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ESE 2025 : Prelims Exam
CLASSROOM TEST SERIES

E & T ENGINEERING

Test 10

Section A : Signals & Systems + Basic Electrical Engineering

Section B : Analog & Digital Communication Systems-1

Section C : Electronic Devices & Circuits-2 + Analog Circuits Topics-2

1.	(c)	**16.**	(c)	**31.**	(c)	**46.**	(c)	**61.**	(b)
2.	(d)	**17.**	(a)	**32.**	(a)	**47.**	(d)	**62.**	(b)
3.	(d)	**18.**	(a)	**33.**	(d)	**48.**	(c)	**63.**	(a)
4.	(b)	**19.**	(c)	**34.**	(c)	**49.**	(c)	**64.**	(b)
5.	(b)	**20.**	(a)	**35.**	(b)	**50.**	(c)	**65.**	(c)
6.	(c)	**21.**	(d)	**36.**	(a)	**51.**	(a)	**66.**	(c)
7.	(b)	**22.**	(b)	**37.**	(d)	**52.**	(d)	**67.**	(d)
8.	(d)	**23.**	(d)	**38.**	(a)	**53.**	(c)	**68.**	(d)
9.	(b)	**24.**	(b)	**39.**	(b)	**54.**	(c)	**69.**	(d)
10.	(c)	**25.**	(d)	**40.**	(d)	**55.**	(a)	**70.**	(a)
11.	(d)	**26.**	(a)	**41.**	(b)	**56.**	(c)	**71.**	(a)
12.	(c)	**27.**	(a)	**42.**	(d)	**57.**	(a)	**72.**	(d)
13.	(c)	**28.**	(d)	**43.**	(c)	**58.**	(c)	**73.**	(b)
14.	(c)	**29.**	(a)	**44.**	(d)	**59.**	(d)	**74.**	(b)
15.	(d)	**30.**	(d)	**45.**	(b)	**60.**	(a)	**75.**	(a)

Detailed Explanation

Section A : Signals & Systems + Basic Electrical Engineering

1. (c)

The derivative of unit impulse is known as unit doublet function.

We know,
$$\dot{\delta}(\alpha t + b) = \frac{1}{\alpha|\alpha|}\dot{\delta}\left(t + \frac{b}{\alpha}\right)$$

$$\Rightarrow \quad \dot{\delta}(5t + 10) = \frac{1}{5\times|5|}\dot{\delta}\left(t + \frac{10}{5}\right) = \frac{1}{25}\dot{\delta}(t+2) = 0.04\dot{\delta}(t+2)$$

Similarly,
$$\dot{\delta}(\alpha t) = \frac{1}{\alpha|\alpha|}\dot{\delta}(t)$$

$$\Rightarrow \quad \dot{\delta}(5t) = \frac{1}{5\times|5|}\dot{\delta}(t) = \frac{1}{25}\dot{\delta}(t) = 0.04\dot{\delta}(t)$$

2. (d)

The given system is,
$$y[m] = \sum_{k=-\infty}^{m+1} x[k]$$

$$y[m] = \ldots\ldots + x[-1] + x[0] + x[1] + \ldots\ldots\ldots + x[m] + x[m+1]$$

$\Rightarrow$ $y[m]$ depends on $x[m + 1]$ i.e. the output depends on the future values of the input. Hence, the given system is "non-causal".

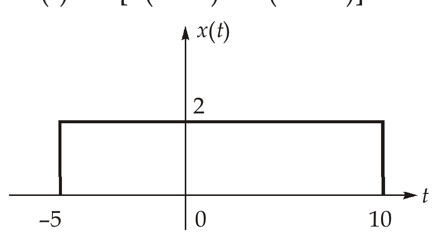
If $x[k] = 1$,
$$y[m] = \sum_{k=-\infty}^{m+1}(1) = \ldots + 1 + 1 + 1 + \ldots\ldots = \infty \text{ (for very large } m)$$

$\Rightarrow$ For a bounded input, the output is unbounded. Hence, the given system is "**unstable**".

