + 4{,}605}{\int}$

$$= 1 + \frac{4{,}605}{40{,}935}$$

$$= 1{,}1125$$

$$t = \frac{1}{f_n} \times n$$

$$= \frac{1}{29{,}72} \times 1{,}1125$$

$$= 0{,}03743$$

$$= 37{,}43 \text{ ms}$$

MODULE

2 *Transducers*

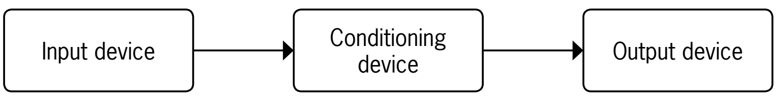
By the end of this module, students should be able to:

- describe the terms related to transducers, data acquisition systems and operational amplifiers in their own words;
- draw and label the diagrams related to the above-mentioned systems; and
- perform calculations related to operational amplifier circuits and interfacing networks.

Exercise 2.1 — SB page 31

1. A transducer is a device that will convert one form of energy into another form of energy.
2. 2.1 A transducer is a device that will convert one form of energy into another form of energy.

   2.2



   The input device will receive a quantity, normally non-electrical. This quantity under measurement is converted into an electrical quantity and will be passed onto the conditioning device in the form of a proportional electrical signal.

   The output device may be a simple indicating meter, a CRO, a chart recorder for visual display or a standard 1 V to 5 V or 4 mA to 20 mA output device.

2.3 Type of transducer, operating principle and application

2.4
- Fundamental transducer parameters (type, range, sensitivity, etc.)
- Physical conditions (corrosion resistance, connection provisions, etc.)
- Ambient conditions (non-linearity, resolution, frequency response, etc.)
- Environmental conditions (temperature, vibration, shock, etc.)
- Compatibility of equipment (zero balance, impedance matching, etc.)

2.5 Determine the physical quantity that needs to be measured.
Which transducer principle is best suited to measure a particular quantity?
What is the level of accuracy that will be required for this measurement?

2.6 Data acquisition systems are used to measure and record signals obtained.

3.

Active transducers		
Thermocouple	An emf will be generated across the junction of two dissimilar metals when the junction temperature is varied.	Temperature, heat flow and radiation
Piezo-electric	An emf will be generated when an external force is applied across certain crystalline materials.	Pressure changes, vibration, acceleration and sound
Photovoltaic cells	An emf will be generated when incident light strikes the surface of a semiconductor junction.	Solar cell and light meter
Moving-coil generator	Motion of a coil (inductor) in a magnetic field.	Velocity

Passive types		
Resistance strain gauge	Resistance of wire changes when subjected to elongation or compression due to external stress being applied.	Measurement of force, torque or displacement
Resistance thermometer	A resistance wire with a high positive temperature coefficient changes with a variance in temperature.	Measurement of temperature and radiant heat
Thermistor	The resistance of certain metal oxides with a negative temperature coefficient changes with a variance in temperature.	Measurement of temperature
Resistance hygrometer	The resistance of conductive material changes with a variance in moisture content.	Relative humidity
Photoconductive cell	The resistance of a cell changes with a variance in incident light.	Photosensitive relay

4. **Transducers**
For translating physical parameters into electrical signals

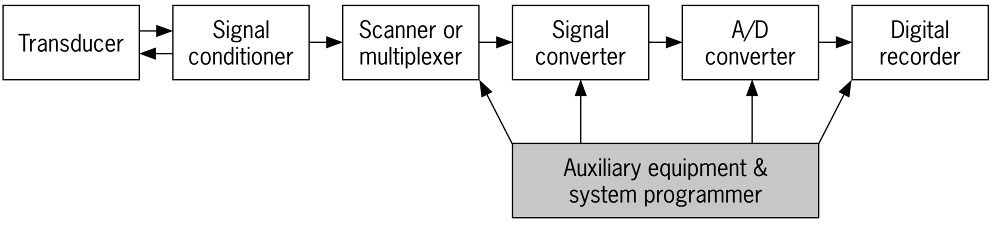
Signal conditioners
For amplification, linearisation, attenuation, impedance transformation and modifying or selecting certain portions of these signals

Visual display units
For continuous monitoring of the input signals; these devices may include CROs, ORCs, panel meters, numerical displays, etc.

Graphic recording instruments
For obtaining permanent records of the input data; these instruments include stylus-and-ink recorders to provide continuous records on paper charts or optical recording systems such as ultraviolet recorders

Computer systems (i.e. magnetic tape or disk)
For acquiring input data, preserving their original electrical form and reproducing them at a later date for more detailed analysis

5.



Transducer
A transducer translates physical parameters into electrical signals acceptable by the acquisition (parameters that you have required before) system. Some typical parameters include temperature, pressure, acceleration, weight displacement and velocity. Electrical quantities, such as voltage, resistance or frequency, may also be measured directly.

Signal conditioner
In general, a signal conditioner includes the supporting circuitry for the transducer. This circuitry may provide excitation power, balancing circuits and calibration elements. An example of a signal conditioner is a strain-gauge bridge balance and power supply unit.

Scanner or multiplexer
The scanner or multiplexer accepts multiple analogue inputs and sequentially connects them to one measuring instrument.

Signal converter
The signal converter translates the analogue signal into a form acceptable by the analogue-to-digital converter. An example of a signal converter is an amplifier for amplifying low-level voltages generated by thermocouples or strain gauges.

Analogue-to-digital (A/D) converter
This converter converts the analogue voltage into its equivalent digital form. The output of the A/D converter may be displayed visually and is also available as voltage outputs in discrete steps for further processing or recording on a digital recorder or computer.

Auxiliary equipment (extra help or support)
This section contains instruments for system programming functions and digital data processing. Typical auxiliary functions include linearisation and limit comparison. These functions may be performed by individual instruments or by a digital computer.

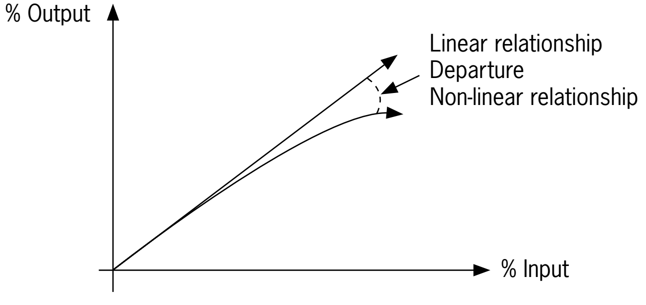
Digital recorder

The digital recorder records digital information on a computer, perforated paper tape, magnetic tape, typewritten pages, pen recorders or a combination of these systems. The digital recorder may be preceded by a coupling unit that translates the digital information into the proper form for entry into the particular digital recorder selected.

6. 6.1 Because most transducers have a low-level voltage output, amplification is required in most instances. A reliable amplification is commonly accomplished by means of operational amplifiers.

   6.2 Because the operation amplifier is limited in terms of the magnitude of the input voltages that can be applied, it will at times be required that the input voltages be attenuated. This attenuation can be accomplished by a suitable resistive network.

6.3



Very few devices are absolutely linear and one must find a way for such devices to appear linear. In order to accomplish this, we need to take the output from the device and provide for an inverse non-linear correction.

6.4 Analogue techniques are used for offsetting in order to shift the reference level of a signal by a pre-set amount.

6.5 Common forms of signal conditioning:

- 1 V to 5 V standard voltage range
- 4 mA to 20 mA standard current range

Note the following:
The two systems may never be mixed.
1 V and 4 mA indicate zero change.

This conditioning is normally applied and performed by operational amplifiers, but it should be kept in mind that whatever form of conditioning is selected, it will be determined by the characteristics of the selected transducer.

6.6 Connecting different types of circuits, different analogue or digital units and inputs or loads to other electronic units all require some sort of interfacing. Interface circuits may be categorised as either driver or receiver units.

A receiver essentially accepts inputs, providing high-input impedance to minimise loading of the input signal.

A driver circuit provides the output signal at voltage or current levels suitable for operating a number of loads.

Voltage-to-current interfacing is used for converting a conditioned voltage signal into a standard current range.

Current-to-voltage interfacing is used for the conversion of a conditioned current signal into a standard voltage range.

A buffer isolates the input from the output in order to eliminate any errors that may occur.

7. 7.1 The operational amplifier (op-amp) is a very high-gain differential amplifier employing voltage feedback to provide a stabilised voltage gain.

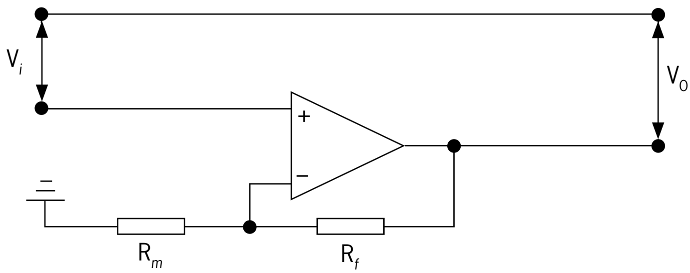
   7.2
   - Very high open-loop gain (no signal-feedback condition)
   - High-input impedance
   - Low-output impedance
   - Wide bandwidth
   - Fast slew (turn or swing round) rate

   7.3
   - Infinite input impedance
   - Infinite bandwidth
   - Zero output impedance
   - Zero open-loop voltage
   - Zero offset voltage

   7.4 Applications include interfacing, instrumentation circuits, analogue computers, summing, differentiating, integrating, inverting, etc.

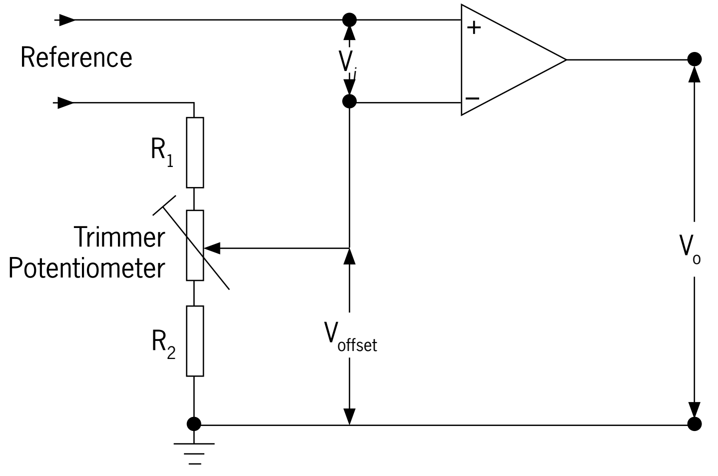
   7.5 The analogue comparator circuit utilises both the inverting and the non-inverting inputs of an op-amp for different incoming analogue signals. The comparator then generates a digital output signal that indicates the relative states between the two different input analogue signals.

8. 8.1

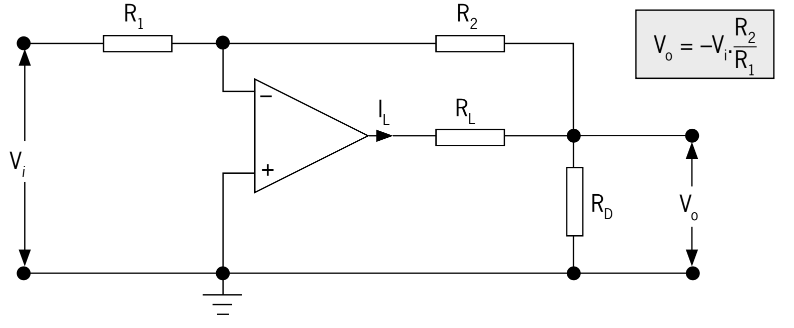


$$V_O = V_i\left(1 + \frac{R_f}{R_m}\right)$$

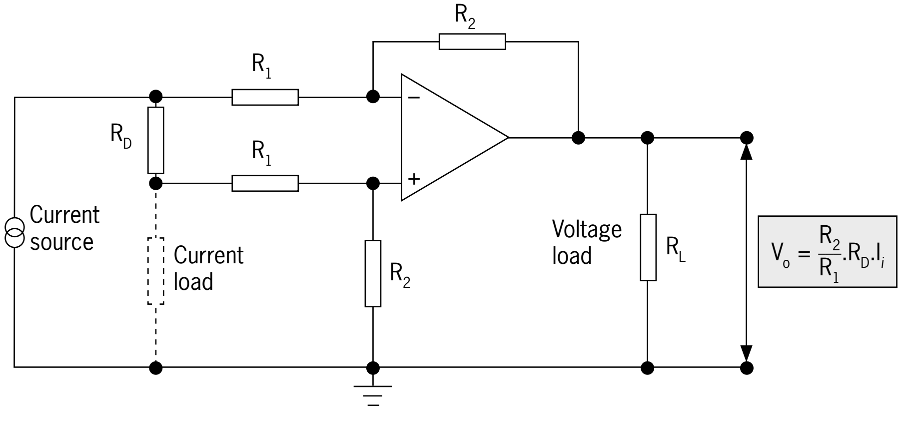
8.2


Reference
+
−
V_i
R_1
Trimmer
Potentiometer
R_2
V_{offset}
V_o

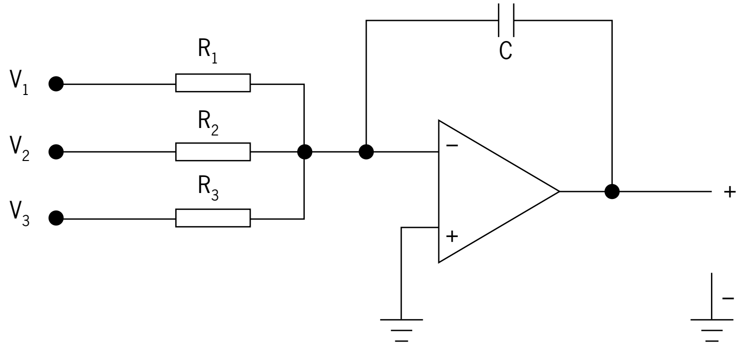
8.3


R_1
R_2
$V_o = -V_i \cdot \frac{R_2}{R_1}$
−
+
I_L
R_L
V_i
R_D
V_o

8.4

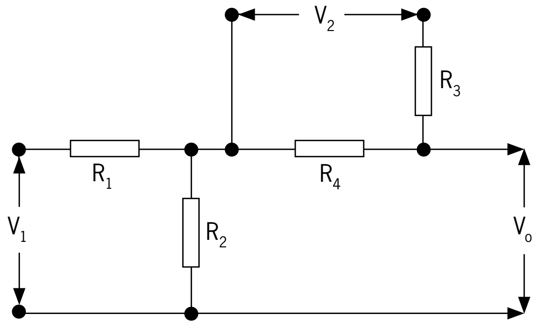

R_2
R_1
R_D
R_1
−
+
Current
source
Current
load
R_2
Voltage
load
R_L
$V_o = \frac{R_2}{R_1} . R_D . I_i$

8.5



$$V_O = \frac{-1}{C \times R_1}\int V_1(t) + \frac{1}{C \times R_2}\int V_2(t) + \frac{1}{C \times R_3}\int V_3(t)$$

9.



$$\boxed{V_o = A.\left\{\left(\frac{V_1.R_2}{R_1 + R_2}\right) + \left(\frac{V_2.R_4}{R_3 + R_4}\right)\right\}}$$

$$V_1 = \frac{R_2}{R_1 + R_2} \times V_1 \qquad \text{and} \qquad V_2 = \frac{R_3}{R_3 + R_4} \times V_2$$

$$V_{OT} = \frac{R_2}{R_1 + R_2} \times V_1 + \frac{R_3}{R_3 + R_4} \times V_2$$

10. $V_{IN} = \frac{1,5\ \text{mV}}{20\ °\text{C}} \times 750\ °\text{C}$

$= \frac{1,5 \times 10^{-3}}{20\ °\text{C}} \times 750\ °\text{C}$

$= 0,057\ \text{V}$

Inverting

$V_O = -\,V_{IN} \times \frac{R_F}{R_M}$

$R_F = R_M \times \frac{V_O}{V_{IN}}$

$= 1,5 \times 10^3 \times \frac{10}{0,057}$

$= 263,16\ \text{K}\Omega$

Non-inverting

$V_O = V_{IN} \times 1 + \frac{R_F}{R_M}$

$R_F = R_M \times \frac{V_O - 1}{V_{IN}}$

$= 1,5 \times 10^3 \times \frac{10 - 1}{0,057}$

$= 236,84\ \text{K}\Omega$

11. For $\Delta t = 20\ °C$ above nominal

$$\Delta R = \Delta t \times R_{NOM} \times \alpha$$
$$= 20 \times 0{,}5 \times 10^3 \times 0{,}001$$
$$= 10\ \Omega$$

$$R_{TH} = R_{NOM} + \Delta R$$
$$= 0{,}5 \times 10^3 + 10$$
$$= 510\ \Omega$$

$$V_{D1} = V_S \times \frac{R_1}{R_1 + R_2} - \frac{R_3}{R_3 + R_{TH}}$$
$$= 10 \times \frac{10 \times 10^3}{10 \times 10^3 + 1 \times 10^3} - \frac{5 \times 10^3}{5 \times 10^3 + 510}$$
$$= 10 \times (0{,}909 - 0{,}907)$$
$$= 20\ mV$$

For $\Delta t = 60\ °C$ below nominal

$$\Delta R = \Delta t \times R_{NOM} \times \alpha$$
$$= 20 \times 0{,}5 \times 10^3 \times 0{,}001$$
$$= 10\ \Omega$$

$$R_{TH} = R_{NOM} - \Delta R$$
$$= 0{,}5 \times 10^3 - 10$$
$$= 480\ \Omega$$

$$V_{D2} = V_S \times \frac{R_1}{R_1 + R_2} - \frac{R_3}{R_3 + R_{TH}}$$
$$= 10 \times \frac{10 \times 10^3}{10 \times 10^3 + 1 \times 10^3} - \frac{5 \times 10^3}{5 \times 10^3 + 480}$$
$$= 10 \times (0{,}909 - 0{,}912)$$
$$= -30\ mV$$

$$V_{DT} = V_{D1} - (-V_{D2})$$
$$= 20 \times 10^{-3} - (-30 \times 10^{-3})$$
$$= 50 \times 10^{-3}\ V$$

12. $$V_O = V_i \times 1 + \frac{R_F}{R_M}$$

$$R_M = \frac{R_F}{\frac{V_O}{V_{IN}} - 1}$$

$$= \frac{1 \times 10^6}{\frac{5}{50 \times 10^{-3}}} - 1$$

$$= 10{,}1\ K\Omega$$

13. $$I_L = \frac{R_2}{R_1 \times R_D} \times V_{IN}$$

$$R_D = \frac{R_2}{R_1 \times I_L} \times V_{IN}$$

$$= \frac{500 \times 10^3}{10 \times 10^3 \times 20 \times 10^{-3}} \times 10$$

$$= 25\ k\Omega$$

$$I_L = \frac{R_F}{R_1 \times R_D} \times V_{IN}$$

$$= \frac{500 \times 10^3}{10 \times 10^3 \times 25 \times 10^3} \times 6{,}8$$

$$= 13{,}6\ mA$$

14. $R_D = \frac{V_O}{I_{IN}}$

$= \frac{5}{20} \times 10^{-3}$

$= 250\ \Omega$

$V_O = R_D \times I_{IN}$

$= 250 \times 13{,}5 \times 10^{-3}$

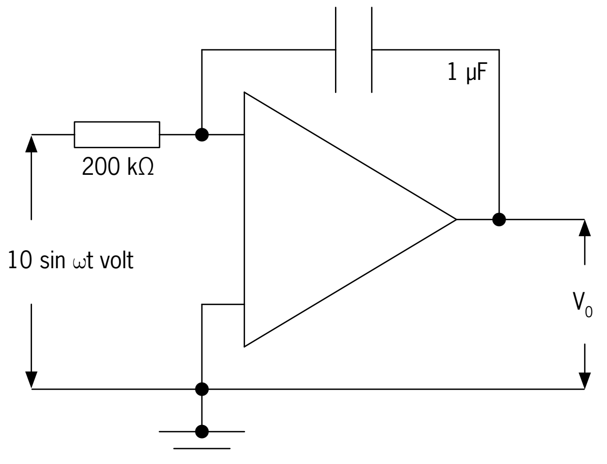
$= 3{,}375\ V$

15. $V_O = -R_F \times \frac{V_1}{R_1} + \frac{V_2}{R_2} + \frac{V_3}{R_3}$

$-7 = -1\ 000 \times 10^3 \times \frac{1}{500 \times 10^3} + \frac{V_2}{1\ 000 \times 10^3} + \frac{3}{1 \times 10^6}$

$V_2 = 2\ V$

16.

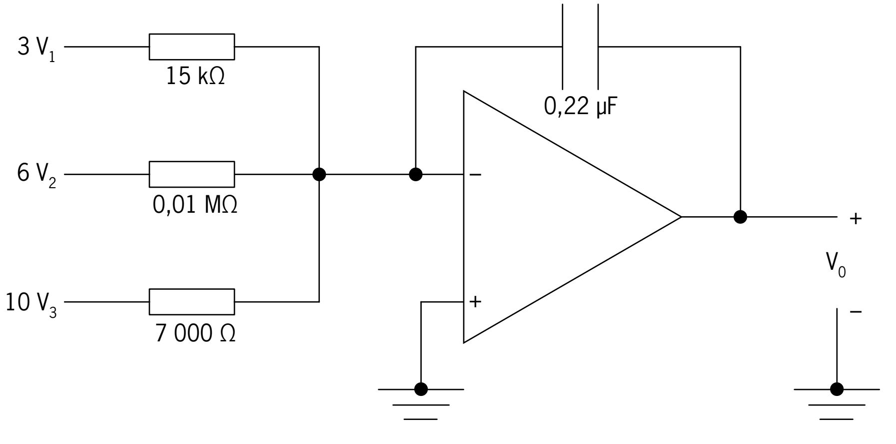


$V_O = -\frac{1}{R \times C \int V_i\,(t)dt}$

$V_O = -\frac{1}{100 \times 10^3 \times 10^{-6} \int 10 \sin \omega t(t)dt}$

$= -100 \times \omega(\cos \omega t - 1)$

17.



i) $V_O = -\left[\dfrac{1}{C \times R_1 \int V_1(t)} + \dfrac{1}{C \times R_2 \int V_2(t)} + \dfrac{1}{C \times R_3 \int V_3(t)}\right]$

$= -\left[\dfrac{1}{0{,}22 \times 10^{-6} \times 15 \times 10^3 \int 3(t)} + \dfrac{1}{0{,}22 \times 10^{-6} \times 0{,}01 \times 10^6 \int 6(t)}\right.$

$\left. + \dfrac{1}{0{,}22 \times 10^{-6} \times 7\,000 \int 10(t)}\right]$

$= -[909{,}091(t) + 2\,727{,}273(t) + 6\,493{,}506(t)]$

$= -10\,129{,}87(t)$ V

ii) $V_O(t) = 10\,129{,}87 \times 200 \times 10^{-6}$

$= -2{,}026$ V

iii) $V_O = \dfrac{dv}{dt}$

$= \dfrac{10\,129}{0{,}1 \times 10^{-3}}$

$= 1{,}013\ V_S^{-1}$

MODULE

3 Ultrasonics

By the end of this module, students should be able to:

- describe the term 'ultrasonic' with reference to its properties and generation;
- draw and describe transducers used in the ultrasonic industry; and
- draw and describe the practical applications of selected ultrasonic devices as applied to the electronic industry.

Exercise 3.1 SB page 43

1. 1.1 Ultrasonic energy refers to acoustic energy above the audio range of 20 kHz.

   1.2 Acoustic energy is generated in a medium by creating alternating regions of high and low pressures (cavitation) that will propagate through the medium at a rate proportional to the density of the medium. The transmission of energy takes place through collisions between the atoms or molecules of the medium.

   1.3 Each group of molecules is caused to propagate under the influence of acoustic energy and will collide with adjacent groups and, in doing so, these molecules transfer their acquired kinetic energy to the adjacent group. In this manner, all acquired kinetic energy is transferred further.

   Ultrasonic energy is best propagated by solids and liquids, attenuated by air and completely stopped by a vacuum.

   1.4 The energy can be transmitted in the form of a narrow beam, and as this energy can be concentrated in a narrow beam, this characteristic enables waves to be generated at high intensities and high efficiency, particularly in solids and liquids.

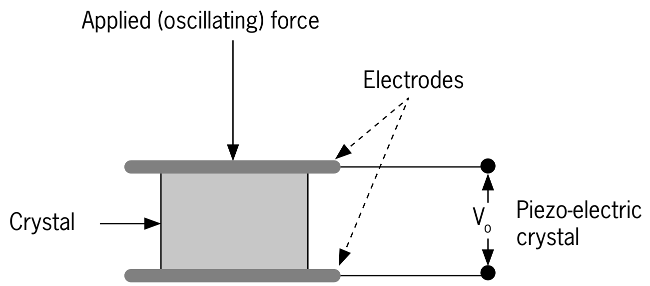
   1.5 Ultrasonic energy behaves similarly to that of electromagnetic energy in that reflection and refraction can take place when energy is transferred from one medium to another medium that has a different density. Absorption and attenuation of ultrasonic energy can also take place when it is propagated over a distance. This attenuation is inversely proportional to the density of the propagating medium and also inversely proportional to the square of the distance travelled.

1.6
- Material flaw detection
- Cleaning and mixing
- Machining, soldering and welding
- Thickness gauging

1.7 Ultrasonic transducers and ultrasonic sensors are devices that generate or sense ultrasound energy.

2. Ultrasonic energy behaves similarly to that of electromagnetic energy in that reflection and refraction can take place when energy is transferred from one medium to another medium that has a different density. Absorption and attenuation of ultrasonic energy can also take place when it is propagated over a distance. This attenuation is inversely proportional to the density of the propagating medium and also inversely proportional to the square of the distance travelled.

3. It is a crystal placed between two electrodes that will generate an emf when subject to mechanical stress.



This ultrasonic energy can be obtained in one of two methods.

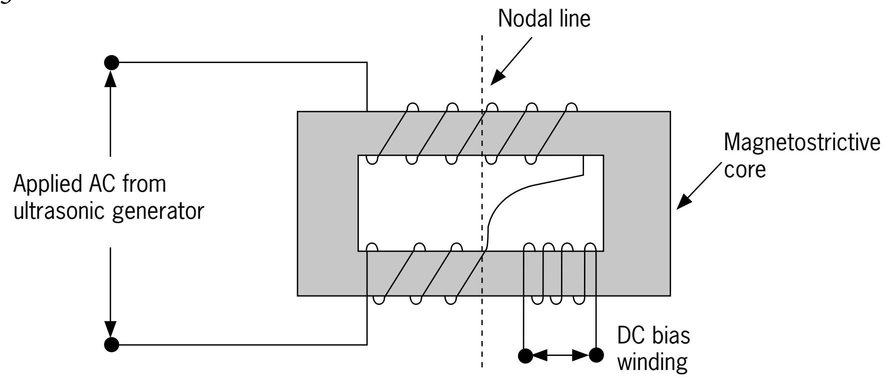
The direct piezo-electric effect: When a crystal is subjected to mechanical stress, it would accumulate an electrical charge that is directly proportional to the magnitude of the applied stress.

The converse piezo-electric effect: When a crystal is subjected to an applied emf, its dimensions will vary by about 0,001% to 0,003% of its original size. The magnitude and polarity of the applied emf will determine the direction and degree of the dimensional variation.

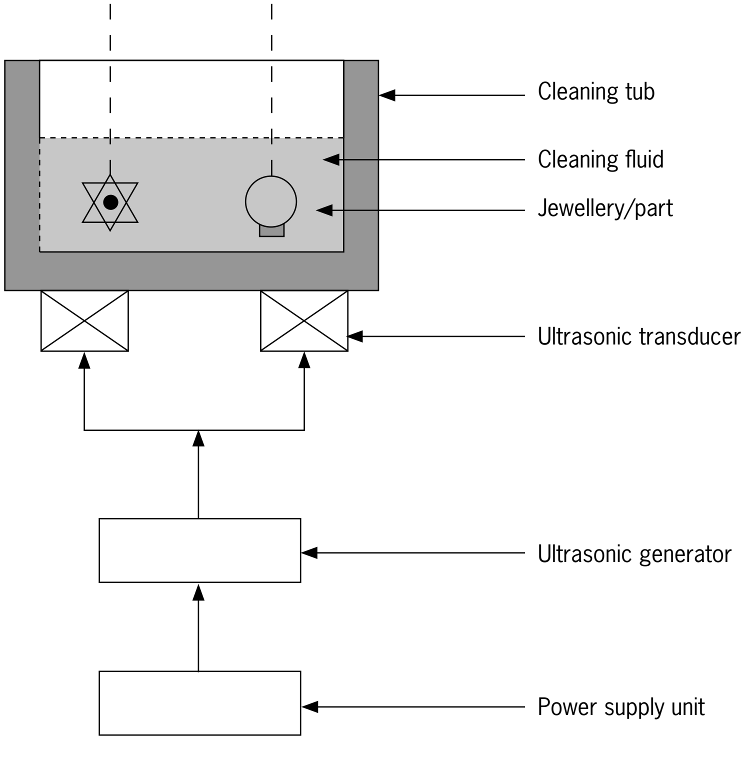
4. 4.1 Certain conductive materials such as cobalt and alloys or iron cobalt are classified as magnetostrictive elements and can have variations in length of ± 0,003% compared to their original length when subjected to oscillating magnetic or electric fields.

4.2 The magnetostrictive core has two windings: a DC winding, which is used to control the AC winding, if required. The function of the DC bias in an ultrasonic application is to allow the magnetostrictive core to expand along the nodal line so that the core can oscillate (expand and contract) when an AC signal from the ultrasonic generator is applied.

4.3



5. 5.1



5.2 The cleaning apparatus consists of a cleaning tub with ultrasonic transducers attached to the sides and/or bottom of the tub and is normally attached to the outside of the tub.

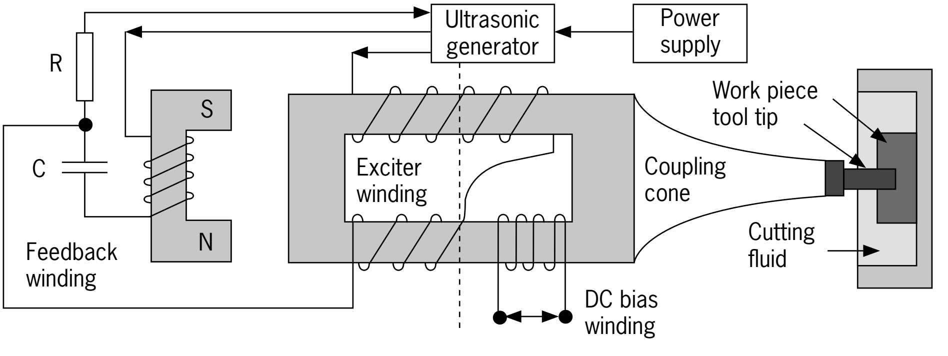
The parts to be cleaned can either be placed in a perforated container or suspended in the cleaning fluid. The ultrasonic transducers are activated to cause cavitation and in this manner the dirt is removed from the objects. When a large number of objects need to be cleaned, a conveyer belt can be used.

5.3 Cavitation is a process or phenomena that takes place when ultrasonic energy is generated within a liquid. This is obtained when alternating regions of high and low pressures generate vapour voids that form as pressure drops and collapse as pressure increases. When these voids collapse, high-energy shockwaves are generated, which can then be used to dispense dirt on objects when immersed in the liquid.

5.4
- It must be a fluid normally used to dissolve dirt.
- It must hold the dirt in suspension to prevent re-depositing.
- It must have good wetting properties to ensure thorough physical contact with the surface(s) to be cleaned.
- It must have a low viscosity so as to minimise resistance to the propagation of the ultrasonic energy.
- It must have a low surface tension to ensure cavitation.

5.5 To mix liquids that:
- do not mix easily; and
- are dangerous to handle.

6. 6.1 Materials that are too hard and brittle to be machined by conventional methods can be shaped by means of abrasion.

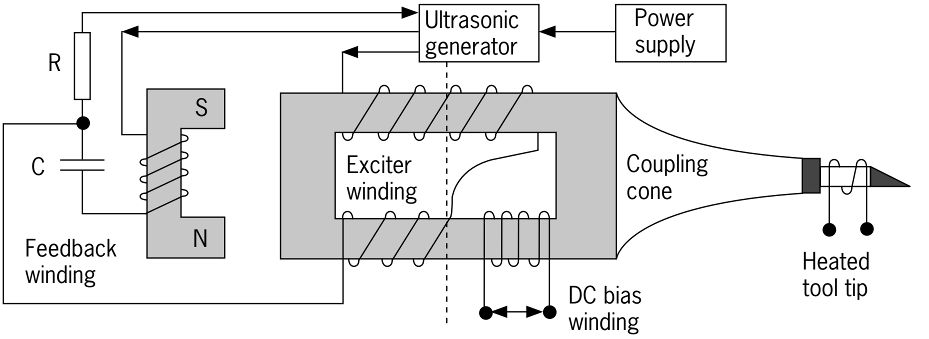


The machine consists of a magnetostrictive core that is clamped at the nodal line. At one end of the core, an exponential cone is attached, which carries the cutting tip. The other end of the core is coupled to the feedback circuit in order to sustain oscillations.

6.2 The tooltip is manufactured from a soft material such as copper and gives the added advantage that the tip can be manufactured into any desired shape for reproduction in the workpiece.

6.3 The workpiece is immersed in a suspension of abrasive powder and water, forming a grind-like paste. The vibration from the ultrasonic generator is transferred through the tip into the paste. The abrasive particles of the paste are agitated in the direction of the vibration and a cavity is gradually sunk into the workpiece, corresponding to the shape of the tooltip.

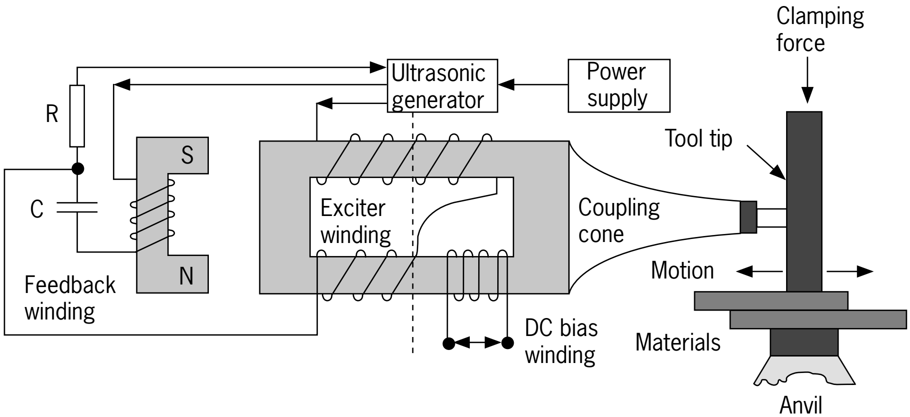
7. The vibration of the soldering tip causes cavitation within the molten solder, which results in the oxide film being broken up. The surface that is covered by the molten solder prevents re-oxidation and the solder is able to flow to all the required sections.



8. The vibration of the soldering tip causes cavitation within the molten solder, which breaks up the oxide film. As the surface is now covered with solder, re-oxidation is prevented, and the solder is able to flow onto the aluminium surface and set before the oxide film develops.
9. Using an ultrasonic machine provides a method where the oxide layer is broken down by the vibration energy, causing deformation of the surfaces on both sides to form a solid bond.

   The metal plates are clamped between the tooltip and a supporting anvil. The vibrating energy is then transferred through the tooltip and is given just enough pressure to keep the plates in close contact for a predetermined period. The pressure applied will vary from a few grams for light materials to several kilograms for heavier materials. The main advantage of not generating heat, as is the case with conventional arc welding, is that the materials do not become deformed.

   Uses for this method of welding include the following:
   - Electric and electronic components
   - Metallic foil splicing
   - Welding of aluminium sheets



10. In medicine, ultrasonics is used as a diagnostic tool to destroy diseased tissue and repair damaged tissue. Ultrasonic waves have been employed to treat various types of rheumatoid arthritis, gout and muscular injuries and to destroy kidney stones. As a diagnostic tool, ultrasonics is often more revealing than X-rays, which do not prove as useful in detecting the subtle density differences found in certain forms of cancer. It is also widely used to produce images of the foetus during pregnancy. When ultrasonic waves are passed through a tissue, the waves are reflected in varying degrees, depending on the density and elasticity of the tissue.

MODULE

4 *X-rays and radio activity*

By the end of this module, students should be able to:

- describe and reproduce the properties of the electromagnetic spectrum;
- discuss the generation, characteristics and power supply unit of a rotating anode X-ray tube with the aid of fully labelled diagrams and practical applications; and
- draw and describe the operating principle of selected particle detectors as applied to the electronics industry.

Exercise 4.1 — SB page 51

1. 1.1 Electromagnetic waves have both electric and magnetic components. Electromagnetic radiation can be arranged in a spectrum that extends from waves of extremely high frequency and short wavelengths to waves with extremely low frequency and long wavelengths. Visible light is only a small part of the electromagnetic spectrum. Electromagnetic radiation is energy waves produced by the oscillation (or acceleration) of an electric charge.

   1.2 The electromagnetic spectrum consists of gamma rays (shortest wavelength), hard and soft X-rays, ultraviolet radiation, visible light, infrared radiation, microwaves and radio waves (longest wavelength).

   1.3 The visible light spectrum is the segment of the electromagnetic spectrum that the human eye can view. Typically, the human eye can detect wavelengths from 380 to 700 nanometres.

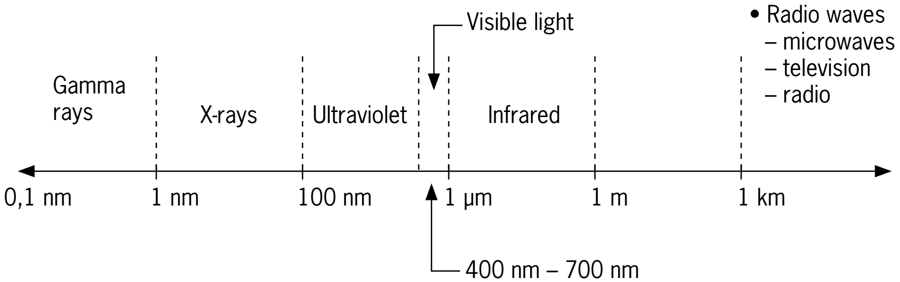
   An X-ray, or X-radiation, is a penetrating form of high-energy electromagnetic radiation. Most X-rays have a wavelength ranging from 10 picometers to 10 nanometres.

   Radio waves are electromagnetic radiation with wavelengths in the electromagnetic spectrum longer than infrared light. Radio waves have frequencies as high as 300 gigahertz to as low as 30 hertz.

1.4 • Electromagnetic radiation
- Conduction
- Convection (transfer of heat)

1.5 • Electromagnetic waves are transverse in nature, as they propagate by varying the electric and magnetic fields such that the two fields are perpendicular to each other.
- Accelerated charges produce electromagnetic waves.
- Electromagnetic waves have a constant velocity in vacuum, equal to 3×10^8 ms^{-1}.
- Electromagnetic wave propagation does not require any material medium to travel.

2.



3.
- In conduction, heat transfer takes place between objects by direct contact.
- In convection, the heat transfer takes place in a fluid.
- In radiation, heat transfer occurs through electromagnetic waves without involving particles.
- The heat transfer occurs through a heated solid object.

4. 4.1 X-rays are produced through radiation of very short wavelengths and therefore have a very high frequency. Because energy is proportional to frequency ($\omega = h \times f$ joules), it follows that X-rays consist of high-energy photons that are also very penetrating due to the short wavelength of the radiation.

4.2 The shorter the wavelength of the X-ray, the greater its energy and its penetrating power. Longer wavelengths, near the ultraviolet-ray band of the electromagnetic spectrum, are known as soft X-rays. The shorter wavelengths, closer to and overlapping the gamma ray range, are called hard X-rays.

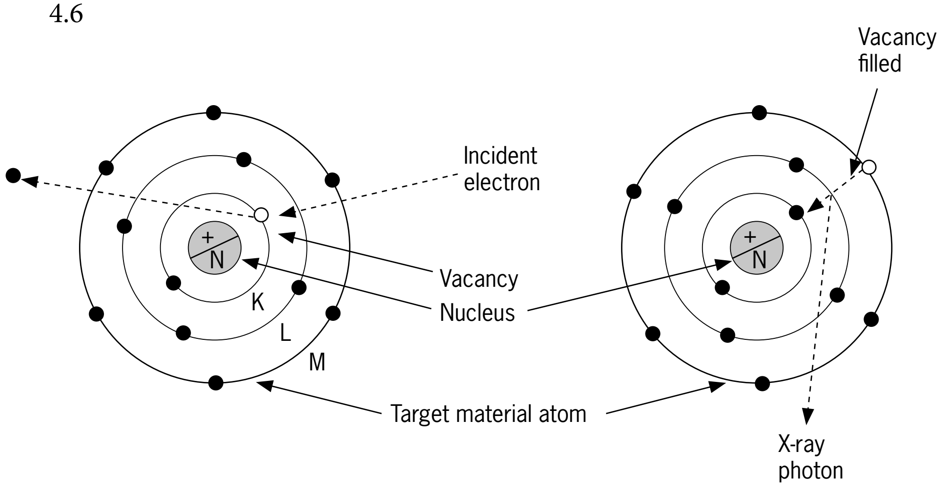
4.3 Both light and X-rays are produced by transitions of electrons that orbit atoms: light by the transitions of outer electrons and X-rays by the transitions of inner electrons.

4.4 X-rays are produced (generated) whenever a beam of high-velocity electrons strikes a material object. Most of the energy of the electrons is lost in heat. The remainder of the energy produces X-rays by causing changes in the target's atoms as a result of the impact. The X-rays emitted can have no more energy than the kinetic energy of the electrons that produce them.

4.5 The X-ray energy that is emitted is proportional to frequency and is expressed by $W = h \times f$.

X-rays therefore consist of high-energy photons and are highly penetrating due to the short wavelength of the radiation.

4.6



X-rays are emitted by atoms that have a high atomic number when they are bombarded by a high-energy electron beam. The X-rays emitted result from two separate effects:

- The deceleration of high-speed electrons as they pass through matter, which results in the emission of energy in a continuous spectrum
- The inner-orbital ionisation of atoms by collision, which results in energy bursts at specific frequencies and wavelengths due to higher-energy electrons moving to vacant spaces in the ionised inner shells and radiating their excess energy in well-defined quanta.

4.7 Absorption of X-radiation by any substance depends on its density and atomic weight. The lower the atomic weight of the material, the more transparent it is to X-rays of given wavelengths.

4.8 The power supply for any X-ray generator consists of a low-voltage alternating current filament transformer, and for the high voltage side either an AC transformer or an AC transformer followed by a suitable high-voltage rectifier system.

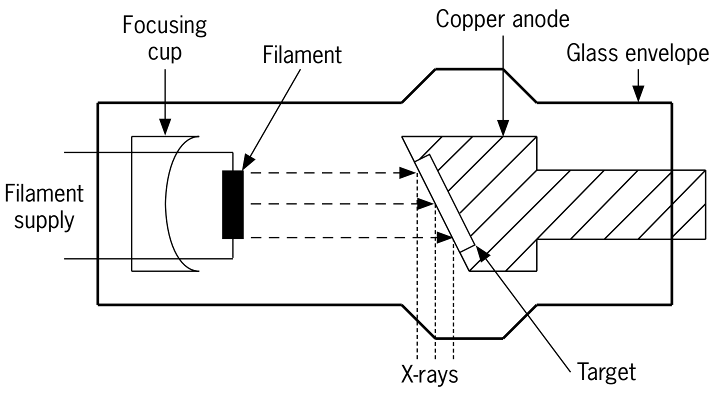
The low-voltage filament transformer must be capable of supplying a high current at low voltage, as the filaments of X-ray tubes are operated under saturated current conditions so that changes in the supply voltage do not affect the electron emission from the filament.

4.9 An auto-transformer is included in the power supply unit so that the unit can be adjusted for variations in the mains supply voltage before exposure is made, as the amount of X-radiation produced is directly proportional to the voltage applied to the anode.

4.10
- Scientific research, i.e. development of quantum mechanics
- Engineering, i.e. non-destructive inspection (NDT or NDI), automatic inspection and testing
- Medicine, i.e. radiographs (X-ray photographs), CAT scans, etc.

4.11 As the bulk of the energy produced by an X-ray tube is in the form of heat and the effectiveness of the tube is furthermore proportional to both the applied voltage and the atomic number of the target material, it becomes evident that an X-ray tube is rather an inefficient device. In fact, in general, less than 2% of all the energies generated in an X-ray tube correspond to actual X-rays emitted. The factors that determine the efficiency of an X-ray are the following:
- The atomic number of the material
- The number of electrons penetrating the target
- The applied (accelerating) voltage

5. X-rays are high-energy electromagnetic waves, but visible light is medium-energy electromagnetic waves. The visible spectrum is very narrow compared to the X-ray spectrum. X-rays can penetrate the human body, but visible light is not capable of doing that.
6. The bulk of the energy produced by an X-ray tube is in the form of heat and the effectiveness of the tube is furthermore proportional to both the applied voltage and the atomic number of the target material.
7. The wavelength and frequency of light are closely related. The higher the frequency, the shorter the wavelength.
8.

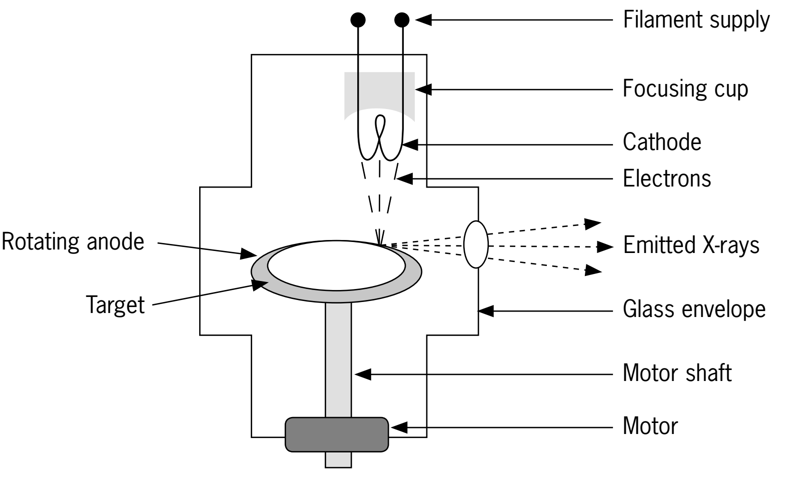


The X-ray tube consists of a helical tungsten filament cathode and a copper anode that contains the target in the form of an insert in the copper anode material. The target is normally from a material such as tungsten or molybdenum.

This target is embedded in the copper anode, as the copper has the ability to act as a heat exchanger in order to remove a large amount of heat that is generated along with the X-rays. The target material can be any of those mentioned and can be changed so as to obtain different X-rays' spectra for special and/or different applications. The glass tube is evacuated and outgassed to a high degree so that the application of high potentials, ± 100 kV between the anode and the cathode, will not produce gas within the tube.