3. (d)

Let
$$x(t) = 2\int_{-\infty}^{t}[\delta(\tau+5) - \delta(\tau-10)]d\tau$$

$$\Rightarrow \quad x(t) = 2[u(t+5) - u(t-10)]$$



$\therefore$ Energy of the signal is,
$$E_x = \int_{-\infty}^{+\infty}|x(t)|^2\,dt$$

$$= \int_{-\infty}^{+\infty}\left|2\int_{-\infty}^{t}[\delta(\tau+5) - \delta(\tau-10)]d\tau\right|^2 dt$$

$$= \int_{-5}^{10} (2)^2\, dt = 4[10-(-5)] = 60$$

$\therefore \quad E_x = 60$ Joules

4. (b)

We know, $\quad u[n] * u[n] = (n + 1)u[n]$

and $\quad x[n - \alpha] * h[n - \beta] = y[n - \alpha - \beta]$

$\therefore \quad u[n + 50] * u[n - 100] = (n + 50 - 100 + 1)u[n + 50 - 100]$

$$= (n - 49)u[n - 50]$$

At $n = 150$,

$$(n - 49)u[n - 50] = (150 - 49)u[150 - 50]$$

$$= 101u[100] = 101 \times 1 = 101$$

5. (b)

Given,

$$\longrightarrow \underbrace{\boxed{S_1(t)} \longrightarrow \boxed{S_2(t)}}_{S(t)} \longrightarrow$$

We know, the overall step response of the cascaded system is,

$$S(t) = \dot{S}_1(t) * S_2(t)$$

$$= S_1(t) * \dot{S}_2(t)$$

$$= \frac{d}{dt}[t\,u(t)] * e^{-t}u(t)$$

$$= u(t) * e^{-t}\, u(t)$$

$$= \int_{-\infty}^{t} e^{-t}u(t)dt = \int_0^t e^{-t}dt = \left[-e^{-t}\right]_0^t$$

$$= (1 - e^{-t})u(t)$$

6. (c)

Properties of Fourier series

$A - 3 \rightarrow$ Convolution in time domain leads to multiplication of Fourier Series Coefficients.

$B - 1 \rightarrow$ Multiplication in time domain leads to convolution in frequency domain.

$C - 4 \rightarrow$ Parseval's relation: The total energy of a periodic signal is equal to the sum of the squares of the magnitudes of its Fourier coefficients.

$D - 2 \rightarrow$ Cross correlation Theorem: The cross-correlation between two signals is equal to the product of Fourier series coefficient of one signal multiplied by complex conjugate of fourier series coefficient of another signal.

7. (b)

Given, $\quad x(t) = e^{-t/2}; \quad 0 < t < \pi$

The exponential Fourier series coefficient is,

$$C_n = \frac{1}{T}\int_0^T x(t)\, e^{-jn\omega_0 t}\, dt, \quad \text{where } \omega_0 = \frac{2\pi}{T}$$

$$= \frac{1}{\pi}\int_0^{\pi}(e^{-t/2})\cdot e^{-jn\frac{2\pi}{\pi}t}\,dt = \frac{1}{\pi}\int_0^{\pi} e^{-\left(\frac{1}{2}+jn2\right)t}\,dt$$

$$= \frac{1}{\pi}\left[\frac{e^{-\left(\frac{1}{2}+jn2\right)t}}{-\left(\frac{1}{2}+jn2\right)}\right]_0^{\pi} = \frac{1}{\pi}\left[\frac{1-e^{-\left(\frac{1}{2}+jn2\right)\pi}}{\left(\frac{1}{2}+jn2\right)}\right]$$

$\therefore$ $$|C_n\cdot\delta(n)| = |C_0\cdot\delta(n)| = |C_0| = \left|\frac{1}{\pi}\left[\frac{1-e^{-\frac{\pi}{2}}}{\frac{1}{2}}\right]\right|$$

$\Rightarrow$ $$|C_n\cdot\delta(n)| = \frac{2}{\pi}\left(1-e^{-\pi/2}\right)$$

8. (d)

Given, $$x(t) = [1 + \cos \pi t];\ -1 \le t \le 1$$

$\Rightarrow$ $$x(t) = [1+\cos\pi t]\,\text{rect}\left(\frac{t}{2}\right)$$

$\Rightarrow$ $$x(t) = \left[1+\frac{e^{j\pi t}+e^{-j\pi t}}{2}\right]\text{rect}\left(\frac{t}{2}\right)$$

$\Rightarrow$ $$x(t) = \text{rect}\left(\frac{t}{2}\right)+\frac{1}{2}\text{rect}\left(\frac{t}{2}\right)\cdot e^{j\pi t}+\frac{1}{2}\text{rect}\left(\frac{t}{2}\right)e^{-j\pi t}$$