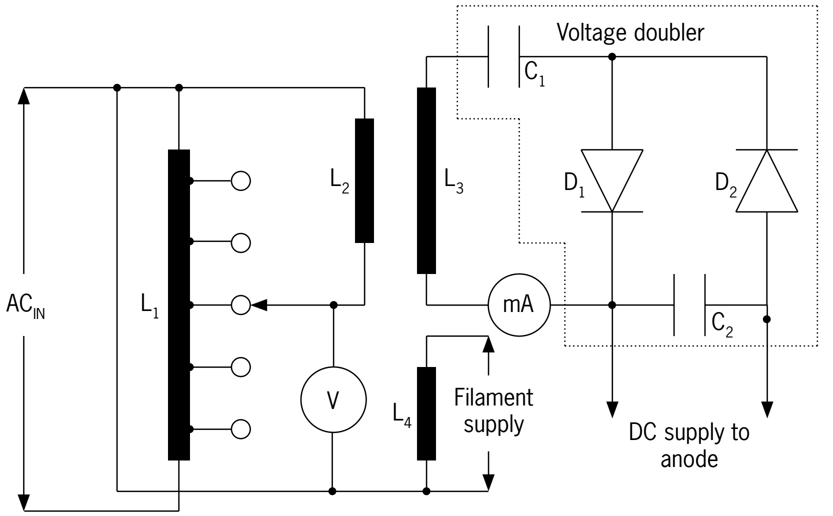
The whole tube could also be immersed in oil in a shockproof housing, thereby also providing electrical insulation and aiding in cooling of the tube. The electrons emitted by the hot filament (cathode) travel towards the target in the anode in a narrow beam at high speed caused by the high anode potential with respect to the cathode. When these electrons strike the target material embedded in the anode, electrons in the target material are dislodged from their normal energy levels.

9.



With the anode at a high positive potential with respect to the cathode, the electrons emitted from the hot filament are accelerated in a narrow beam and travel towards the tungsten target at high speed. When these electrons collide with the atoms of the target, they dislodge electrons from their normal energy levels in the target material. These excited electrons gain energy and move into higher energy levels, producing mainly heat. Only electrons releasing energy by falling from a higher energy level to the inner-most energy level (K-shell) will produce X-ray energy. In order to move from a higher energy level to a lower level, a vacancy (hole) should exist that was caused by the electrons in the target material being excited from the high-velocity electrons supplied by the cathode.

10.



11. i) $\lambda = \dfrac{1{,}2395 \times 10^{-6}}{V}$

$= \dfrac{1{,}2395 \times 10^{-6}}{150 \times 10^{3}}$

$= 8{,}236 \times 10^{-12}$ m

ii) $\eta = 1{,}4 \times 10^{-9} \times Z \times V \times 100$

$= 1{,}4 \times 10^{-9} \times 42 \times 150 \times 10^{3} \times 100$

$= 0{,}88\%$

iii) I = number of electrons × q

$= 1{,}5 \times 10^{19} \times 1{,}602 \times 10^{-19}$

$= 2{,}403$ A

X-ray power $= P_T \times \eta$

$= 360{,}45 \times 10^{3} \times 8{,}8 \times 10^{-3}$

$= 3{,}172$ kW

$P_T = V \times I$

$= 150 \times 10^{3} \times 2{,}403$

$= 360{,}45$ kW

P dissipated $= P_T$ – X-ray power

$= 360{,}45 - 3{,}172$

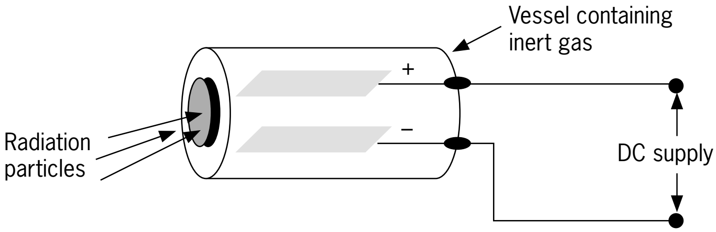
$= 357{,}28$ kW

12. Its purpose is the deceleration of high-speed electrons as they pass through matter, resulting in the emission of energy in a continuous spectrum, as well as inner-orbital ionisation of atoms by collision, resulting in energy bursts at specific frequencies and wavelengths due to higher-energy electrons moving to vacant spaces in the ionised inner shells and radiating their excess energy in well-defined quanta, thereby representing wavelengths that depend only on the structure of the target atoms.

13. Some energetic electrons lose all their acquired energy in a single process, emitting photons corresponding to the energy lost when these electrons ionise target atoms by collision in the inner shells. The vacancies created are filled by outer-shell electrons falling in from the higher energy levels and emitting photons with energy equal to the difference between the respective orbital energy levels.

Exercise 4.2 **SB page 58**

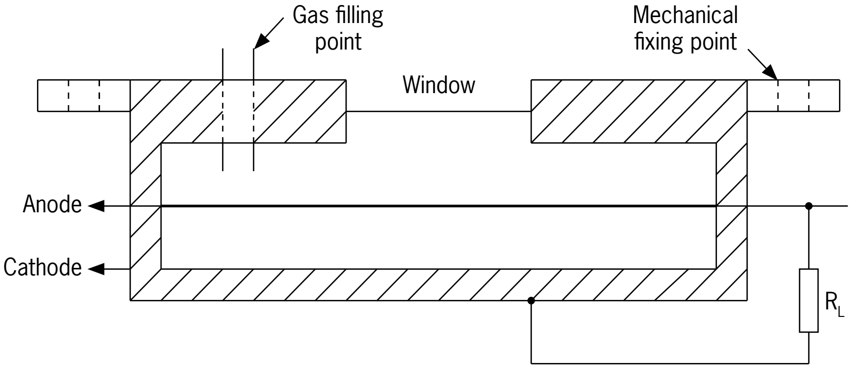
1. 1.1



The ionisation chamber essentially consists of a closed vessel containing a gas and is equipped with two electrodes at different electrical potentials. The electrodes, depending on the type of instrument, may consist of parallel plates or coaxial cylinders, or the walls of the chamber may act as one electrode and a rod inside the chamber acts as the other.

When radiation or ionising particles enter the chamber, they ionise gas between the electrodes. The ions that are thereby produced migrate to the electrodes of opposite sign (negatively charged ions move towards the positive electrode and vice versa), creating a current that may be amplified and measured directly with an electrometer – an electroscope equipped with a scale – or amplified and recorded by means of electronic circuits.

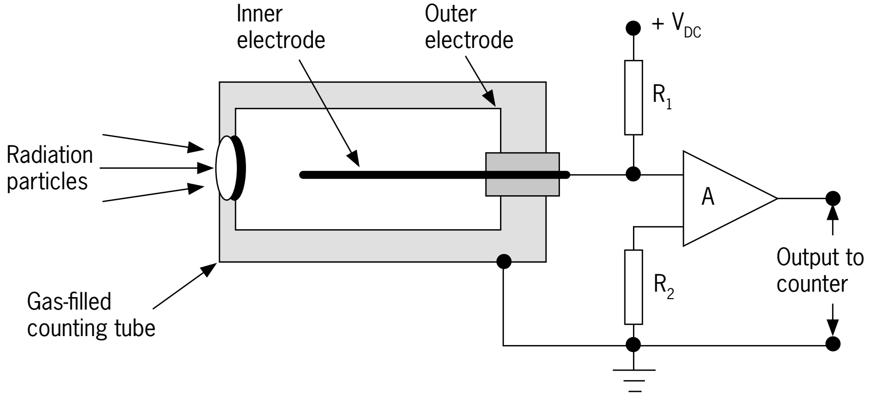
1.2



Photons penetrate the window and pass into the gas, inside which interactions with the gas atoms result in the creation of a number of ion pairs, namely electrons and partially ionised gas atoms. Anodes in the detector volume are held at a positive potential with respect to the rest of the detector. The anodes are usually thin metal wires, and their electric field causes the electrons to drift towards the anodes where the field strength is highest.

The energy of the electrons increases and collisions with other gas atoms cause further ionisation, producing more electrons. These secondary electrons themselves drift and acquire enough energy to cause further ionisation and electrons, and so a large cloud of electrons arrives at the anode in a process known as an avalanche. The quantity of charge produced in the avalanche is great enough to be detectable across a load resistor.

1.3



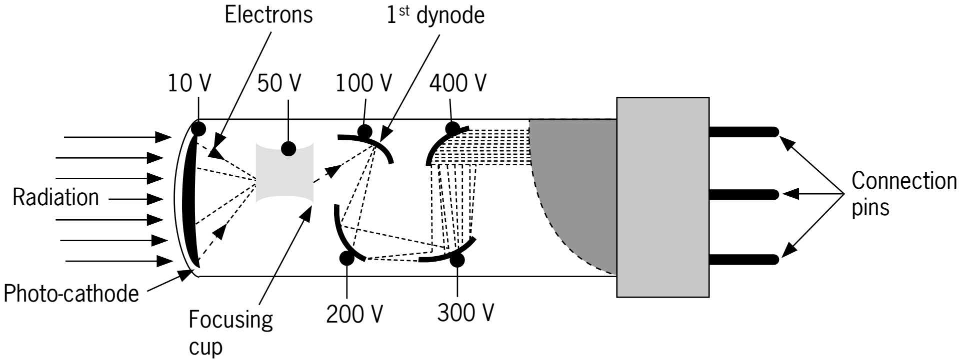
Radiation moves about randomly outside the detector tube. Some of the radiation enters the window at the end of the tube. When radiation collides with gas molecules in the tube, it causes ionisation: some of the gas molecules are turned into positive ions and electrons.

The positive ions are attracted to the outside of the tube. The electrons are attracted to a metal wire running down the inside of the tube maintained at a high positive voltage. As the electrons head for the wire, some of them collide with other gas molecules, splitting them into ions and more electrons. This causes a type of chain reaction in which even a single particle of radiation can produce avalanches of electrons in rapid succession; this process is known as a Geiger discharge.

Many electrons travel down the wire, causing a burst of current in a circuit connected to it. The electrons cause a meter needle to deflect and, if a loudspeaker is connected, you can hear a loud click every time particles are detected. The number of clicks you hear gives a rough indication of how much radiation is present.

Before the counter can detect any more radiation, it needs to be restored to its original state through a process called quenching, which cancels out the effects of the Geiger discharge. Sometimes this is achieved by having a second gas, called a quenching gas, often a halogen inside the tube, or it can be done using an external circuit with a very large resistance.

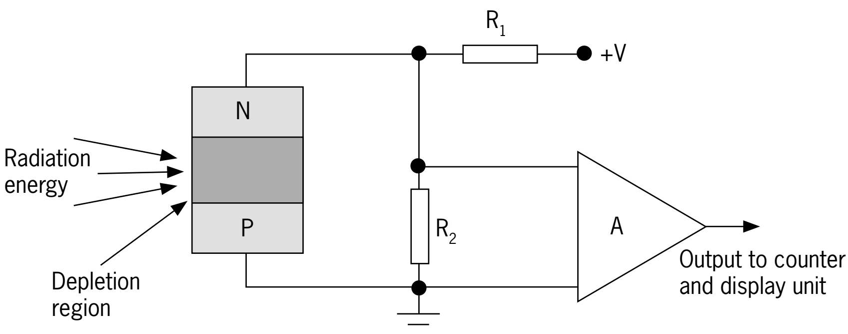
1.4



A photo-multiplier uses the principle of fluorescence, where radiation with energies equal to or greater than those of visible photons causes certain phosphorous materials to emit a visible glow when radiation energy excites the phosphorous materials.

The light energy emitted by the phosphorous material is allowed to fall onto the cathode of the photo-multiplier. Electrons, proportional to the incident incoming radiation, will be liberated from the surface. Each liberated electron is accelerated to an intermediate anode or dynode that is also coated with an emissive layer where several electrons are liberated for each incident electron. These liberated electrons are now further accelerated towards further dynodes, each at a higher potential than the previous one. This process is repeated until an avalanche of electrons are received at the final anode, which then generates a pulse proportional to the energy of the incoming radiation. A multi-channel analyser will now accept these generated pulses, sorts them into sizes and keeps track of the number of pulses. The energy spectrum of the incoming radiation entering the analyser can therefore be recorded and studied.

1.5



The semiconductor detector consists of a reverse-biased PN junction. When radiant energy is allowed to penetrate the depletion region, electron-hole pairs are generated and momentarily increase the conduction across the junction and produce a pulsed output that is amplified and counted.

2. The ionisation chamber is the simplest of all gas-filled radiation detectors and is widely used for the detection and measurement of certain types of ionising radiation, e.g. X-rays, gamma rays and beta particles.

   The proportional counter is a type of gaseous ionisation detector device used to measure particles of ionising radiation. The key feature is its ability to measure the energy of incident radiation by producing a detector output pulse that is proportional to the radiation energy absorbed by the detector due to an ionising event; hence the detector's name. It is widely used where energy levels of incident radiation must be known, such as in the discrimination between alpha and beta particles and accurate measurement of X-ray radiation dose.

   The Geiger counter is an instrument used for detecting and measuring ionising radiation. Also known as a Geiger-Mueller counter (or Geiger-Müller counter), it is widely used in applications such as radiation dosimetry, radiological protection, experimental physics and the nuclear industry. It detects ionising radiation such as alpha particles, beta particles and gamma rays using the ionisation effect.

3. The biological radiation effects, if observed when ionising radiation interacts with living tissue by transferring energy to molecules of cellular matter, is described as follows:

   The cellular function may be temporarily or permanently damaged as a result of such interaction, or the cell may be destroyed. The severity of the injury depends on the type of radiation, the absorbed dose, the rate at which the dose was absorbed and the radiosensitivity of the tissues involved. The effects are the same, whether from a radiation source outside the body or from material within. Precautionary measures when working with radioactive substances include the following:

   - Designate and label areas for working with radioactive material.
   - Label all containers with a radioactive material label and specify the isotope.
   - No eating, drinking or smoking is allowed in the laboratory.
   - No mouth pipetting of radioactive material is allowed.
   - Use spill trays and absorbent covering.
   - Use fume hoods for handling potentially volatile material.
   - Use a glove box for handling large quantities of volatile material.
   - Wear a laboratory coat, disposable gloves and laboratory safety glasses.
   - Use gloves appropriate for the chemicals to be handled.

4. Magnetic fields affect the gain of the photomultiplier tube because some electrons may be deflected from their normal path between stages and therefore never reach a dynode or, eventually, the anode. In scintillation counting applications, this effect may be disturbing and mu-metal magnetic shields are therefore often placed around the photomultiplier tube.

5. The semiconductor detector can be much smaller and the operating voltage is also lower.
6. Supply voltage = V per stage × number of stages

$$\begin{aligned}\text{Supply voltage} &= \text{V per stage} \times \text{number of stages}\\ &= 100 \times 12\\ &= 1\,200 \text{ V}\end{aligned}$$

$$\begin{aligned}A &= \int^{n}\\ &= 2^{12}\\ &= 4\,096\end{aligned}$$

$$\begin{aligned}\text{Sensitivity} &= \text{cathode sensitivity} \times A\\ &= 20 \times 10^{-6} \times 4\,096\\ &= 81{,}92 \text{ mA/lm}\end{aligned}$$

$$\begin{aligned}I_M &= \frac{V}{R}\\ &= \frac{1}{250}\\ &= 4 \text{ mA}\end{aligned}$$

$$\begin{aligned}\text{Maximum safe illumination} &= \frac{I_M}{\text{sensitivity}}\\ &= \frac{0{,}4 \times 10^{-3}}{81{,}92 \times 10^{-3}}\\ &= 48{,}883 \text{ m/lm}\end{aligned}$$

MODULE

5 Automatic inspection, testing and NDT

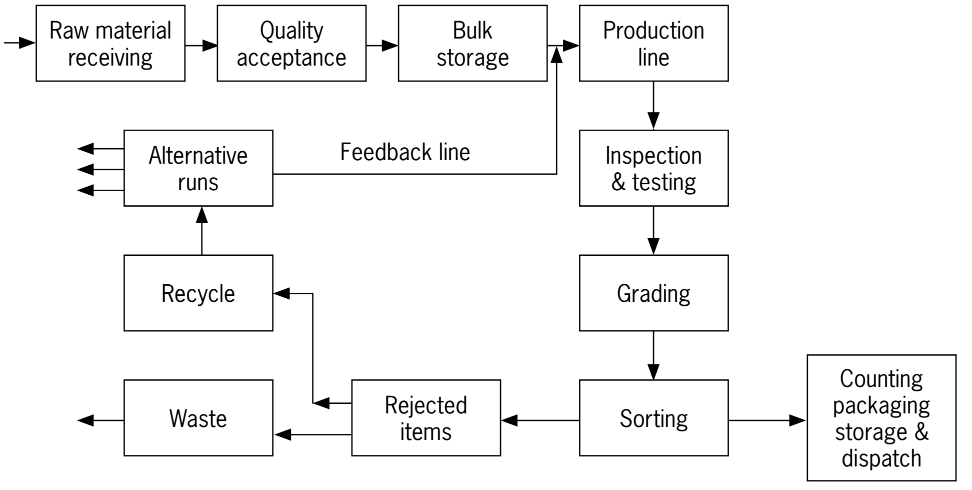
By the end of this module, students should be able to:

- describe selected automatic inspection, testing and grading devices with the aid of suitably labelled diagrams as applied to the electronics industry;
- reproduce the operating principle of a typical metal detector with the aid of labelled block diagrams and give detailed explanations as applied to timber, mining, security and other related industries; and
- draw and describe selected NDT procedures employing an X-ray tube and ultrasonics as applied to the medical, welding and various other related industries.

Exercise 5.1 SB page 71

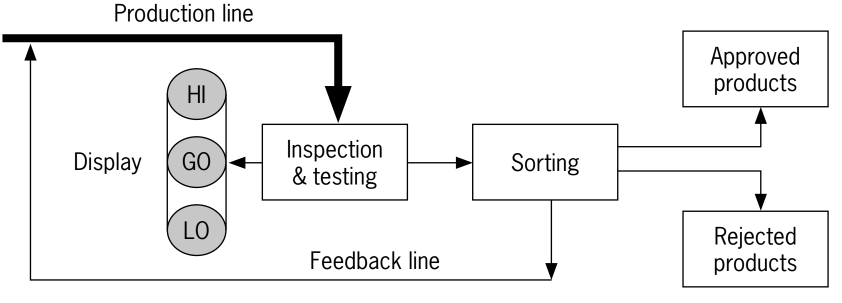
1. - Perform as many processes as possible in a single stage
   - Consider the availability and the economic factors of labour versus automation
   - Consider the character of the final product, such as size, mass, structure, density, colour, texture, etc.
   - Consider the shelf life and demand of the product
   - Consider the packaging and storage methods
   - Consider the rejection cycles of the product; that is, if it is rejected in the mainstream production line, what are the alternatives?

2.



Material required for the production process is received and quality assurance is carried out to ensure that the correct material and quality have been received before the material is stored. From the storage facility the material is moved to the production line for working. Material may also be received from the alternate run, which could be a different production process or products that did not conform to the specifications during the initial production process. After the production line process, the manufactured items are inspected, tested and graded in order to comply with the design specifications. All the items are then sorted and split according to rejection or acceptance. If accepted, the items are counted, packaged, stored if required and then dispatched. Rejected items are sorted as waste if they do not conform to specifications, or they are sorted to be recycled and will then move to the alternate run.

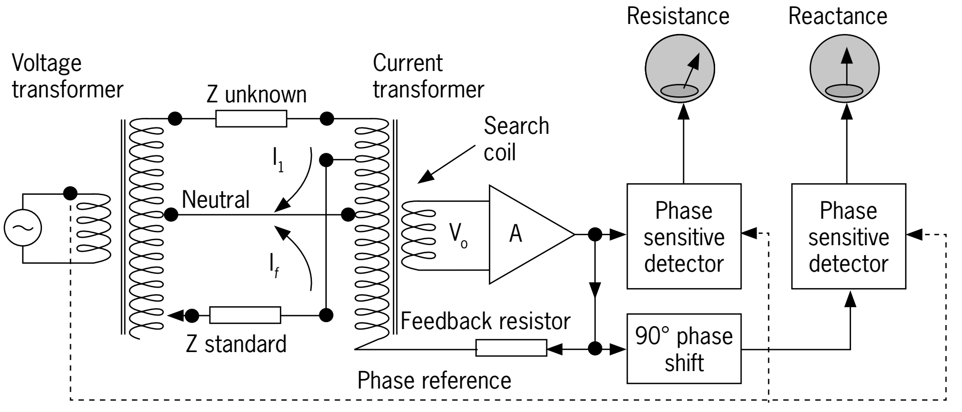
3.



The two-way system simply accepts or rejects items, whereas the three-way system sorts the rejected items into those that can be returned to the production line for reworking and those that have to be scrapped. A practical example here would be machined articles that are too large (high), correct dimensions (go) or too small (low). The high rejects could be returned for re-working, while the low rejects would be scrapped.

Should the display be changed to display only a GO/NOGO condition and the feedback line be removed, the three-term system has been converted into a two-term system. Practical applications of typical two-term systems include the various glass and bottling industries, such as Consol Glass, SA Breweries, etc.

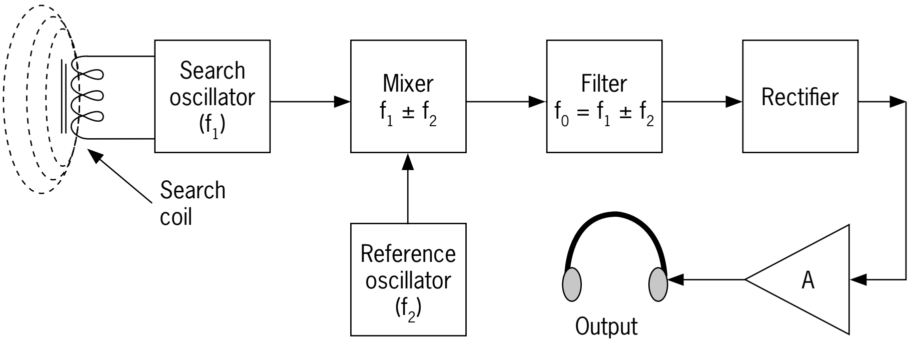
4.



The transformer ratio arm bridge utilises a current balancing technique. One winding on a transformer core receives alternating current (I_1) produced by a known voltage and unknown impedance. A second core winding is driven by a balancing current (I_f) produced by a known variable voltage and known standards of resistance and capacitance. A further core winding acts as a search coil to allow determination of a null in the core flux.

The multi-winding current transformer acts as a current balance detector. The output of the search coil is amplified to provide a bridge-balancing voltage, which is applied through a resistance to the feedback winding. The initial current (I_1) derived from the voltage transformer and unknown impedance is opposed by the feedback current (I_f) and the condition of equality is limited only by the finite (limited) gain to the amplifier and stability considerations. The amplifier output is also connected to two phase-sensitive detectors, one of which incorporates a unity gain, 90° phase shift network. A stable oscillator provides the signal for the bridge and reference voltage for the detectors. Analogue detector voltage outputs are formed, which independently show the amount of variation of the resistive and reactive terms from the value set by the standard impedance and can represent capacitance, inductance or resistance according to the type of bridge network used. Tapped voltage transformers in cascade can provide digital taps feeding impedance standards, in addition to the auto-balance circuit, enabling a wide range of measurements to be made.

5.

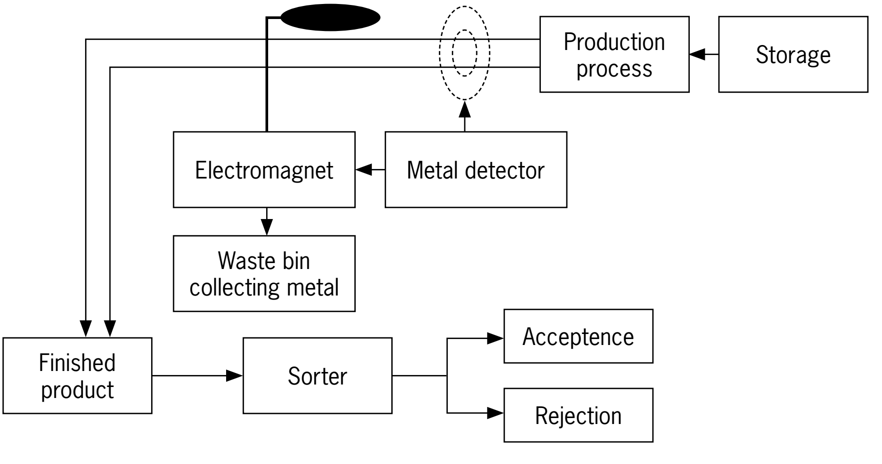


The search coil oscillator, f_1, is initially tuned to the same frequency as the reference oscillator, f_2, when no metal is in the vicinity of the search coil. Because these two frequencies are equal, the output from the mixer and filter circuit is zero ($f_0 = f_1 - f_2$). As soon as metal parts enter the detector range of the search coil, the frequency of the search coil oscillator changes and the output frequency, f_0, is then the difference between the two frequencies, f_1 and f_2. This output is then fed to a rectifier and is amplified. The amplified output signal can then be used to initiate the required action, such as stopping the conveyor belt while sounding the alarm or operating an electromagnet or gate to reject the metal part or contaminated package.

Applications include the following:

- Mining industry
- Processed foods
- Timber industry
- Military
- Security
- Private enterprise

6.



7. Radioactive isotopes are also increasingly used for non-destructive testing. They produce radiation of similar wavelength to X-rays and a piece of radioactive cobalt provides a portable source of radiation that will do the job, as well as the cumbersome X-ray generator.
8. Alpha particles (α):

   This type of radiation emits two protons, and one neutron is emitted as a single charge and is positively charged. The emission causes its atomic number to be reduced by two and its atomic mass by four. These particles have the least penetration and have the following effects:

   - They can be stopped or attenuated by a thin sheet of paper.
   - They are deflected a little by a magnetic field.
   - The range of these particles in air at atmospheric pressure is short and definite.
   - They ionise air along their path, thereby delivering 1 000 times more ions than beta particles for the same distance.

   Beta particles (β):

   In this type of radiation, the nucleus emits a negatively charged particle and converts a neutron into a proton. The atomic number is increased by one, but the mass remains unchanged. Beta particles have the following effects:

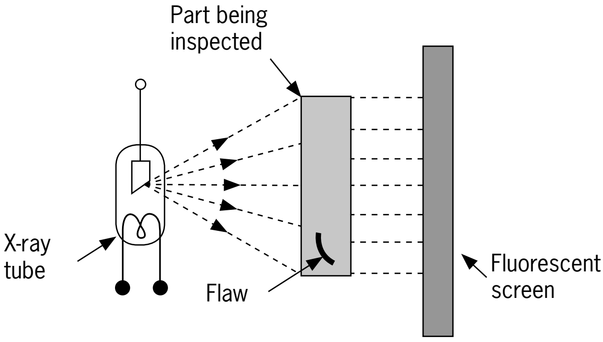
   - They can be stopped or attenuated by ± 1 metre of concrete.
   - They can be stopped or attenuated by ± 1 cm of lead.
   - They have greater penetrating power of materials than alpha particles.
   - They are deflected by magnetic and electric fields.
   - They ionise air along their path, thereby delivering 1 000 times more ions than gamma particles for the same distance.
   - They have a longer range than alpha particles, but their path is not so well defined.

   Gamma particles (γ):

   After each adjustment of charges in the nucleus, excess energy of uncharged high-frequency electromagnetic radiation is emitted in the form of photons with a very short wavelength. Because this energy is very high, it also has a great penetrating depth. Gamma particles have the following effects:

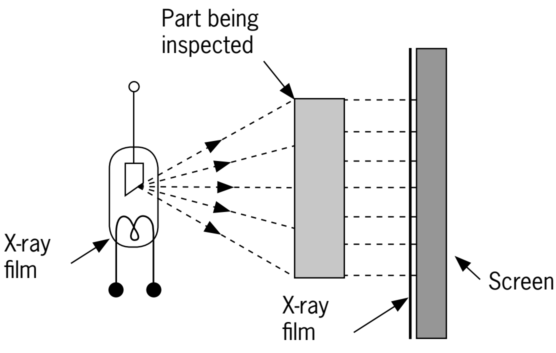
   - They are diffracted by crystals.
   - They can penetrate large thicknesses of metals.
   - They have little ionising power in gas.
   - They are not deflected by magnetic fields.
   - They are stopped by ± 15 cm lead, ± 44 cm thick concrete or ± 1,8 m deep water.
   - They undergo negligible absorption in air.
   - Their intensity varies inversely as the square of the distance between the source and a detector or anything else.

9. a)



The operator views the image on a fluorescent screen in a darkened room. Flaws and cracks in castings, for example, show up as dark lines or shadows on the image of the casting seen on the fluorescent screen. This method is convenient, because if the operator detects a flaw in the object being tested, it is immediately marked and rejected as faulty. However, viewing a fluorescent screen of this type for long periods causes a high degree of eye strain and worker fatigue. Another disadvantage of this method is that the operator is exposed to X-radiation unless great care is taken in making sure that all safety shields are in place and the operator wears lead-lined protective clothing at all times.

b)

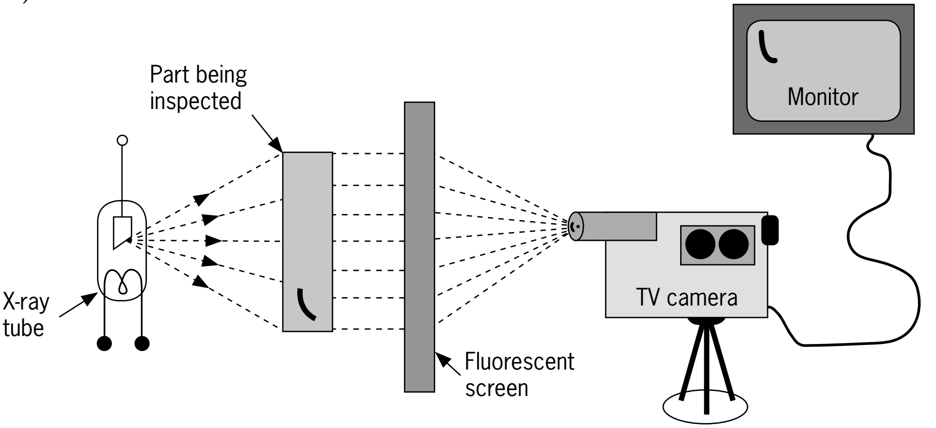


The fluorescent screen is replaced by a film holder containing a sheet of X-ray-sensitive film. The film is positioned with the object being examined between the film and the X-ray tube. After exposure, the film is developed in the usual method of processing photographic film and after being dried, it is viewed by placing it on a brightly lighted sheet of glass in the same manner as done with medical X-ray films.

While slower than the direct viewing method due to the developing time, the indirect method has several advantages:

- A permanent record is made of the flaw in a particular object.
- A skilled operator can view a large batch of X-ray films in rapid succession after they have been exposed and developed by a semi-skilled operator who does not have to interpret the developed films.
- There is less danger of exposure to X-radiation to the operator.

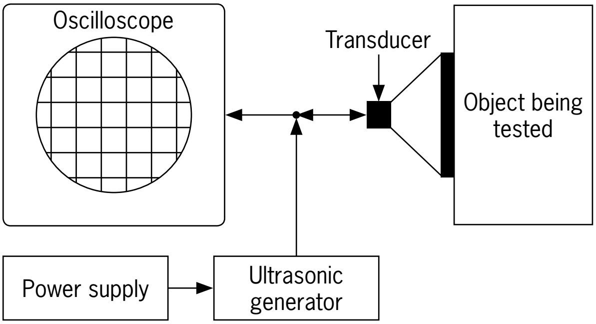
c)



This method employs an X-ray tube and fluorescent screen, but instead of the screen being viewed directly, a remote-controlled television camera is focused onto the image produced by the X-ray tube and the final viewing is done on a monochrome monitor with high resolution. The image on the monitor screen is bright enough to be viewed in daylight, although subdued light is preferred. The bright image on the screen is easy to interpret, making the detection of flaws in castings and other objects easier and quicker. The fact that X-ray films no longer have to be developed speeds up the testing procedure. As remote-controlled television cameras have been used to focus on the object, it has the added advantage of zooming in on any suspect area.

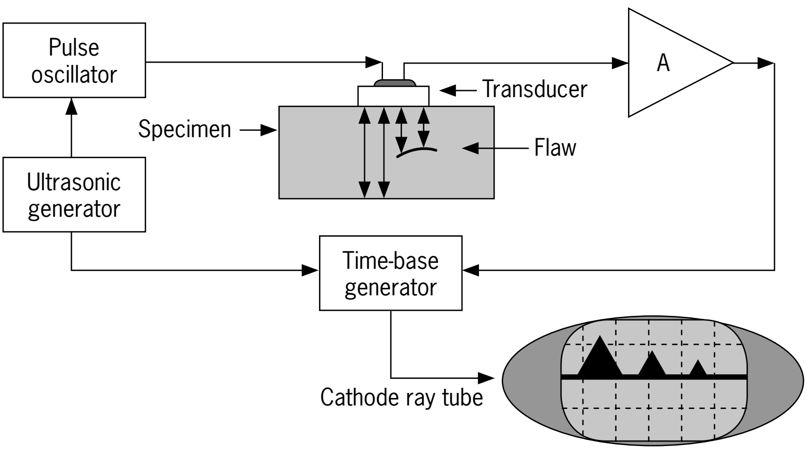
10. Minute cracks, checks and voids, too small to be seen by X-rays, are located using ultrasonic techniques. An ultrasonic test instrument requires access to only one surface of the material to be inspected and can be used with either straight-line or angle-beam testing techniques.
11. Sonar refers to any detection system based on the reflection of sound waves in a medium. A typical sonar system emits ultrasonic pulses with the aid of a transmitter and then 'listens' with the aid of a receiver for reflected pulses from the opposite surface of the specimen.

12. **Sonar method**



A frequency of ± 100 kHz is used and beamed through an object that needs to be tested for a possible flaw. A marker pulse is generated and displayed on the screen of the oscilloscope when a pulse is transmitted. Any possible flaw present in the object being tested will also generate a pulse on the screen of the oscilloscope. A third pulse will be generated on the screen of the oscilloscope when the ultrasonic energy is picked up by the transducer on its return. Ultrasonic testing has the advantage that testing can be done from one side only, as ultrasonic energy will travel around bends and corners.

Resonance method



The resonance method is principally used for thickness measurements when the two sides of the material being tested are smooth and parallel. If the frequency of the ultrasonic wave is such that its wavelength is twice the thickness of a specimen (fundamental frequency), then the reflected wave will arrive back at the transducer in the same phase as the original transmission so that strengthening of the signal, or resonance, will occur.

13. $d = \frac{v \times t}{2}$

$v = \frac{2 \times d}{t}$

$= \frac{2 \times 540 \times 10^{-3}}{276 \times 10^{-6}}$

$= 3{,}913 \times 10^{3}\ m/s^{-1}$

$d = \frac{v \times t}{2}$

$= \frac{3{,}913 \times 10^{3} \times 0{,}21 \times 10^{-3}}{2}$

$= 410{,}8\ mm$

$f = \frac{1}{t}$

$= \frac{1}{276 \times 10^{-6}}$

$= 3{,}623\ kHz$

MODULE

6 Electronic safety devices

By the end of this module, students should be able to:

- describe the terms related to intrinsic safety and electronic safety devices as applied to the various industries;
- draw and describe selected electronic safety devices; and
- draw and describe selected series and shunt protection diagrams with specific reference to intrinsic safety and related industries.

Exercise 6.1 SB page 81

1. The operating controls of a press can be arranged so that an operator has to press two 'buttons', spaced more than a hand-span apart, in order to operate the machine. A device such as this appears to be safe enough and will fulfil the function of ensuring that both the operator's hands are out of the work (danger) area. However, workers have been known to force one control switch in the 'closed' position and then operate the press one-handedly using only the second switch.

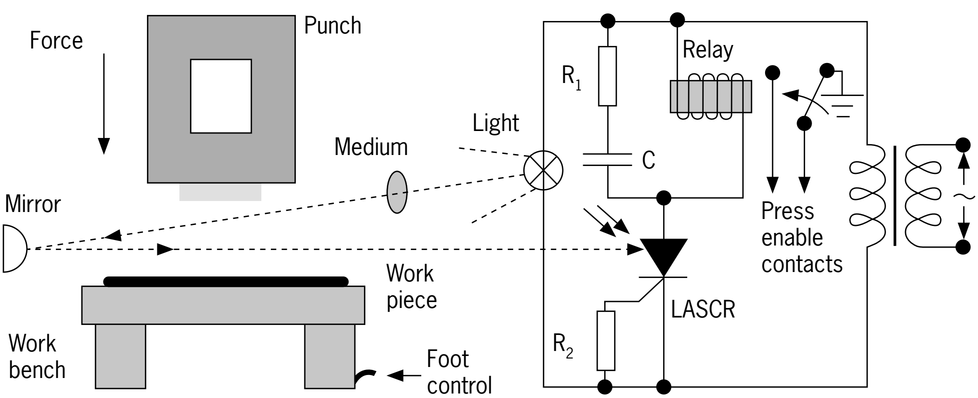
   Electronic safety devices, on the other hand, can be made more sophisticated and independent of manual effort, making it more difficult for personnel operating the machine to by-pass the safety device.
2. To prevent injury or death of workers
3. Inductive or capacitive proximity switches are sensors that are able to detect the presence of nearby objects without any physical contact. When installed as a safety device, the equipment will not function if there is any kind of interference, e.g. human interference, sound, light or infrared radiation. Electromagnetic fields may also be utilised by the sensor to detect a target and then make equipment inoperative.
4. To prevent injury or death of workers

   This could happen through the inadvertent operation of a machine. The objective of electronic and other safety devices is to prevent such inadvertent operation, when, for example, a worker's hands are in the work area of a press, which could crush the worker's hands if the press is closed before the worker was able to remove his/her hands.

To prevent damage to machinery

Safety devices also prevent damage to the machine itself, either through unskilled operation or unsafe conditions being inadvertently introduced. An example of this would be an electrical trip to prevent the flow of welding current in a spot welder unless the cooling water was flowing through the electrodes.

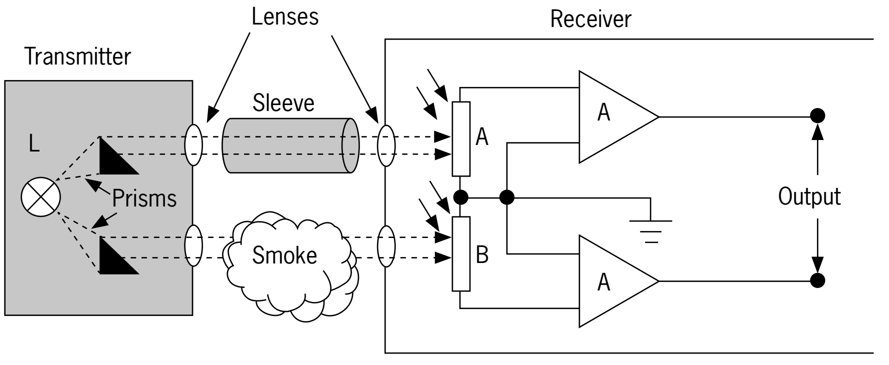
5. There are two important factors to consider when installing any such safety device:
   - The safety device must be fail safe, i.e. if the safety device itself is faulty, operation of the equipment should not be possible.
   - The operator should not be able to by-pass the safety device, if, in his/her opinion he/she can work faster or better without the safety device being operational.
6. The photo-electric device is dependent on light for its operation. Use is made of optical relays, which consist of:
   - a transmitter in the form of a light source;
   - a medium in the form of an optic fibre cable for transmitting the light energy; and
   - an optical receiver such as a photo-diode, photo-transistor, light dependent resistor or any other photo-sensitive component.
7. 7.1



There are many industrial processes involving large hydraulic presses, which can be made safe to operate by means of a beam of light shining across the work area and impinging on a photocell. Should any object, such as the operator's hand, interfere with the beam of light, no light energy falls on the photocell, which will therefore not become conductive. Naturally, the beam of light must not be directed so that the actual work to be pressed interferes with the light path, otherwise the safety system will become operative and it will be impossible to activate the press. The placement of the light beam is a compromise between safety and efficient operations but if it is possible to operate the press with the workpiece and the worker's hand still in the danger area, then an alternative method of making the press safe must be found.

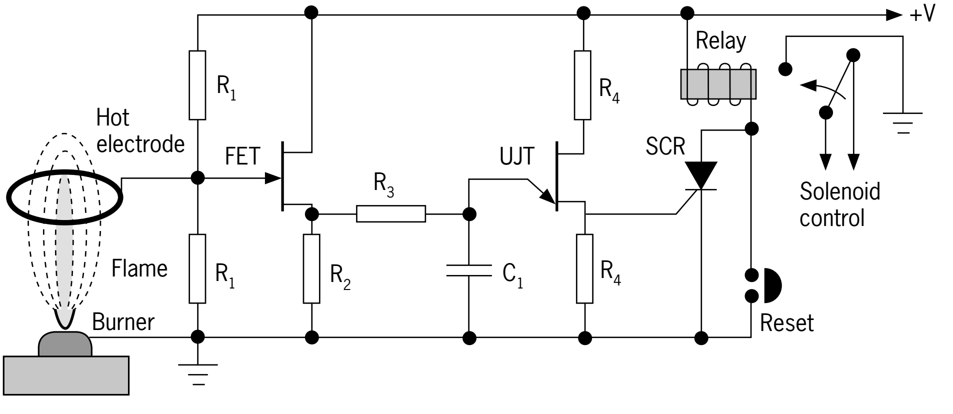
The resistance of the LASCR is lowered when light falls on it, enabling sufficient current to flow through the relay coil, keeping the relay contact in the closed position, thereby allowing the operator to operate the press by closing the manual switch. Should the light beam be interrupted by a human hand or should the lamp of the light source fuse or power to the photo-conductive cell circuit fail, the relay contacts open and it will not be possible to activate the relay control circuit and operate the press.

7.2



Two light-sensitive transducers, A and B, are coupled to differential amplifiers. A light source illuminates both cells by optical means such as optical fibres, mirrors, lenses or prisms. Any brightness variation from the transmitter will therefore act similarly on both channels and therefore cancel at the output. The presence of smoke may be detected if the smoke intercepts the light path transmitted by only one channel. The second light path is protected by screening it off against outer interference with the aid of a protective sleeve. It would therefore be possible to measure the light intensity difference of two light signals transmitted from a single light source and received by two light-sensitive transducers.

7.3



A ring-shaped conductor made of a refractory metal such as tungsten and exposed to the flame acts as the hot (positive) electrode. The grounded burner forms the second (negative) pole. Because flames are high-impedance mediums, the flame probe is connected to a high-impedance input stage, using an FET. The FET input electrode is biased by a high-impedance input voltage divider and grounded by the flame. Should the flame be extinguished, the bias on the source-follower FET-controlled UJT is reduced; the thyristor triggered, energising the relay and opening the burner 'enable' contacts ensuring no further fuel arrival. The entire circuit is DC-fed. A new operating cycle can be started by activating the reset button and thereby applying commutation to the SCR in order to de-energise the safety circuit. With an automatically controlled furnace, the fuel arrives after a preset time and the electric igniter is struck. Should no flame appear, an automatic delay timer will be activated, followed by another ignition trial. After three or four unsuccessful trials, an error signal is produced in the control room.

8. Positive protection is applied to any electronic system, whereby operation of the system can only be justified if all safety conditions are simultaneously fulfilled, expressed by a logic AND function. This implies that the machine and all of its safety devices are connected in serial mode, e.g. the overcurrent protection of a motor. In a microwave oven the magnetron heater is in standby mode (assuming power is connected) so that heat is produced when the on/off switch is activated. This, however, is possible only if the switch is activated and the oven door is closed, the cooling system is on and the tube is preheated.

   Negative protection, on the other hand, is represented by a logic or function. This implies that the machine and its safety device are connected in parallel. An example of this type of protection is the no-voltage protection as applied to industrial DC motors.

9. Intrinsically safe systems are designed to prevent high-energy sources (electrical, chemical or a combination thereof) from placing ignition-capable currents or voltages on wires leading into hazardous areas. Intrinsic safety describes not only the protection technique for electrical wiring and equipment, but also includes the protection technique of flammable atmospheres. Intrinsic safety describes the explosion protection technique that prevents faults in low power measurement and control systems from igniting flammable atmospheres in hazardous areas. Intrinsically safe devices and systems operate by restricting the electrical energy available in hazardous areas to a level at which any spark or hot spots caused by circuit faults or accidents are too weak to produce ignition.

10. Intrinsically safe equipment and wiring are incapable of releasing sufficient electrical energy under normal and abnormal conditions to cause ignition of a specific hazardous atmospheric mixture. Abnormal conditions include accidental damage to any part of:

    - the equipment or wiring;
    - insulation or other failure of electrical components;

- application of overvoltage;
- adjustment and maintenance; or
- other similar conditions.

11. 11.1 Intrinsically safe devices and systems operate by restricting the electrical energy available in hazardous areas to a level at which any spark or hot spots caused by circuit faults or accidents are too weak to produce ignition.

    11.2 Abnormal conditions include accidental damage to any part of the equipment or wiring, insulation or other failure of electrical components, application of overvoltage, adjustment maintenance operation and other similar conditions.

    11.3 Safe energy levels cannot be defined in any simple form. Ignition depends on the specific gas, gas concentration, voltage, current, energy storage elements, contact material, contact size and the speed of opening or closing of contacts.

    11.4 All aspects of the loop must be viewed as components of a loop that are capable of causing or creating an unsafe condition.

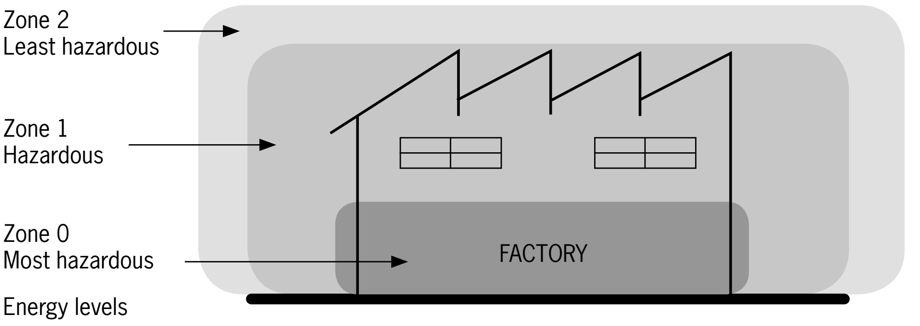
    11.5 The purpose of the margin of safety under the worst possible conditions is to ascertain whether or not each condition constitutes a possibility of ignition. With the use of the actual circuits, faults are introduced by short- or open-circuiting components with the field leads connected to a test apparatus. Measured voltages, currents, inductance, capacitance, etc. are also compared with the results of previous tests under similar circuit conditions.

    11.6 Transformers can be constructed so that they can be presumed not to be fail-unsafe. A typical approach uses a heavy-grounded shield between primary and secondary or equivalent constructions. Certain bodies specify that such transformers must pass a 1-minute insulation test at 1 000 V plus twice the rated voltage after a burnout test. The burnout test is conducted with all secondary windings short-circuited and primary windings at rated voltage for 6 hours or until burnout, whichever occurs first.

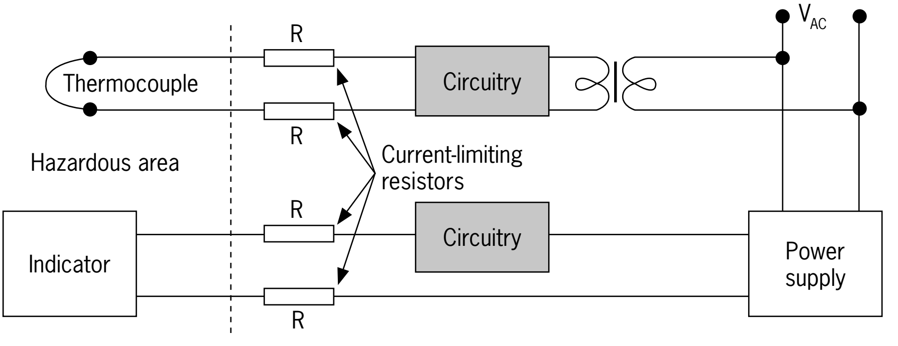
    11.7 Components can be encapsulated to prevent accidental touching of components against other parts of the circuit. However, when such encapsulation is done, one must ensure that the circuit that is encapsulated is in proper working condition, as it cannot be repaired after encapsulation.

12. - Circuit analysis
    - Evaluation
    - Construction review
    - Transformer construction
    - Isolation by encapsulation

13.

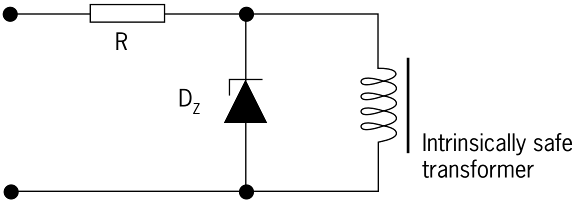


14. 14.1



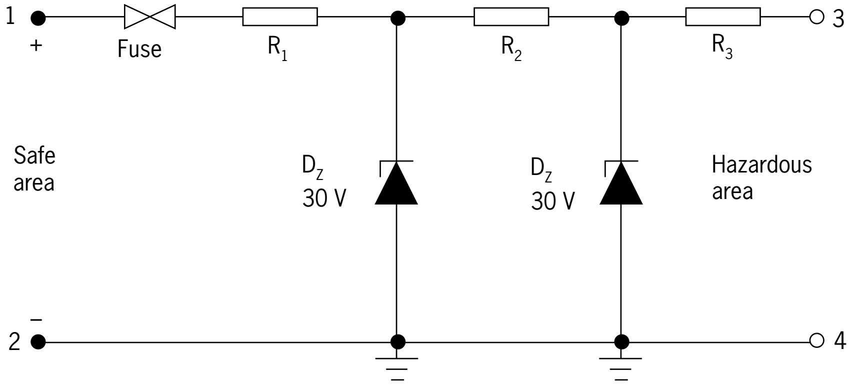
Current-limiting resistors are commonly used as a form of series protection in restricting the amount of current reaching hazardous areas.

14.2



In most cases, two or more diodes are required. Failure of a diode is not usually self-revealing and protection could be lost because of a single failure. When two or more diodes are used, the mounting should be so robust that there is a zero probability that all diodes could be made inoperative by a single wiring or connection failure.

14.3



It is so designed to ensure that when the mains supply is applied to terminal 1 (zone 2 area), the voltage and current at terminal 3 (zone 0 area) are safe.

Devices to be connected to the safe terminals (3 and 4) must be approved for use with the barrier protection device. If the barrier is designed to limit against full mains supply power, then the equipment in the non-hazardous area may be selected, connected and intermingled without regard to safety in the field circuits.

MODULE

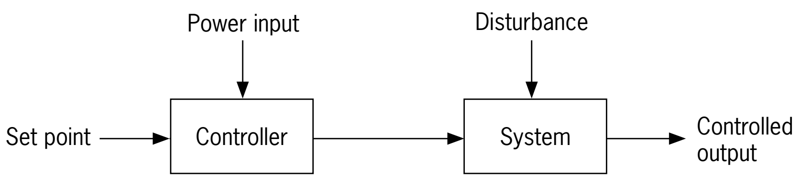
7 Electronic power control

By the end of this module, students should be able to:

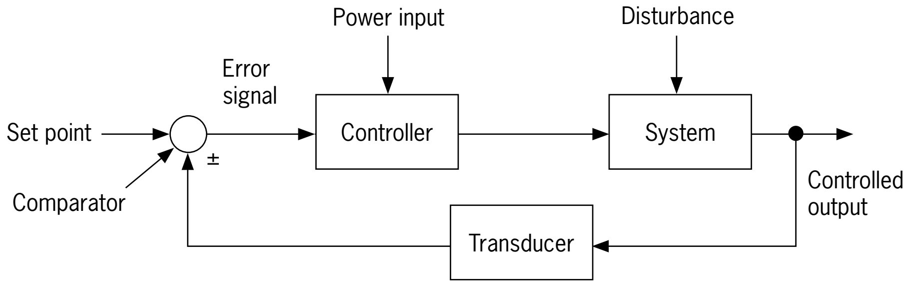
- describe the operating principles and characteristics of closed-loop control systems with regard to damping and stability;
- draw and label block diagrams and describe the operation principle of selected industrial process control systems as applied to the manufacturing industry; and
- draw, label and describe the operating principles of both industrial- and office-type uninterrupted power supply units as applied to the electronic industry.

Exercise 7.1 **SB page 98**

1.



The open-loop process control system is one in which a human operator is used to make changes or to take corrective action during the manufacturing process. An open-loop control system does not employ feedback and the control element has no reference data concerning the variable being controlled.



A closed-loop system is one that senses the output, compares the output against a desired condition and then corrects the system to achieve the desired output without the aid of a human operator.

A closed-loop control system can be described as an assembly of components used to maintain a desired output by controlling the energy input. Closed-loop systems regulate the energy supplied to the process.

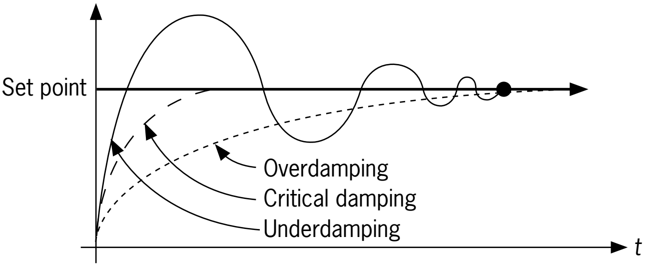
A closed-loop control system is error-actuated; that is, there must be a change in the measured value of the controlled variable before corrective action can occur. The quality of the system determines the maximum error that must be present before corrective action will take place.

2. The first characteristic of a closed-loop control system would be the degree of closeness between the desired value (set point) and the measured value of the controlled variable, regardless of the frequency and magnitude of load or set point changes.

   The second characteristic is the speed of response (or setting time); that is, the time interval between the detection of the error and the completion of the corrective action.

   The third characteristic is the offset, also known as the residual error or steady-state error. The residual error is the final difference between the desired value and the measured value of the controlled variable after the system has responded to a measurable change in the desired value or controlled output.

3. The residual error should respond to an increase in gain of the controller, but increasing the gain makes the system more sensitive, and as a result, may increase the maximum value of the error as well as the settling time. Another effect of increasing the controller gain is to change the type of damping applied to the system in response to disturbances.

4. 

Overdamped

Low gain; the dynamic or transient response is very slow and a large residual error may be present. The system would require a relatively long period to reach its set point; however, no oscillations (overshoot) will be present.

Critically damped

Low to medium gain; the least amount of damping that produces an output without any additional overshoot or oscillations that is required to reach steady state in the shortest possible time.

Underdamped

High gain; the gain has been increased and the output overshoots and oscillates around the set point with a diminishing amplitude response. Any further increase in gain will result in the system becoming completely uncontrolled and unstable. This type of damping also results in a relatively long recovery period, although its initial response to any change in the input is fast.

5. A system is defined as stable if its impulse response approaches zero as time approaches infinity.

   A system is defined as being unstable if, with zero input, the output increases indefinitely.

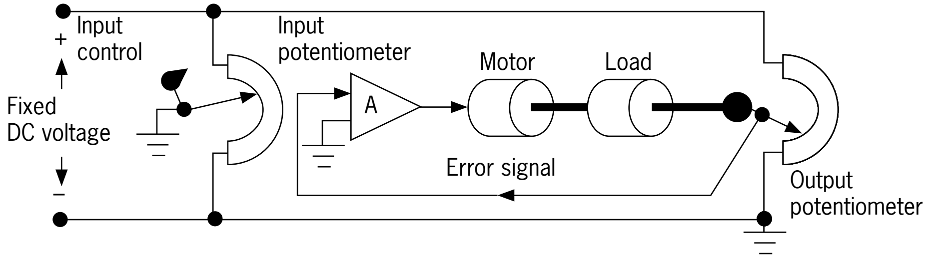
6. If the output of a system has continuous oscillation of constant peak-to-peak amplitude, the system is considered to be neutrally stable.

7. 7.1 A system is stable if its impulse response approaches zero as time approaches infinity.

   7.2
   - Regulator systems
   - Follow-up systems
   - Process control
   - Servomechanisms
   - Sequential control
   - Numeric control

   7.3 Process control is a term commonly applied to the control of variables in the manufacturing processes. Process control applications maintain such variables as temperature, flow rates, pressures, viscosity, density, levels, etc., as applied to chemical plants, oil refineries, food processing, blast furnace operations, vehicle production lines and nuclear power plants.

   7.4



The input potentiometer is adjusted to the required setting (set point). This sets up a difference signal voltage across the input of the amplifier. The error voltage is amplified and drives the servo motor. The shaft of the motor is mechanically coupled to the load and output potentiometer, which is displaced at the same angular displacement as the motor shaft. A new voltage setting is thereby produced on the output potentiometer.

When the voltage signal from the output potentiometer equals the input voltage across the input potentiometer, then $V_{IN} = V_O$, and the system has no (zero) error signal. The two voltages are now balanced and cancel each other so that no differential voltage appears across the output of the amplifier.

7.5 Numeric control uses a set of predetermined instructions or parameters to control the necessary sequence of manufacturing operations.

Numerically controlled systems employ stepper motors and/or servomechanisms. The controller, a simple computer, sequences a machine tool through various operations, such as vertical, horizontal and lateral displacement.

7.6 A sequential control system is a system that performs a series of operations in a prescribed sequence.

8. **Computer-aided design (CAD)**

The development of CAD was a natural tendency in the evolution of computer technology. A project to be designed is entered into the computer, which then uses the formulae and procedures, stored in its memory, to do all the necessary calculations required to design the project. The final design, along with recommended manufacturing techniques and methods, is then forwarded to the production line for manufacture.

Computer-aided manufacturing (CAM)

A CAM system is any system that uses a central control computer, also called a host, linked to other intelligent devices in a manufacturing environment.

9. 9.1 The simplest method is to use an existing design that is incorporated into the computer and has been tried and tested. The engineer only has to enter the required parameters of the new design and the computer will change the existing design to conform to the new requirements and/or dimensions. This system is extensively used by architects when designing new buildings. The computer drafts the plans according to the given building specifications and calculates building costs and material quantities.

Where an existing model is not suitable, the engineer can enter new specifications or parameters and the computer will use data from its memory to calculate new values and produce a new design.

Certain formulae, data or specifications may require changing to accommodate the new design. The engineer will update the data and the computer will produce the design.

When a new project is to be designed, it may require the engineer to start from scratch. The CAD system will act as the basic drawing board for the design with new formulae and data entered as required. All information is continuously displayed, updated and stored. When the engineer is satisfied, he/she can test the project by simulation and then print the final design. The design is then available in the computer memory for future reference.

9.2 Software distributors promote and sell CAD packages that contain design programs, formulae, symbols, data, instruction manuals and procedures required for the design of either a variety of aspects or systems dedicated to a single industry.

These programs allow the engineer to draft, test and simulate the design before it reaches the final stage. New designs and developments are stored in the computer memory to form a reference library (data base) for future projects and design.

9.3
- Electronic engineering
- Electrical engineering
- Civil engineering
- Mechanical engineering

10. **Existing design**

The simplest method is to use an existing design that is incorporated into the computer and has been tried and tested. The engineer only has to enter the required parameters of the new design and the computer will change the existing design to conform to the new requirements and/or dimensions. This system is extensively used by architects when designing new buildings. The computer drafts the plans according to the given building specifications and calculates building costs and material quantities.

Use existing data

Where an existing model is not suitable, the engineer can enter new specifications or parameters and the computer will use data from its memory to calculate new values and produce a new design. Certain formulae, data or specifications may require changing to accommodate the new design. The engineer will update the data and the computer will produce the design.

Generate new data

When a new project is to be designed, it may require the engineer to start from scratch. The CAD system will act as the basic drawing board for the design with new formulae and data entered as required. All information is continuously displayed, updated and stored. When the engineer is satisfied, he/she can test the project by simulation and then print the final design. The design is then available in the computer memory for future reference.

Simulation and testing

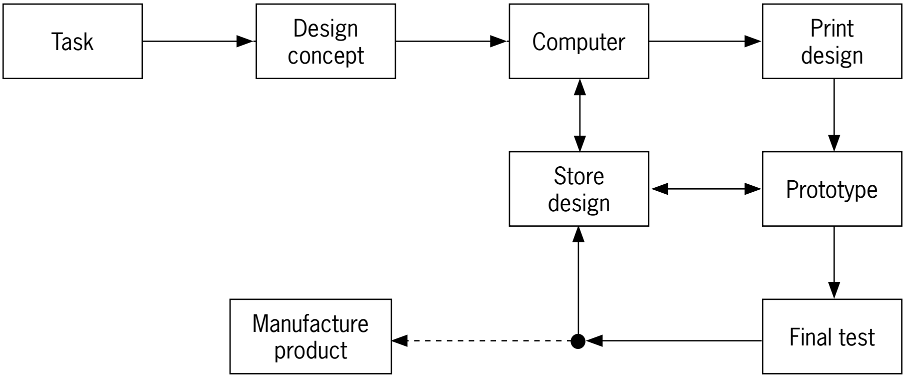
Another advantage of the CAD system is that any physical process that can be defined by mathematical formulae can be simulated by the computer.

All stages of the design development can therefore be tested before proceeding to the next stage and prior to producing the final design.

Design implementation

When the design engineer is satisfied with the final stage of design, a prototype can be built and tested according to the procedures specified by the CAD system. The values obtained during testing should, within close tolerances, duplicate the values specified by the computer. Final modifications can be made prior to the manufacturing process.

11.



12. **Design**

A comprehensive process analysis is the first step to achieve an optimal cost/benefit ratio in the entire production process.

Topics such as the following should be handled and solved through dialogue:

- Production process
- Availability of production lines
- Ergonomic working places
- Economic value stream
- Intelligent machine utilisation
- Short setup times
- Optimal lot size

Production safety

- Compressed air
- Electrical feeding and distribution
- Cooling systems
- HVAC systems (heating, ventilation and air conditioning)
- Material drying and conveying
- Lighting
- Cranes

Intralogistics includes aspects such as the following:

- Storage container and transportation units, e.g. carton boxes, small load carriers and pallets
- Material flow and warehousing, e.g. transport systems (forklift trucks, tugger trains, hand trucks, AGV systems, conveyer bands) and storage systems (block storage, rack systems, automated high rack warehouse)
- Receiving and shipping areas and public road access, e.g. connection to road and rail

13. **High data transmission rate**

The transmission rate of data is measured in bits per second, referred to as the baud rate. A computer that transmits 1 200 bits per second is said to be working at 1 200 baud. Generally speaking, the baud rates available for use with computers are standard values. The CAM system must be able to support a relatively high transmission rate for data, and baud rates of up to 55 000 are in use today.

Low baud rates would slow the operation of the system considerably. In order to ensure efficient operation, a baud rate of 9 600 or higher is desirable.

Fibre optic cables seem to be the best for use in an industrial environment. Transmission losses and propagation delays are minimal. Furthermore, fibre optic cables are impervious to the types of electronic interference normally found in factories.

A universal language

Effective communication with industrial robots, as well as other devices, is essential for a CAM system. Robots, numerical controllers and other devices must be able to communicate with the host computer quickly and efficiently. Obviously, a universal language is an absolute necessity. Robot languages, however, have been developed for specific robots and robot applications. In fact, there are probably as many robot languages as there are robots.

Now consider the number of high- and low-level languages available for use with computers; Fortran, Cobol, BASIC (and its variations) and PASCAL are just a few. The central computer in a CAM system must be able to communicate with other computers within the plant to perform inventory management, accounting, design and other functions, and to communicate with humans.

Communication interface

Interfacing the central computer with intelligent peripherals is a complicated task. Because of the incompatibility of the various devices that are connected to the central computer, it may be necessary to purchase or manufacture a custom interface network or device. This can necessitate the use of one or a number of interface devices.

These devices are known by a variety of names, such as UART, USART, PIA, ACIA, or PPI, RS 232, etc., but no matter what the device, their primary functions are to provide:

- a means for interconnecting computers and robot controllers as well as other peripheral devices that operate at different baud rates;
- a means for converting parallel data to serial data and vice versa; and
- a means for synchronous or asynchronous data transmission and reception.

In many cases, computers may already have one or more of these devices on board. In other instances, although not on board, they are readily available and easy to install, or, where the device is not available for a particular computer or robot controller, it is then necessary to develop the circuitry necessary to use the device.

Communication organisation

When the central computer is used for factory-wide control, it is necessary not only that it communicates with a great number of devices, but also that these devices are able to communicate with the central computer. As the central computer can only 'listen' to one peripheral at a time, some type of structured input/output system must be organised.

It is possible to use device polling. In this type of system, the central computer checks, or polls, each device in turn to determine whether the device needs service. This type of system has a number of disadvantages:

- The central computer is tied up in a continual device polling loop. This limits the central computer to performing just one operation, which severely restricts the potential of the central computer
- Because the computer polls each device sequentially, there is the possibility that an emergency situation will be neglected while some routine housekeeping is being done.

Because than using device polling, it is recommended that an interrupt-driven system be used. In this type of system, each peripheral generates an interrupt when service is needed. However, because of the number of devices and the variety of possible reasons for requesting an interrupt, there must be some way of assigning a priority to each type of interrupt from each device.

14.
- Provide a means for interconnecting computers and robot controllers as well as other peripheral devices that operate at different baud rates
- Provide a means for converting parallel data to serial data and vice versa
- Provide a means for synchronous or asynchronous data transmission and reception

15. **Device polling**

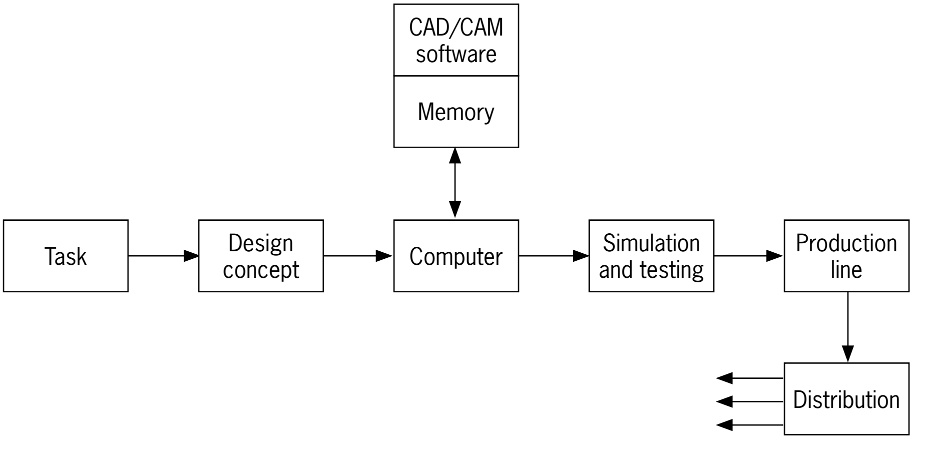
In this type of system, the central computer checks, or polls, each device in turn to determine whether the device needs service. This type of system has a number of disadvantages such as:

- The central computer is tied up in a continual device polling loop. This limits the central computer to performing just one operation, which severely restricts the potential of the central computer
- Because the computer polls each device sequentially, there is the possibility that an emergency situation will be neglected while some routine housekeeping is being done.

Interrupt-driven system

In this type of system, each peripheral generates an interrupt when service is needed. However, because of the number of devices and the variety of possible reasons for requesting an interrupt, there must be some way of assigning a priority to each type of interrupt from each device.

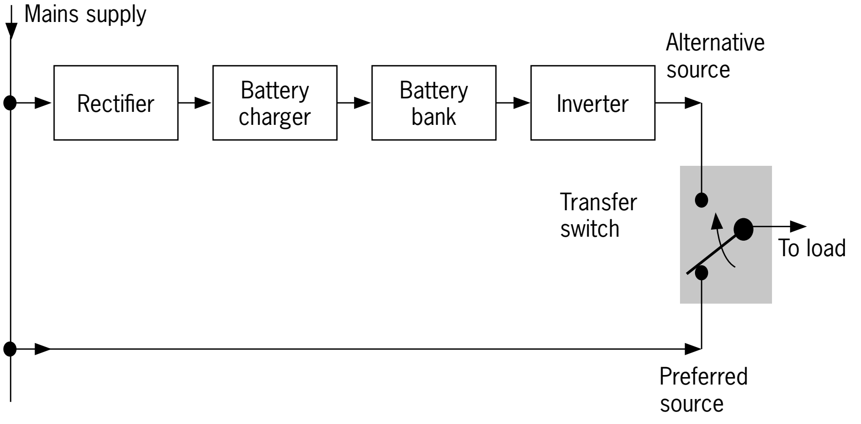
16.



- Products can be modified on the production line by making changes to the CAD system.
- The CAM system will automatically reproduce design changes on the production line.
- Twenty-four-hour productive production without 24-hour skilled supervision is possible.
- Machinery and plant can operate without skilled labour, subject to breakdowns, routine services and power failure.

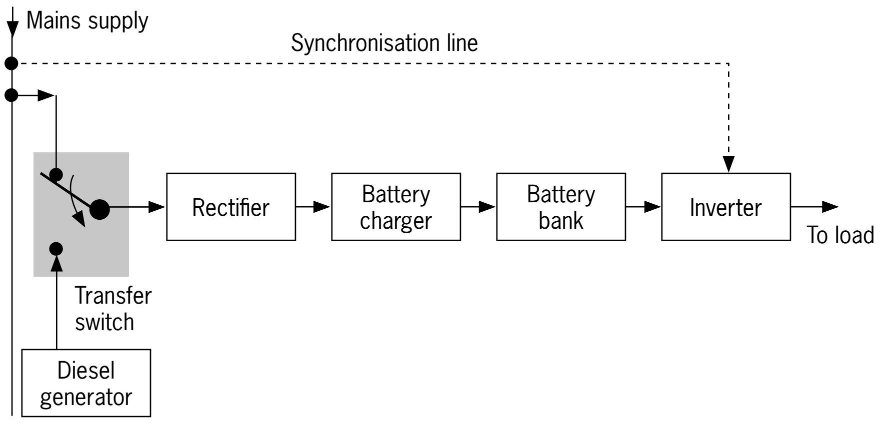
17. An uninterruptible power supply or uninterruptible power source (UPS) is an electrical apparatus that prevents shut down and damage to critical equipment and systems and a load when the input power source or mains power fails.

18. 18.1



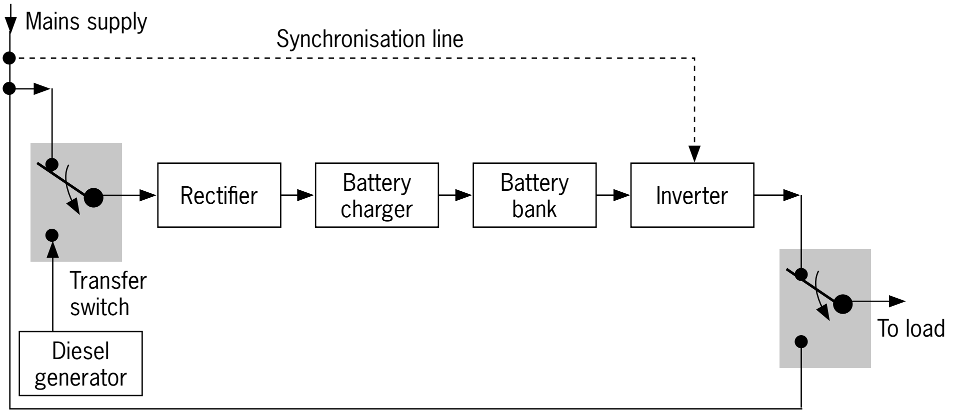
The transfer system is used in applications where a switch-over time of several cycles is acceptable, usually in the range of 4 to 12 cycles. The load is normally supplied from the preferred AC source via an automatic bus transfer switch to the load, with the battery being charged via the rectifier and battery charger. When the power fails, the automatic bus transfer switch connects the inverter, which is operated from the battery supply, to the load. This arrangement is satisfactory for applications where a short power interruption is not critical.

18.2



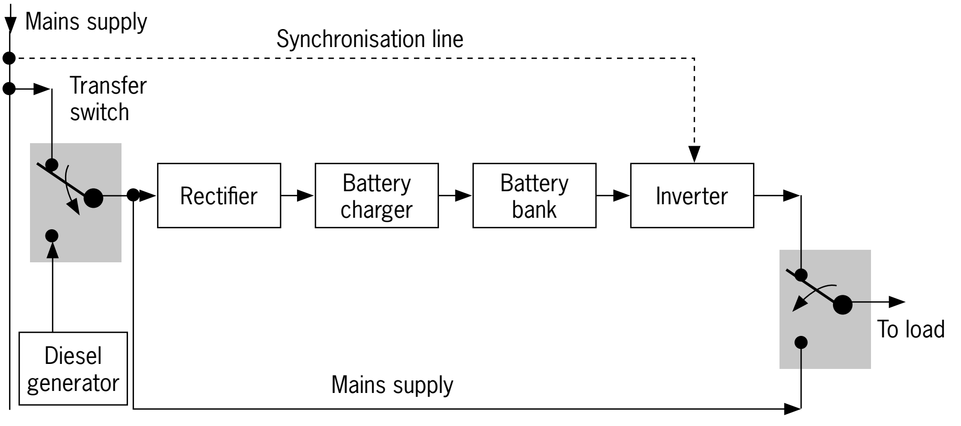
In the forward system, the load is supplied from the mains line via the charger, battery and inverter. In the event of a power failure, the static transfer switch automatically transfers from the AC mains supply to the output of the diesel generator. When power is restored, the load is transferred back to the mains supply via the inverter and the diesel generator is disconnected.

18.3



In the reverse system, the load is supplied from the inverter and is only transferred to the mains supply as a result of an inverter failure.

18.4



A combination of the forward and reverse system effectively gives double protection in the event of power or inverter failure.

19. **Transfer system**

This system is used in applications where a switch-over time of several cycles is acceptable, usually in the range of 4 to 12 cycles. The load is normally supplied from the preferred AC source via an automatic bus transfer switch to the load, with the battery being charged via the rectifier and battery charger. When the power fails, the automatic bus transfer switch connects the inverter, which is operated from the battery supply, to the load.

Advantage

This arrangement is satisfactory for applications where a short power interruption is not critical.

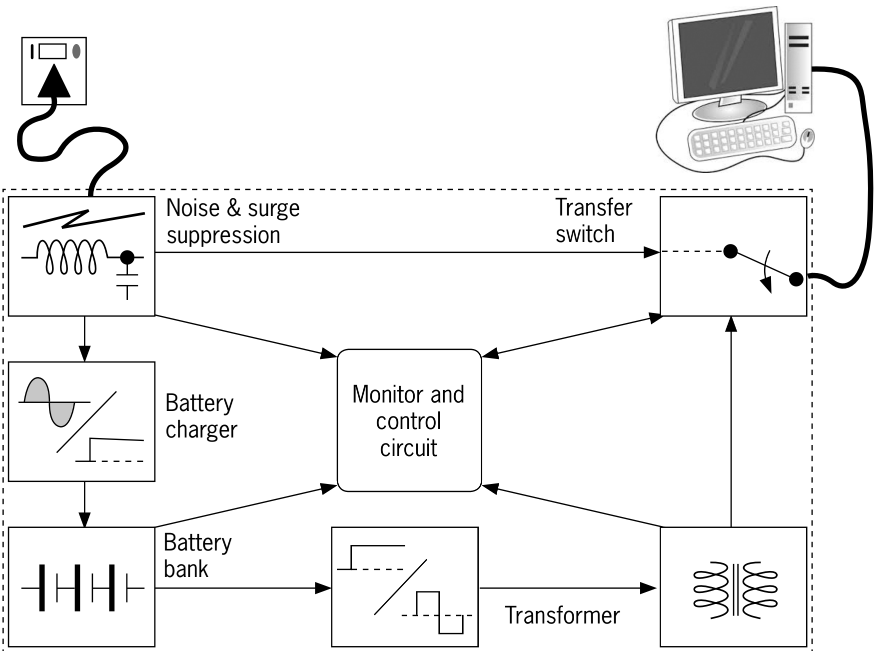
Continuous type

Advantage

This type of UPS ensures that there is no interruption of power to the load and will also sustain interruptions of longer duration.

In the event of power failure, the inverter is supplied by the battery and there is no interruption of power to the load. Because the ability of the battery to support load is limited (usually between 10 and 60 minutes), a standby generator may be automatically started up after a predetermined time interval and connected to the rectifier by means of an automatic change-over switch. The standby generator has effectively replaced the mains supply and maintains the charge on the battery.

20.



21. **Load transfer switch**

The load transfer switch is an electrodynamic relay that serves to transfer the computer equipment (load) from the utility to the UPS's alternate power source rapidly in the event of a mains failure. When the mains are restored to within safe limits, the switch acts to retransfer the load to the utility (mains). Except for the user control switches, the transfer switch is the only moving part in the UPS. The time required for the relay to transfer the load to either power source is much faster than that which is required by any modern computer or computer peripheral device.

Battery charger

In the event of utility failure, the UPS supplies power to the equipment that is derived from energy taken from a battery. The UPS's battery charger converts the alternating current supplied by the utility into a direct current that is compatible with the battery. The charger maintains the battery at a constant voltage to ensure that the battery will have the capacity to support the load as often as possible. This charging method, known as 'float charging', provides maximum battery service life and minimal internal heating. The battery charger normally operates whenever the UPS is plugged in, whether or not the UPS is turned on.

Battery

The UPS's battery is an energy source much like the battery in an automobile. Also, like most automobile batteries, the UPS's battery is a modern maintenance-free lead-acid type; it is sealed and leak-proof. The battery has a typical service life of three to six years. The service life is extended when the UPS is kept below 30 °C. In the event of utility failure, the battery supplies the inverter with the required energy.

Inverter

The UPS must convert the battery's energy (DC) into a form that all computer equipment can rely upon during a utility failure. The inverter converts the battery's DC into AC using solid-state devices, controlled using a technique known as 'pulse width modulation'. This technique is highly efficient, which means that little battery power is wasted in the conversion process. Hence, equipment can run for reasonable periods from the UPS before the battery's capacity is exhausted.

Transformer

The UPS's transformer is an electrical component that steps up the output voltage of the inverter to the normal utility line voltage (220-V AC). In addition, it serves to isolate the UPS from equipment failures.

Monitoring and control electronics

This block is the 'brain' of the UPS. The monitoring and control circuitry detects utility failures such as blackouts, sags and brownouts; synchronises the inverter's output frequency and phase to that of the utility; detects low-battery voltage conditions; directs the load transfer switch; and governs all user controls, indicators and computer interface functions.

22. Power failure
23. Computer VA = 250 × 2 = 500 VA

Computer	VA = 250 × 2	= 500 VA
Monitor	VA = 100 × 1,5	= 150 VA
Printer	VA = 250 × 0,3	= 75 VA
Total		= 725 VA

∴ purchase a 750-VA UPS

MODULE

8 Thyristor devices and SCR speed control

By the end of this module, students should be able to:

- draw, label and describe the operation principles of the various rectifier, inverter, converter and cycloconverter devices as applied to the electronic and related industries;
- analyse the characteristics and manufacturer's specifications of SCR devices as applied to DC, single-phase and three-phase circuits; and
- draw, label and describe the operating principles of both AC and DC motor speed control circuits with specific reference to SCR devices.

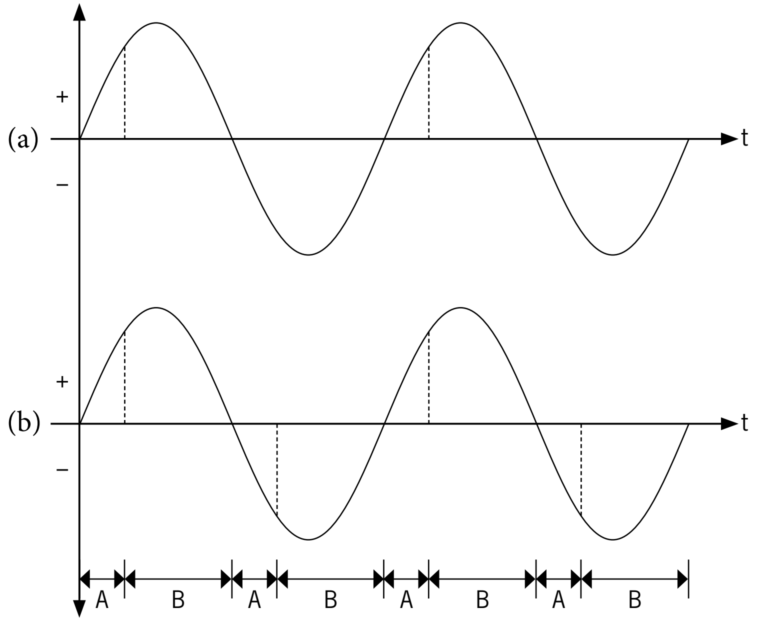
Exercise 8.1 SB page 108

1. In order to trigger a thyristor positively in the shortest time, it is desirable to have a gate current with a fast rise time up to the maximum permitted value. This rise time is best achieved by pulse techniques, where the firing circuit generates a fast rise pulse of sufficient length to allow the anode current enough time to reach its latching value.
2. Firing angle α is the angle at which the SCR gets turned on and starts conducting.

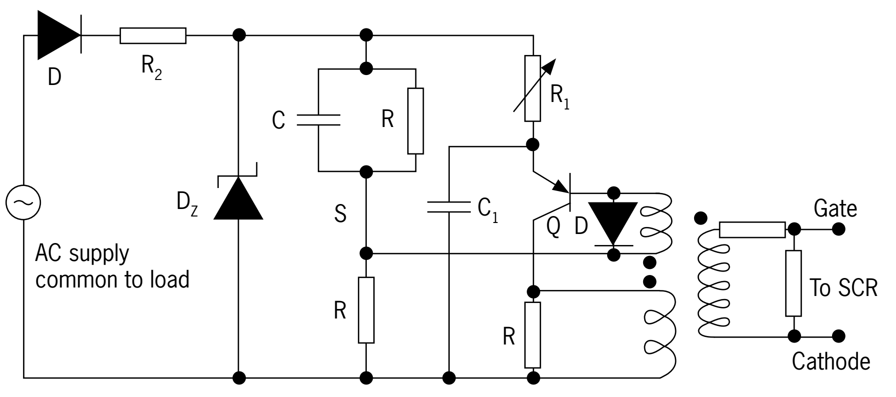
   Conduction angle β is the number of degrees in a half-cycle AC wave during which an SCR is turned on and conducting.

   Therefore, the portion of the supply signal during which the SCR is forward-biased is made up of two parts, namely the firing angle (α) and the conduction angle (β).

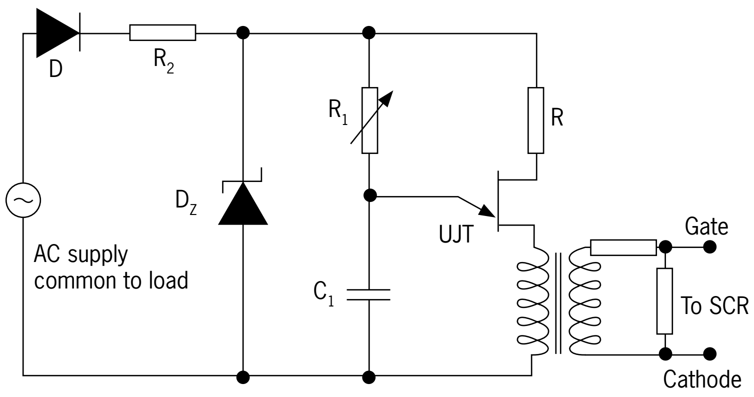
In both waveforms, the area indicated by A is called the delay angle or firing angle and the area indicated by B is called the conduction angle.



3. a)

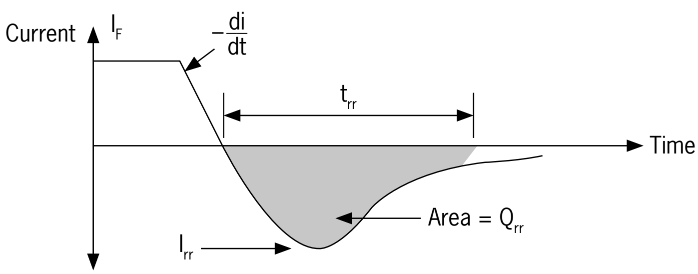


b)



Both circuits are supplied from an AC source that is clipped by the Zener diode D_z to give a level voltage to the R_1C_1 series circuit. Resistor R_2 drops the voltage difference between the supply and the Zener diode. In both circuits, the voltage on capacitor C_1 will rise exponentially at a rate determined by the value of resistor R_1.

4.



Given a forward diode or thyristor current I_F, which is turned off at a given $\frac{di}{dt}$ rate, the current will go into reverse until the carrier storage charge Q_{rr} is recovered. At defined values of I_F and $\frac{di}{dt}$, a particular device will have rated values of Q_{rr} (the reverse recovery charge) associated with a reverse recovery time t_{rr} and a reverse recovery current I_{rr}.

When triggered in the usual manner, the junction of the thyristor will break-over first in the region of the gate electrode. If the total anode current were established immediately, the current density in this region would be excessive, resulting in damage by overheating.

The rate of rise of the anode current must be limited to the time taken for the junction breakdown to spread completely across the slice; a time typically 10 μ. A thyristor will also have a $\frac{di}{dt}$ rating, which is not to be exceeded.

5.
- The loss during forward conduction, which is a function of the forward voltage drop and conduction current; this is the major source of loss at mains and lower-frequency operation
- The loss associated with the leakage of current during the blocking state

- The loss occurring in the gate circuit as a result of the energy input from the gating signal; in practice, with pulse firing, these losses are negligible
- The switching loss that is, the energy dissipated in the device during turn-on and turn-off, which can be significant when switching is occurring at a relatively high frequency, say at 1 kHz

6. The losses will lead to heat generation within the device and consequently a rise in temperature until the rate of heat dissipation matches the loss. The heat generated in the junction area is transferred to the base and then to a heatsink.

   The temperature level to which the base can be allowed to rise is not high enough for much heat dissipation to take place by radiation. Hence, most of the heat transfer takes place by convection; to air in the case of the metal-finned heatsink. Heat transfer will take place from a higher-temperature region to a lower-temperature region.

7. - Unlimited voltage and current ratings
   - Instant turn-on and turn-off times
   - Zero leakage current
   - Zero conduction and switching losses
   - Zero gate firing power requirement
   - Ability to withstand current overloads and voltage transients
   - Easy to protect against spurious turn-on and fault conditions
   - Low cost and ease of assembly

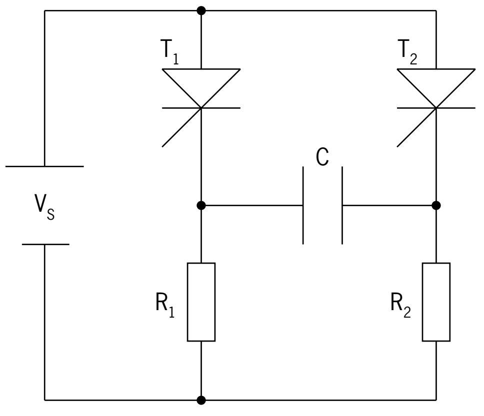
8. The important criteria in circuit applications very often depend on the parameters of ratings, conduction losses, switching losses, switching times, control strategy and cost.

   The conventional thyristor has the highest ratings of all devices, is robust, has low conduction losses and is inexpensive, but is slow to turn on and cannot be turned off other than by cessation of its load current.

   For applications linked to the public electricity supply at 50 Hz or 60 Hz, such as rectifiers, the conventional thyristor is the first choice, its capability of withstanding high forward and reverse voltages being essential to this application.

Exercise 8.2 — SB page 112

1.

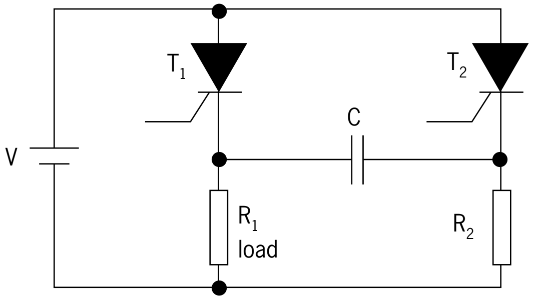


$$V = E - 2E \times e^{-t/\tau}$$
$$0 = E - 2E \times e^{-t/\tau}$$
$$-E = -2E \times e^{-t/\tau}$$
$$\frac{-E}{-2E} = e^{-t/\tau}$$
$$\frac{1}{2} = -e^{-t/\tau}$$
$$\ln\frac{1}{2} = {}^{-t/\tau}$$
$$-0{,}693 = {}^{-t/\tau}$$
$$\tau = \frac{t}{0{,}693}$$
$$= \frac{80 \times 10^{-6}}{0{,}693}$$
$$= 0{,}11544 \text{ ms}$$

$$\tau = R \times C$$
$$C = \frac{\tau}{R}$$
$$= \frac{0{,}11544 \times 10^{-3}}{5} \text{ (use R = 5 and not } 5R_1\text{)}$$
$$= 23{,}088\ \mu F$$

2. The process used for turning off a thyristor is called commutation. By the commutation process, the thyristor operating mode is changed from forward-conducting mode to forward-blocking mode. So, the thyristor commutation methods or thyristor commutation techniques are used to turn off the thyristor.
3. In order to turn off the conducting SCR, the following conditions must be satisfied.
   - The anode or forward current of the SCR must be reduced to zero or below the level of holding current.
   - A sufficient reverse voltage must then be applied across the SCR to regain its forward-blocking state.

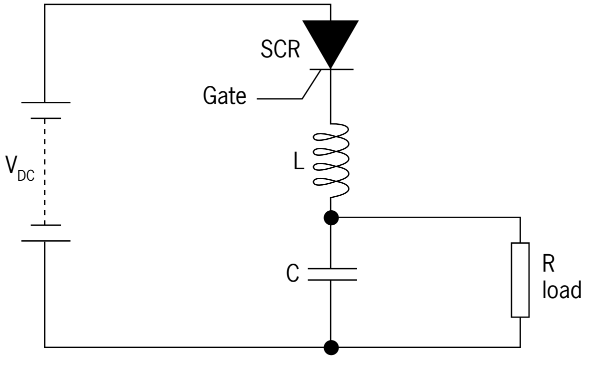
4. • Parallel capacitance commutation
   • Series resonant commutation
   • Parallel resonant commutation

5. 

The principle is that firing thyristor T_1 connects the battery to load resistor R_1 and at the same time enables capacitor C to charge via resistor R_2. Firing thyristor T_2 places the charge on capacitor C across thyristor T_1, turning it off. Thyristor T_2 will remain on, with current flow via resistor R_2, and capacitor C oppositely charging via resistor R_1. Firing thyristor T_1 now connects the battery to load resistor R_1 and at the same time turns thyristor T_2 off by placing capacitor C across it.

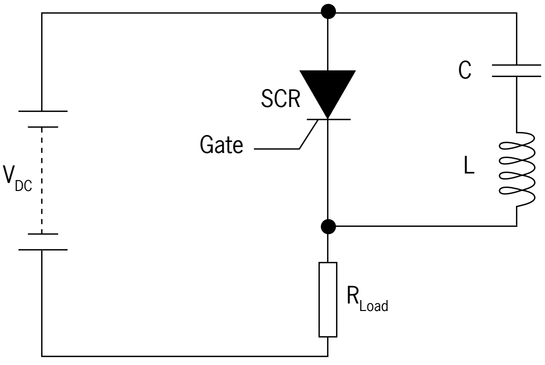
6. The disadvantage of this circuit is the loss in resistor R_2 as it carries current throughout the load off-period. The loss can be minimised by making resistor R_2 large compared to resistor R_1, but this will lengthen the charging time of the capacitor, hence limiting the rate at which the load can be switched.

7. **Series resonance commutation**



The self-oscillating property of a capacitor-inductor combination can be utilised to turn off the load thyristor at a given time after turn-on without the need for a second or auxiliary thyristor. The series resonant circuit must be underdamped for the thyristor current to attempt to reverse and turn-off to take place.

Parallel resonance commutation

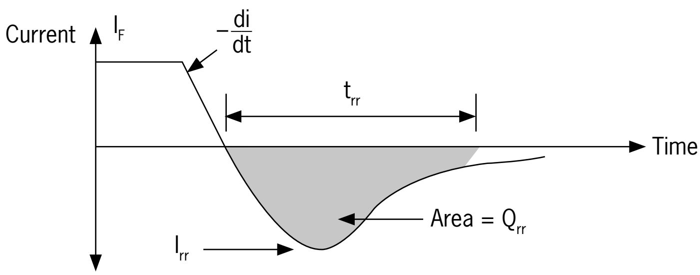


In the parallel resonance circuit, the capacitor is charged to the level of the battery voltage when connected. Firing the thyristor connects the battery to the load and at the same time sets up an oscillation in the LC circuit. Provided the oscillating current is greater than the load current (V/R), the thyristor current will attempt to reverse, and turn-off then takes place. For the first half cycle, the thyristor current will increase, but will later be reduced to zero during the early part of the reverse half cycle of oscillation.

It is important that the circuit for R has a value such that the RLC series circuit is critically damped: ($R^2 = 4\ L/C$). If R were reduced, it would rapidly reach the case of the load current being greater than the oscillating LC current. If R were increased, then the charging of capacitor C would be slow, making the minimum off-time too long.

Exercise 8.3 SB page 123

1.



Given a forward diode or thyristor current I_F, which is turned off at a given $\frac{di}{dt}$ rate, the current will go into reverse until the carrier storage charge Q_{rr} is recovered. At defined values of I_F and $\frac{di}{dt}$, a particular device will have rated values of Q_{rr} (the reverse recovery charge) associated with a reverse recovery time t_{rr} and a reverse recovery current I_{rr}.

2. Given a high enough rate of rise of voltage, say 100 V/μs, the minority carriers will be accelerated across the junction at a rate high enough to fire (trigger)

the thyristor into the on-state, even though there is an absence of gate current. A thyristor will have a rated value of $\frac{dV}{dt}$, which must not be exceeded.

3. The rate of rise of the anode current must be limited to the time taken for the junction breakdown to spread completely across the slice; a time typically 10 μ. A thyristor will also have a $\frac{di}{dt}$ rating, which is not to be exceeded.
4. The loss during forward conduction, which is a function of the forward voltage drop and conduction current; this is the major source of loss at mains and lower-frequency operation
5. The switching loss; that is, the energy dissipated in the device during turn-on and turn-off, which can be significant when switching is occurring at a relatively high frequency, say at 1 kHz
6. No. The heat generated in the junction area is transferred to the base and then to a heatsink. The temperature level to which the base can be allowed to rise is not high enough for much heat dissipation to take place by radiation.

   Hence, most of the heat transfer takes place by convection; to air in the case of the metal-finned heatsink. Heat transfer will take place from a higher-temperature region to a lower-temperature region.

7. $$I_M = \frac{V}{R} = \frac{100}{20} = 5\text{ A}$$

   $$\tau = \frac{L}{R} = \frac{0{,}5}{20} = 0{,}025\text{ s}$$

   $$i = I_M\,(1 - e^{\frac{-t}{\tau}}) = 5(1 - e^{\frac{-30 \times 10^{-6}}{0{,}025}}) = 6\text{ mA}$$

   The SCR will fail to remain on, as the latching current must be 0,05 A.

   $$I_{shortfall} = I_{LATCH} - i = 0{,}05 - 6 \times 10^{-3} = 0{,}044\text{ A}$$

   $$R = \frac{V_S}{I_{shortfall}} = \frac{100}{0{,}044} = 2{,}27\text{ k}\Omega$$

8. The angle must always be in radians, or you can simply use π as 180°.

   $$I_{PEAK_{60°}} = \frac{I_{MEAN} \times 2 \times \pi}{\theta} = \frac{170 \times 2 \times 180}{60} = 1\,020\text{ A}$$

   $$I_{RMS} = I_{PEAK}\sqrt{\frac{\theta}{2\pi}} = 1\,020\sqrt{\frac{60}{2 \times 180}} = 416{,}41\text{ A}$$

   $$I_{PEAK_{120°}} = \frac{I_{MEAN} \times 2 \times \pi}{\theta} = \frac{218 \times 2 \times 180}{120} = 654\text{ A}$$

   $$I_{RMS} = I_{PEAK}\sqrt{\frac{\theta}{2\pi}} = 654\sqrt{\frac{120}{2 \times 180}} = 377{,}5\text{ A}$$

$$I_{PEAK_{180°}} = \frac{I_{MEAN} \times 2 \times \pi}{\theta} = \frac{250 \times 2 \times 180}{180} = 500 \text{ A}$$

$$I_{RMS} = I_{PEAK}\sqrt{\frac{\theta}{2\pi}} = 500\sqrt{\frac{180}{2 \times 180}} = 353{,}55 \text{ A}$$

$$I_{PEAK_{240°}} = \frac{I_{MEAN} \times 2 \times \pi}{\theta} = \frac{277 \times 2 \times 180}{240} = 415{,}5 \text{ A}$$

$$I_{RMS} = I_{PEAK}\sqrt{\frac{\theta}{2\pi}} = 415{,}5\sqrt{\frac{240}{2 \times 180}} = 399{,}25 \text{ A}$$

$$I_{PEAK_{300°}} = \frac{I_{MEAN} \times 2 \times \pi}{\theta} = \frac{295 \times 2 \times 180}{300} = 354 \text{ A}$$

$$I_{RMS} = I_{PEAK}\sqrt{\frac{\theta}{2\pi}} = 354\sqrt{\frac{300}{2 \times 180}} = 323{,}155 \text{ A}$$

$$I_{PEAK_{360°}} = \frac{I_{MEAN} \times 2 \times \pi}{\theta} = \frac{305 \times 2 \times 180}{360} = 305 \text{ A}$$

$$I_{RMS} = I_{PEAK}\sqrt{\frac{\theta}{2\pi}} = 305\sqrt{\frac{360}{2 \times 180}} = 305 \text{ A}$$

9. $$R_{TH} = \frac{T_1 - T_2}{P_{LOSS}} = \frac{125\text{ °C} - 35\text{ °C}}{50\text{ W}} = 1{,}8\text{ °C/W}$$

$$T_{HR} \text{ of heatsink} = R_{TH} - R_{JUNCTION} = 1{,}8\text{ °C/W} - 1{,}2\text{ °C/W} = 0{,}6\text{ °C/W}$$

$$T_B = T_S + (P_{LOSS} \times TH_R) = 25\text{ °C} + (50\text{ W} \times 0{,}6\text{ °C/W}) = 65\text{ °C}$$

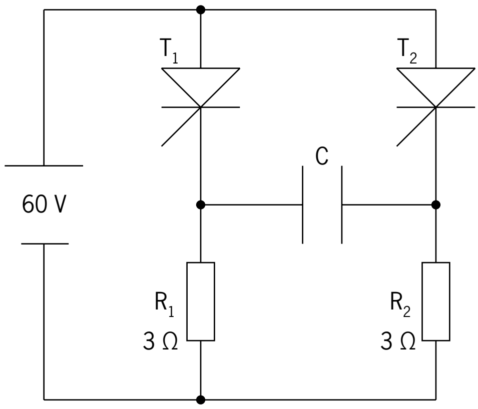
10. R_{TH} = thermal resistance + heatsink resistance
$= 1{,}6$ °C/W + 2 °C/W
$= 3{,}6$ °C/W

$$P_{LOSS} = \frac{T_j - T_a}{R_{TH}} = \frac{152\text{ °C} - 35\text{ °C}}{3{,}6\text{ °C/W}} = 25\text{ W}$$

11. Rise in temperature = $Z_{TH} \times P_{LOSS}$
$= 3{,}6$ °C/W × 1 800 W
$= 81$ °C

Tj = cold conditions + rise in temp
= 35 °C + 81 °C
= 116 °C

12. 12.1



$$V = E - 2E \times e^{-t/\tau}$$
$$0 = E - 2E \times e^{-t/\tau}$$
$$e^{-t/\tau} = \frac{0 - E}{2E}$$
$$\frac{-t}{\tau} = \ln 0{,}5$$
$$t/\tau = 0{,}693$$
$$\tau = \frac{t}{0{,}693}$$
$$= 36{,}067\ \mu s$$

$$\tau = R \times C$$
$$C = \tau/R$$
$$= 36{,}067 \times 10^{-6}/3$$
$$= 12\ \mu F$$

12.2

$$V = E - 2E \times e^{-t/\tau}$$
$$\frac{0{,}8 \times E - E}{-2 \times E} = \frac{-2 \times E \times e^{-t/\tau}}{-2 \times E}$$
$$\frac{-0{,}2 \times E}{-0{,}2 \times E} = e^{-t/\tau}$$
$$0{,}1 = e^{-t/\tau}$$
$$\text{Lin } 0{,}1 = -t/\tau$$
$$-t = \ln 0{,}1 \times \tau$$
$$= -2{,}302 \times 36{,}067 \times 10^{-6}$$
$$= 83\ \mu s$$

13. A rectifier circuit is one that links an AC supply to a DC load; that is, it converts an alternating voltage supply into a direct voltage. The direct voltage so obtained is not normally at 100% level, as from a battery source, but contains an alternating ripple component superimposed on the mean (DC) level.

The uncontrolled rectifier circuits contain only diodes, giving a DC load voltage fixed in magnitude relative to the AC supply voltage magnitude.

In the fully controlled rectifier circuits, all the rectifying elements are thyristors. In these circuits, by suitable control of the phase angle at which the thyristors are triggered, it is possible to control the mean (DC) value of, and to reverse the, DC load voltage.

The half-controlled rectifier circuits contain a mixture of thyristors and diodes, which prevent a reversal of the load voltage but allow adjustment of the direct (mean) voltage level.

14. Advantages of AC motor speed control:
    - High efficiency against low motor cost
    - Compact design and light weight with high torque
    - Low maintenance
    - Wide range of motor designs available

    Advantages of DC motor speed control:
    - A wide range of speed control
    - Can operate under constant or variable torque conditions
    - Rapid acceleration, deceleration and reversing
    - Open- or closed-loop operation
    - Operate under regenerative conditions
    - Range from fractional kW to 7 500 kW
15. A free-wheeling, flywheel or by-pass diode is best described as a commutating diode, as its function is to commutate (or transfer) load current away from the rectifier whenever the load voltage goes into a reverse state.

16. 16.1

$$\begin{aligned} V_{MAX} &= V_{RMS} \times \sqrt{2} \\ &= 32 \times \sqrt{2} \\ &= 45{,}25\ \text{V} \end{aligned}$$

16.2

$$\begin{aligned} V_{MEAN} &= \frac{V_M}{2 \times \pi\,(1 + \cos\alpha)} \\ &= \frac{339{,}4}{2 \times 3{,}142\,(1 + \cos 45°)} \\ &= 92{,}2\ \text{V} \end{aligned}$$

$$\begin{aligned} V_M &= V_{RMS} \times \sqrt{2} \\ &= 240 \times \sqrt{2} \\ &= 339{,}4\ \text{V} \end{aligned}$$

$$\begin{aligned} PIV &= V_M \times \sqrt{2} \\ &= 240 \times \sqrt{2} \\ &= 339{,}4\ \text{V} \end{aligned}$$

$$\begin{aligned} I_{RMS} &= \frac{I_M \times \sqrt{2}}{2} \\ I_M &= \frac{I_{RMS} \times 2}{\sqrt{2}} \\ &= \frac{10{,}6 \times 2}{\sqrt{2}} \\ &= 14{,}99\ \text{A} \end{aligned}$$

16.3

$$\begin{aligned} PIV &= V_{RMS} \times \sqrt{2} \\ &= 220 \times \sqrt{2} \\ &= 311\ \text{V} \end{aligned}$$

$$\begin{aligned} V_{MEAN} &= 0{,}637 \times V_{MAX} \\ &= 0{,}637 \times 311 \\ &= 198{,}107\ \text{V} \end{aligned}$$

16.4 $V_M = 1,414 \times V_{RMS}$
$= 1,414 \times 230$
$= 325,269$ V

$V_{MEAN} = 2 \times V_M \cos \alpha - 2\ SRC_V$
$= 2 \times 325,269 \cos 30° - 2 \times 1,2$
$= 176,9$ V

$PIV = V_{RMS} \times \sqrt{2}$
$= 230 \times \sqrt{2}$
$= 325,269$

16.5 $V_{MEAN} = \frac{V_M}{\pi}(1 + \cos \alpha)$
$= \frac{169,7}{3,142}(1 + \cos 135°)$
$= 15,8$ V

$V_{MAX} = V_{RMS} \times \sqrt{2}$
$= 120 \times \sqrt{2}$
$= 169,7$ V

$PIV = V_{RMS} \times \sqrt{2}$
$= 120 \times \sqrt{2}$
$= 169,7$ V

Thyristor current rating $= 0,707 \times I_{MAX}$
$= 0,707 \times 25$
$= 17,67$ A

16.6 $I_{MAX} = 25$ A
$V_L = V_{PH} \times \sqrt{3}$
$= 120 \times \sqrt{3}$
$= 207,84$ V

$V_{MEAN} = \frac{3}{\pi} \times V_L \times \sqrt{2}$
$= \frac{3}{3,142} \times 207,84 \times \sqrt{2}$
$= 280,65$ V

16.7 $PIV = V_L \times \sqrt{2}$
$= 311 \times \sqrt{2}$
$= 439,82$ V

$V_{MEAN} = \frac{3}{\pi} \times V_L \times \sqrt{2} \times \cos \alpha - 3 \times SRC_V$
$= \frac{3}{3,142} \times 311 \times \sqrt{2} \times \cos 60° - 3 \times 1,5$
$= 205$ V

16.8 $V_{MEAN} = \frac{3\sqrt{3}}{2\pi} \times V_{PH} \times \sqrt{2} \times (1 + \cos \alpha) - 3 \times SRC_V$
$= \frac{3\sqrt{3}}{2 \times 3,142} \times 180 \times \sqrt{2} \times (1 + \cos 90°) - 3 \times 1,5$
$= 209$ V

$V_L = V_{PH} \times \sqrt{3}$
$= 180 \times \sqrt{3}$
$= 311,76$ V

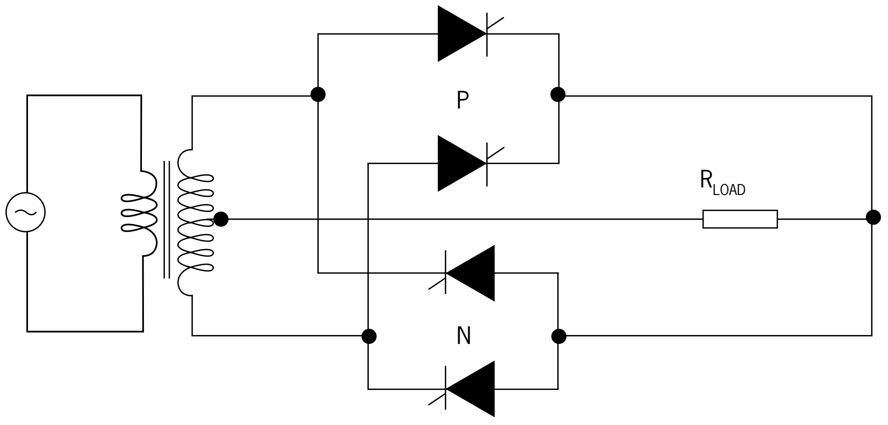
$PIV = V_L \times \sqrt{2}$
$= 311,76 \times \sqrt{2}$
$= 440,89$ V

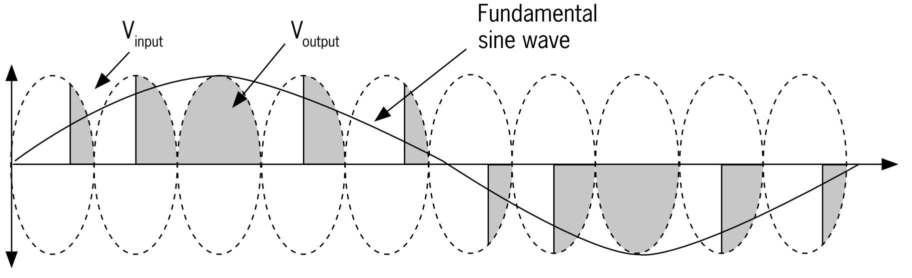
17. A converter is a device for altering the nature of an electric current or signal, especially from AC to DC or vice versa.

    Cyclo-conversion is concerned mostly with direct conversion of energy to a different frequency by synthesising a low-frequency wave from appropriate sections of a higher-frequency source.

    Inverters cover those circuits that have a DC source and by appropriate switching of rectifying devices enable an alternating voltage to be synthesised for feeding to an AC load.

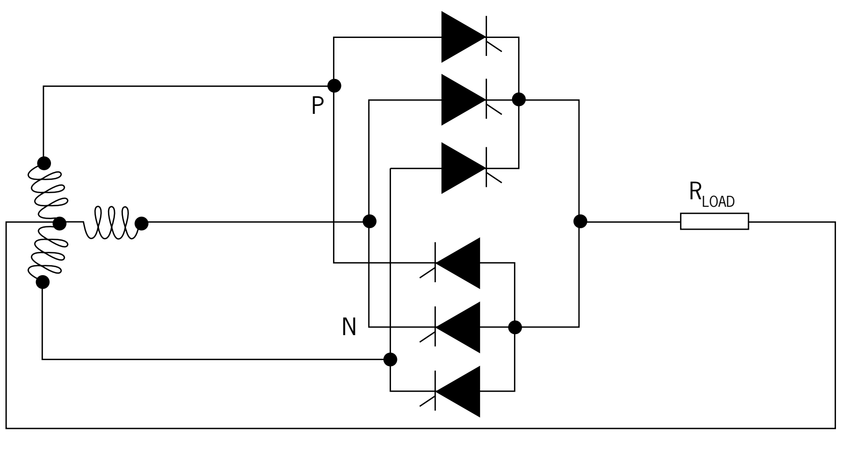
18.



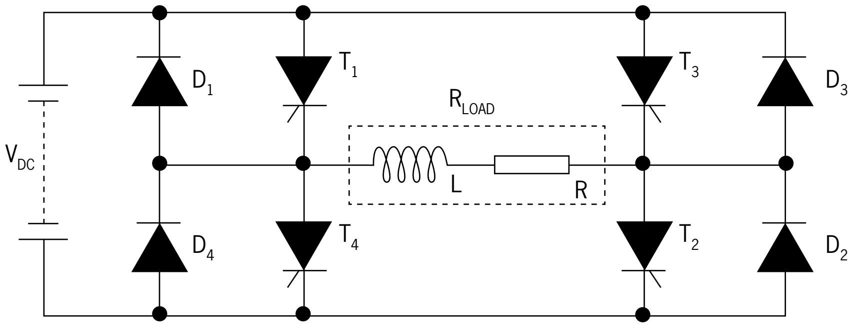


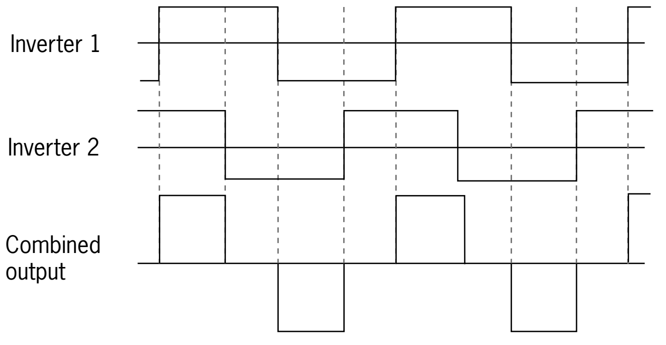
Each converter is a bi-phase half-wave connection; the positive group labelled P and the negative group (for reverse current) labelled N.

The load-voltage waveform is constructed on the basis that group P only conducts for five half cycles. For the next five half cycles, group N only conducts to synthesise the negative half cycle to the load. A close approximation to a sine wave can be synthesised by phase, delaying the firing of the thyristors. The figure above clearly illustrates an output frequency with the fundamental waveform superimposed on top of the output frequency.



19.

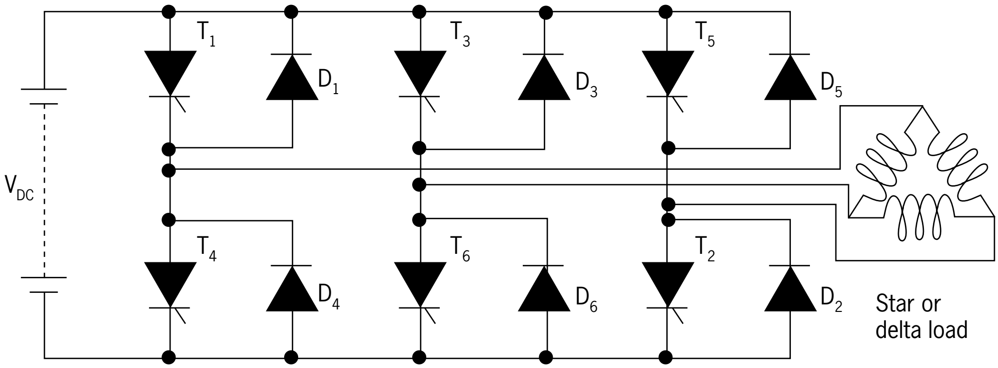




The thyristors are fired by a continuous train of gate pulses for 180° of the inverter output voltage. Looking at the latter end of the positive half cycle, the load current is positive and growing exponentially; however, when thyristors T_3 and T_4 are gated to turn off thyristors T_1 and T_2, the load voltage reverses, but not the load current. The only path for the load current is via diodes D_3 and D_4, which connect the DC source to the load, giving a reverse voltage, with the stored inductive energy of the load being returned to the DC source until the load current falls to zero. Once the load current ceases, thyristors T_3 and T_4 can conduct to feed power into the load; the load current now growing exponentially.

Because the thyristors require re-firing at the instant of the load current zero, a train of firing pulses is required at the gates; at this instant, it can be at any time during the half cycle.

20.

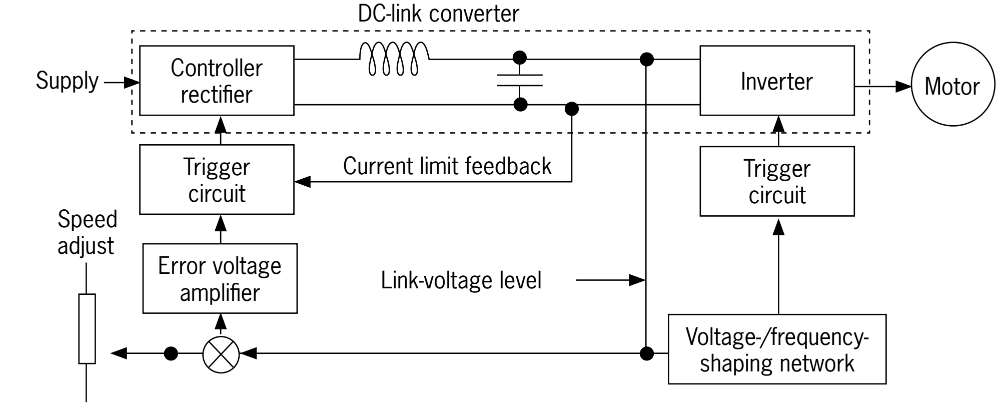


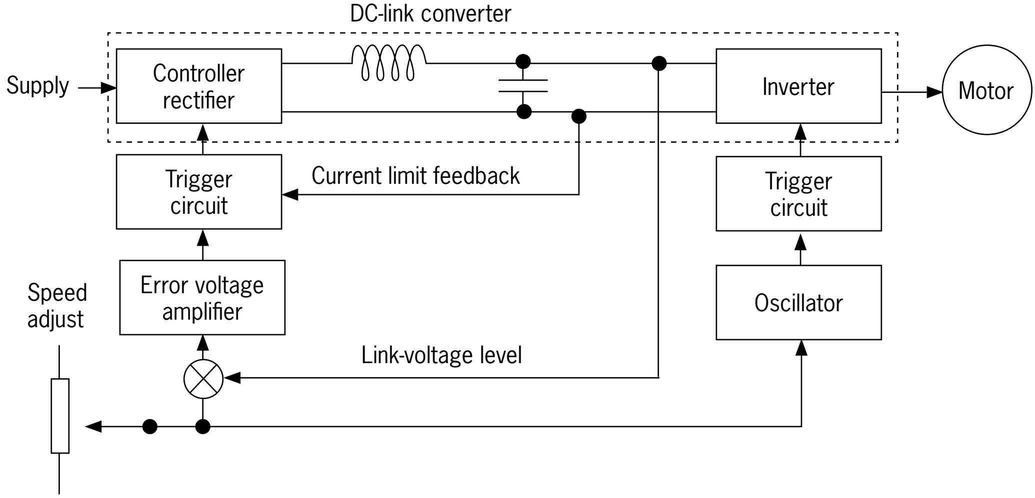
21. Thyristors are mainly used in devices where the control of high power, possibly coupled with high voltage, is demanded. Their operation makes them suitable for use in medium- to high-voltage AC power control applications, such as lamp dimming, power regulators and motor control.
22. For effective AC motor speed control, compared to that of DC motor speed control, a greater complexity of power and control equipment is required.

    With inverter-fed variable-speed drives, energy savings are possible by matching the voltage to the power demand at a given speed so as to maximise efficiency when underloaded. Practical examples in the industry where savings are possible are in pump and fan drives, where the torque is proportional to the square of the speed, with the motor loss a minimum at all speeds. To make economic sense, the annual savings in the cost of losses must exceed the annual capitalised cost of the additional power electronic equipment.
23. A DC link is a connection that connects a rectifier and an inverter found in converter circuits. The AC supply of a specific frequency is converted into DC. This DC, in turn, is converted into AC voltage. The DC link is the connection between these two circuits.

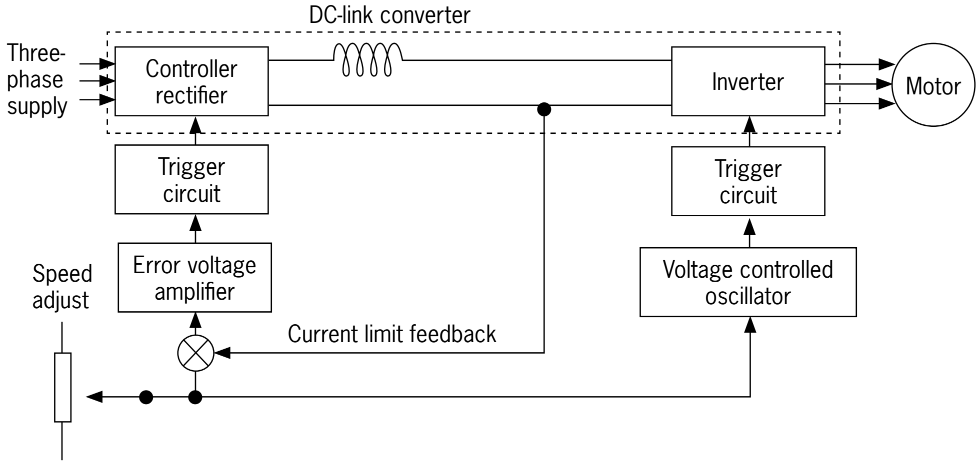
    The purpose of the DC-link capacitor is to provide a more stable DC voltage, limiting fluctuations as the inverter sporadically demands heavy current. A design can use different technologies for DC-link capacitors, such as aluminium electrolytic, film and ceramic types.

24.


DC-link converter
Supply
Controller rectifier
Inverter
Motor
Trigger circuit
Current limit feedback
Trigger circuit
Speed adjust
Error voltage amplifier
Link-voltage level
Voltage-/frequency-shaping network


DC-link converter
Supply
Controller rectifier
Inverter
Motor
Trigger circuit
Current limit feedback
Trigger circuit
Speed adjust
Error voltage amplifier
Oscillator
Link-voltage level

25.


DC-link converter
Three-phase supply
Controller rectifier
Inverter
Motor
Trigger circuit
Trigger circuit
Speed adjust
Error voltage amplifier
Voltage controlled oscillator
Current limit feedback

26. 26.1 Regenerative braking is applied to devices where there is a need to preserve energy, as is the case with battery-operated vehicles. With this type of braking, the motor's supply is removed, which turns the motor into a generator and returns the energy to the DC source.

    26.2 Plugging is the braking technique of reversing the armature connections to the supply via a resistor for the motor to rotate finally in the reverse direction. This braking technique is normally used in traction applications and high-speed drives.

    26.3 Dynamic braking is applied to those areas where rapid slowdown is required and regenerative braking is not possible. When the stop button is activated, a contactor places a fixed resistor across the armature of the motor while at the same time disconnecting the armature supply (or rectifier circuit) either mechanically or electronically. Using a fixed resistor results in a linear drop in braking torque with speed, because the motor generates a voltage proportional to the speed of rotation.

MODULE

9 Programmable logic controllers

By the end of this module, students should be able to:

- describe the operating principle of a PLC and its control unit with the aid of fully labelled block diagrams as applied to the various electronics and related industries;
- identify PLC ladder diagrams and symbols; and
- represent schematically practical examples using multi-elements and rungs with the aid of ladder diagram programming as applied to all industries employing PLCs.

Exercise 9.1 SB page 140

1. 1.1 When relays and contactors are individually wired together to solve typical control problems, it is known as hard wiring.

   1.2 When developing a program to replace the functions of a hard-wired control system, it is known as soft wiring.

   1.3 The conventional control technology employing relay and contactor controls as well as inter-wired solid state component controls are referred to as hard-wired control systems, where the wiring represents the program. When a wiring diagram of the mentioned relay and contactor controls with its associated wiring (conductors) are studied, it is referred to as relay logic.

   1.4 Once a program has been developed and 'debugged' and a listing printed, you will be looking at the completed and functional ladder logic.

2. PLCs are general-purpose microprocessor controllers designed to operate in environments such as encountered in factories, manufacturing, agriculture, star-delta start-up control, process control, traffic light control, etc.

   The PLC accepts data from input devices (sensors, limit switches, push buttons, proximity switches, etc.) and then performs logical decisions in an orderly repetitive sequence determined by the program that has been entered into its memory by the user. The PLC also provides output signals for the control of machines, plants, processes, etc.

The input modules convert electrical signals received from the input devices into logic levels for processing by the CPU (central processing unit) and the output modules convert the signals received from the CPU into the appropriate electrical signals for the control of the various field devices. The input and output (I/O) modules also provide electrical isolation of signals in the CPU from noise typically found in the factory environment.

3. **The programmer**

   This could consist of either a computer-type keyboard that is plugged into the basic unit, from which the programs are written into the memory devices of the basic unit or a hand-held unit that also plugs into the basic unit. The hand-held unit generally consists of a liquid crystal display (LCD), which displays the program as it is entered; element by element, rung by rung (line by line), page by page in ladder diagram format.

   The program can be written either into the random access memory (RAM) of the basic unit or into the programmable read-only memory (PROM) module.

   Programmable controller (basic unit)

   The basic unit is a stand-alone programmable controller. The basic unit would feature the following:

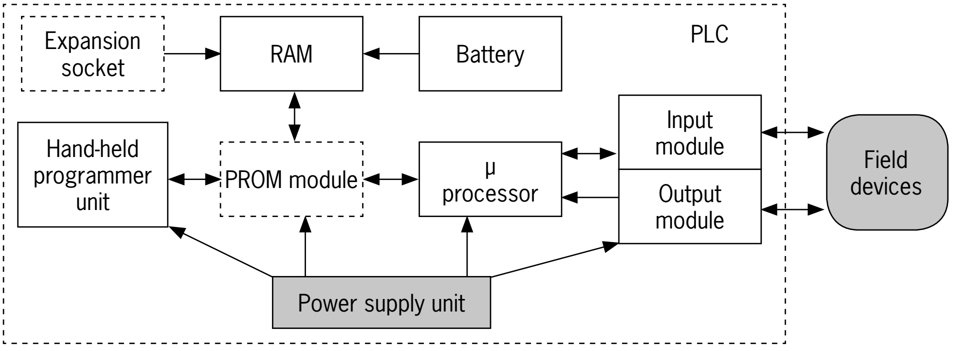
   - Terminal blocks for input and output modules
   - A power socket
   - Alarm contacts
   - A PROM module plug
   - A plug for the programmer cable
   - Expansion connectors
   - LED diagnostic indicators, etc.

   The input module enables the CPU to test the state of sensors and related detector devices connected to the PLC input terminals.

   The output module controls the power-switching devices, using the binary information stored in memory. The processor updates the state of the outputs.

   Basic units are available in various sizes, depending on the number of I/O slots allocated to the specific unit, e.g. 8 I/O unit, 16 I/O unit, 20 I/O unit or 40 I/O unit. A system requiring more I/O slots can be obtained by adding expansion units.

4.



Power supply

Provides the required power requirements

Programmer

Could be a hand-held unit or a computer keyboard

RAM and PROM

The program can be written either into the RAM of the basic unit or into the PROM module.

Microprocessor

The microprocessor is the central unit of a computer system.

Functions of the microprocessor:

- Controlling all other parts of the machine and sending timing signals
- Transferring data between memory and I/O devices
- Fetching data and instructions from memory
- Decoding instructions
- Performing arithmetical and logical operations
- Executing programs stored in memory
- Performing communication among the I/O devices

Expansion socket

Alternatively known as a bus slot or expansion port, an expansion slot is a connection or port inside a computer on the motherboard. It provides an installation point for a hardware expansion card to be connected.

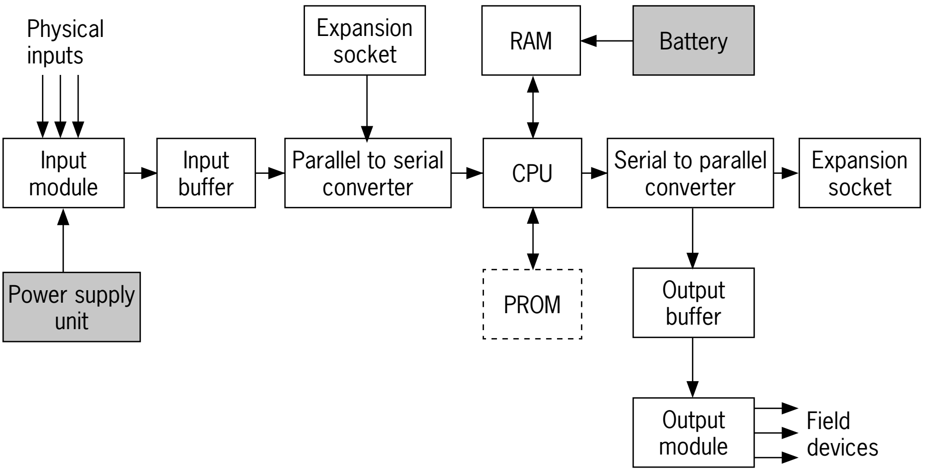
Input-output modules

An input module detects the status of input signals such as push buttons, switches, temperature sensors, etc. An output module controls devices such as relays, motor starters, lights, etc.

Field devices

Relays, motor starters, robotics, lights, etc.

5.



The CPU and its functions

The CPU is a microprocessor containing circuitry that performs logical decision-making functions; that is, it reads in the status of the control system, makes a decision based on the logic that has been programmed and then provides the actuating portion of the control system with its decision.

In order to ensure reliable operation, the CPU performs a self-test of its internal operation, which is done by a circuit referred to as a watch-dog timer. The purpose of this timer is to ensure that internal circuit and memory faults do not cause the CPU to enter an endless loop because of hardware failure. Watch-dog timers are built into all CPUs and will shut the CPU off and shut down the I/O modules if the scan time exceeds a certain predetermined time (assuming 200 ms, ± 20 ms).

Scanning begins at the very first logic element you have programmed; that is, at the top of the ladder diagram proceeding sequentially through the ladder diagram, as was programmed, and not affected by specific reference to coil numbers.

Scanning continues sequentially through the logic memory without jumping over any logic until the end of the program is reached. After the logic is solved, the I/Os are serviced. Each I/O address is taken in turn (X0, Y0, X1, Yl, etc.) with the output being the status of the just-completed logic scan. Peripheral devices, if connected, are then serviced. Finally, the CPU performs a self-test on its functions, verifying the operation of its hardware. The total scan time depends on the program size.

Memory devices

User accessible memories are normally contained in compact plug-in cartridges. Memory devices and cartridges available usually take the form of one of the following: CMOS RAM, EPROM or E^2PROM devices. Internal CMOS RAM memory devices are usually battery backed-up mounted on the inside of the base unit.

CMOS RAM memory

Complimentary metal-oxide semiconductor random access memory (CMOS RAM) is the most common type of memory used in PLCs for storing both logic and data. Although CMOS RAM is fast, low-power memory that can easily be read and written into (examined and/or changed), it is volatile memory; that is, it will lose its data when power is removed.

In order to overcome this problem and having to reload every time, the CMOS RAM memory is usually provided with a long-life battery. This battery is not used when power is applied to the system, and due to the low power drain of the CMOS RAM device, a single new lithium battery can maintain data without application of power for two to five years, with a shelf life of about 10 years.

EPROM and E^{2}PROM memories

These are programmable read-only memory devices, being low-power, fast and retentive upon loss of power, therefore not requiring a battery back-up to hold data. The data in these memories can be read at any time, but in order to change the data, some special action is required on the part of the user. These cartridges are usually programmed by inserting them into the basic unit. Both these cartridges are supplied by the manufacturer in an erased state.

Power requirements and specifications

In general, the power supplies within the basic unit will accommodate AC inputs of 115 V 50/60 Hz and/or 230 V 50/60 Hz ± 10%, and a DC input of 24 VDC. The power to the controller should be continuous, as a lengthy interrupt may terminate the operation of the program. The internal voltage source will operate from 24 VDC and would also provide a regulated + 5 VDC used to power the memory modules of the basic unit and CPU.

Noise and the input/output (I/O) circuitry

Electrical noise, such as spikes on power lines, interference picked up from field wiring and inductive kick-backs from loads, is very prevalent in industrial applications. Because the CPU operates at rather low voltage levels (5 V), this noise could have a serious effect on the operation of the PLC if it is allowed to reach the internal circuits of the CPU. Both the input and the output modules protect the CPU, using optical isolation, from electrical noise entering through the I/O modules.

6. 6.1 The CPU is a microprocessor containing circuitry that performs logical decision-making functions; that is, it reads in the status of the control system, makes a decision based on the logic that has been programmed and then provides the actuating portion of the control system with its decision.

   6.2 The purpose of a watch-dog timer is to ensure that internal circuit and memory faults do not cause the CPU to enter an endless loop because of hardware failure. Watch-dog timers are built into all CPUs and will shut the CPU off and shut down the I/O modules if the scan time exceeds a certain predetermined time (assuming 200 ms, ± 20 ms).

6.3 Scanning begins at the very first logic element you have programmed; that is, at the top of the ladder diagram and left to right, proceeding sequentially through the ladder diagram, as was programmed, and is not affected by specific reference to coil numbers. Scanning continues sequentially through the logic memory without jumping over any logic, unless specified, until the end of the program is reached.

6.4 The scan cycle is the cycle of which the PLC gathers the inputs, runs your PLC program, and then updates the outputs. This will take some time and is called the scan speed, often measured in milliseconds. The duration of the time it takes for the PLC to make one scan cycle is called the scan time of the PLC.

6.5 Complimentary metal-oxide semiconductor random access memory (CMOS RAM) is the most common type of memory used in PLCs for storing both logic and data Although CMOS RAM is fast, low-power memory that can easily be read and written into (examined and/or changed), it is volatile memory; that is, it will lose its data when power is removed.

In order to overcome this problem and having to reload every time, the CMOS RAM memory is usually provided with a long-life battery. This battery is not used when power is applied to the system, and due to the low power drain of the CMOS RAM device, a single new lithium battery can maintain data without application of power for two to five years, with a shelf life of about 10 years.

6.6 This is a programmable read-only memory devices, being low-power, fast and retentive upon loss of power, therefore not requiring a battery back-up to hold data. The data in this memory can be read at any time, but in order to change the data, some special action is required on the part of the user. This cartridge are usually programmed by inserting it into the basic unit. This cartridge is supplied by the manufacturer in an erased state.

6.7 In general, the power supplies within the basic unit will accommodate AC inputs of 115 V 50/60 Hz and/or 230 V 50/60 Hz ± 10%, and a DC input of 24 VDC. The power to the controller should be continuous, as a lengthy interrupt may terminate the operation of the program.

The internal voltage source will operate from 24 VDC and would also provide a regulated + 5 VDC used to power the memory modules of the basic unit and CPU.

6.8 Electrical noise, such as spikes on power lines, interference picked up from field wiring and inductive kick-backs from loads, is very prevalent in industrial applications. Because the CPU operates at rather low voltage levels (5 V), this noise could have a serious effect on the operation of the PLC if it is allowed to reach the internal circuits of the CPU. Both the input and the output modules protect the CPU, using optical isolation, from electrical noise entering through the I/O module.

7.

Robotics	Data collection	Injection moulding
Conveyers	Traffic control	Factory automation
Cutting	Combustion control	Extrusion
Weaving	Security systems	Sewage treatment
Bottling	Nuclear plants	Railroad switching

8. 8.1 Address

A location in the PLC memory accessible by the program

8.2 Compare

The compare function will compare two identical programs stored in two different locations. A bit-by-bit comparison is made.

8.3 Diagnostic

Display functions that are available to the operator to facilitate trouble shooting with programs and field devices

8.4 Edit

To intentionally modify the user program in a PLC

8.5 Element

A recognised schematic symbol used in a PLC program, such as contacts, coils, timers, etc.

8.6 Ladder diagram

A standard industrial method of representing electrical control schematic diagrams

8.7 Latch coil

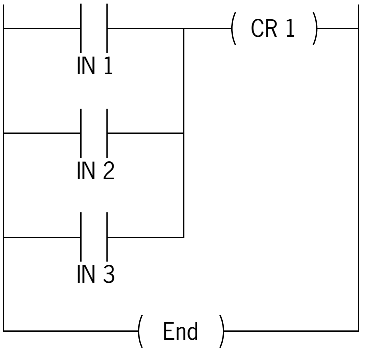
A circuit used to maintain an output coil energised after the initial energising signal was removed and requires an alternative input signal to de-energise the energised output coil

8.8 Rung

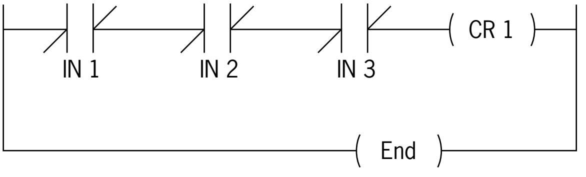
One line of a ladder diagram containing relevant programming information

8.9 A delay-off timer will be on at the commencement of the program and has a fixed predetermined period before turning off.

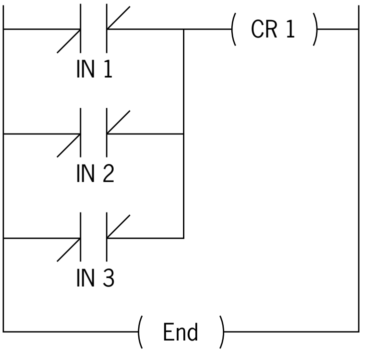
9. 9.1 OR gate



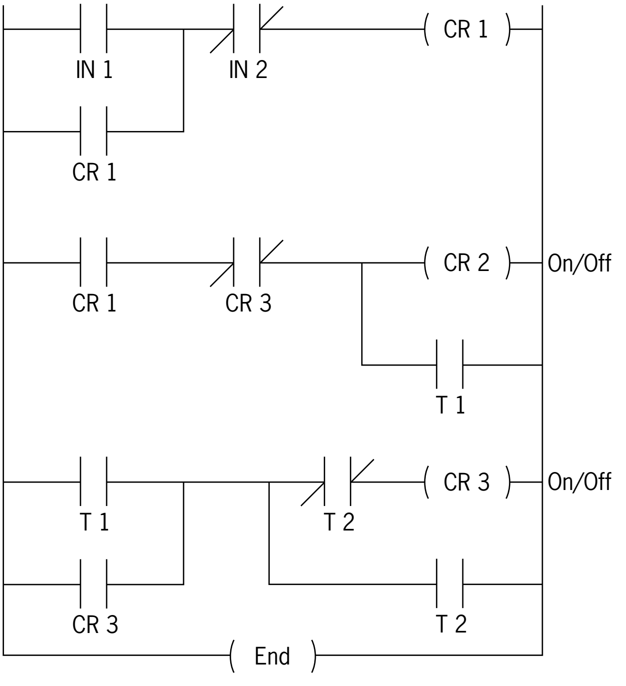
NAND gate



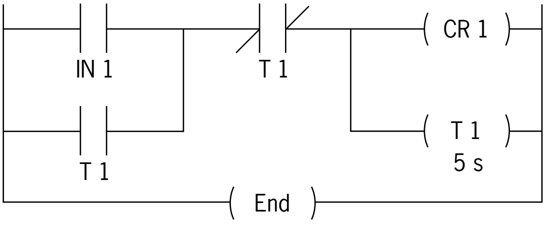
NOR gate



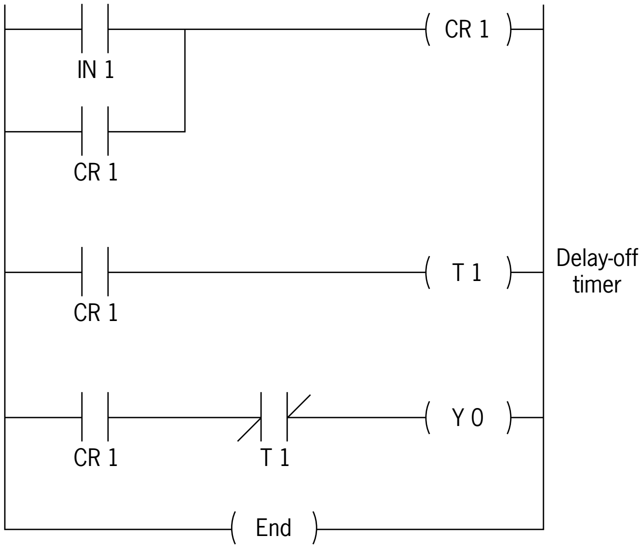
9.2


IN 1
IN 2
CR 1
CR 1
CR 1
CR 3
CR 2
On/Off
T 1
T 1
T 2
CR 3
On/Off
CR 3
T 2
End

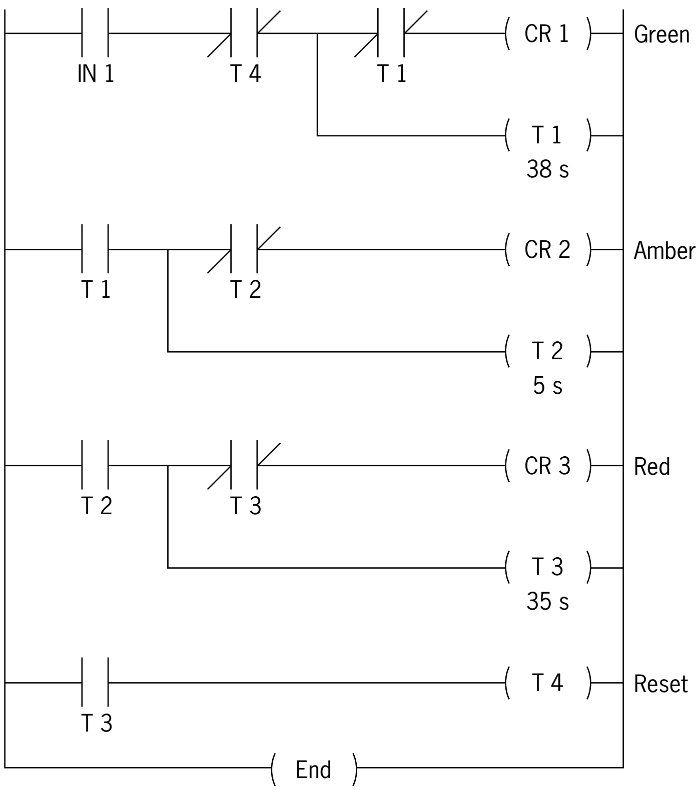
9.3


IN 1
T 1
CR 1
T 1
T 1
5 s
End

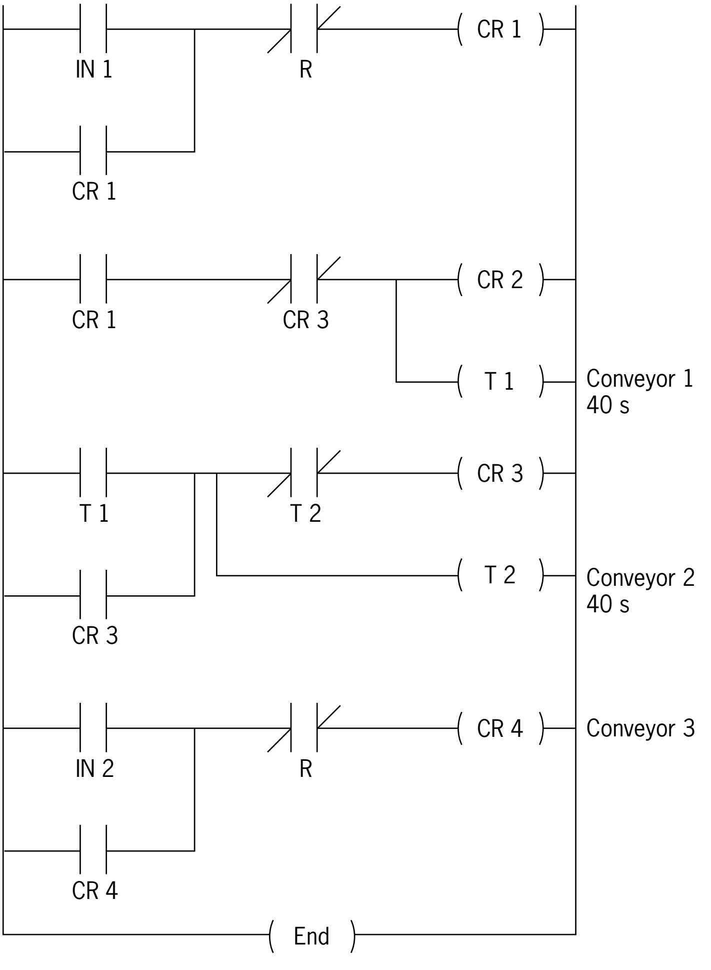
OR


CR 1
IN 1
CR 1
T 1
Delay-off timer
CR 1
Y 0
CR 1
T 1
End

9.4


CR 1
Green
IN 1
T 4
T 1
T 1
38 s
CR 2
Amber
T 1
T 2
T 2
5 s
CR 3
Red
T 2
T 3
T 3
35 s
T 4
Reset
T 3
End

10.


IN 1
R
CR 1
CR 1
CR 1
CR 3
CR 2
T 1
Conveyor 1
40 s
T 1
T 2
CR 3
CR 3
T 2
Conveyor 2
40 s
IN 2
R
CR 4
CR 4
Conveyor 3
End