We know, $$A\,\text{rect}\left(\frac{t}{T}\right)\longleftrightarrow ATSa\left(\frac{\omega T}{2}\right)$$

$\Rightarrow$ $$\text{rect}\left(\frac{t}{2}\right)\xleftrightarrow{F.T} 2Sa(\omega)$$

$\therefore$ The Fourier transform of $x(t)$ using frequency shifting property is,

$$\left[x(t)e^{j\omega_0 t}\longleftrightarrow X(\omega-\omega_0)\right]$$

$$X(\omega) = 2Sa(\omega)+\frac{1}{2}[2Sa(\omega-\pi)+2Sa(\omega+\pi)]$$

$\Rightarrow$ $$X(\omega) = 2Sa(\omega) + Sa(\omega - \pi) + Sa(\omega + \pi)$$

9. (b)

The periodic signal can be expressed using Fourier series as

$$x(t) = \sum_{n=-\infty} C_n e^{jn\omega_0 t}$$

We know that,

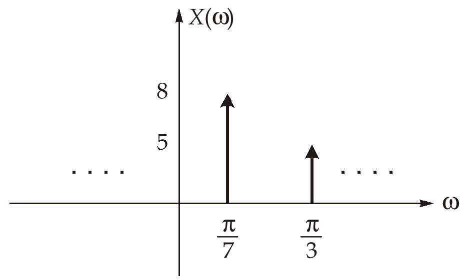
$$e^{jn\omega_0 t}\longleftrightarrow 2\pi\delta(\omega-n\omega_0)$$

Fourier transform of periodic signal is,

$$X_p(\omega) = 2\pi \sum_{n=-\infty}^{\infty} C_n \delta(\omega - n\omega_0)$$

Given spectrum,

$$X(\omega) = 8\delta\left(\omega - \frac{\pi}{7}\right) + 5\delta\left(\omega - \frac{\pi}{3}\right)$$



The frequency components in the spectrum are integer multiples of the fundamental frequency ω_0. Thus, the fundamental frequency,

$$\omega_0 = GCD\left(\frac{\pi}{7}, \frac{\pi}{3}\right) = \frac{\pi}{21}$$

$\Rightarrow$ The fundamental period,

$$T = \frac{2\pi}{\omega_0} = \frac{2\pi}{\left(\frac{\pi}{21}\right)} = 42 \text{ sec}$$

10. (c)

Given, $$x(t) = \cos(2\pi t) = \frac{1}{2}e^{j2\pi t} + \frac{1}{2}e^{-j2\pi t}$$

The Fourier series coefficients of $x(t)$ are,

$$C_1 = \frac{1}{2}, C_{-1} = \frac{1}{2}$$

Similarly, $$y(t) = 0.5\cos\left(2\pi t - \frac{\pi}{2}\right)$$

$$= 0.5\left[\cos(2\pi t)\cdot\cos\frac{\pi}{2} + \sin(2\pi t)\cdot\sin\frac{\pi}{2}\right]$$

$$= 0.5\sin(2\pi t)$$

$$= 0.5\left[\frac{e^{j2\pi t} - e^{-j2\pi t}}{2j}\right]$$

$$= \frac{1}{4j}e^{j2\pi t} - \frac{1}{4j}e^{-j2\pi t}$$

The Fourier series coefficients of $y(t)$ are,

$$d_1 = \frac{1}{4j}, d_{-1} = \frac{-1}{4j}$$

The cross-correlation between two signals is equal to the product of Fourier series coefficient of one signal multiplied by complex conjugate of Fourier series coefficient of another signal. Therefore, the Fourier series coefficients of $r_{xy}(\tau)$ are $c_n^* d_n$ given as,

$$c_1^* d_1 = \frac{1}{2} \times \frac{1}{4j} = \frac{1}{8j}$$

$$c_{-1}^* d_{-1} = \frac{1}{2} \times \frac{-1}{4j} = \frac{-1}{8j}$$

11. (d)

Given signal, $$x(t) = \sin\omega_0 t \cdot u(t-\tau)$$

$\Rightarrow$ $$x(t) = \sin\omega_0(t-\tau+\tau) \cdot u(t-\tau)$$
$$= [\sin\omega_0(t-\tau) \cdot \cos\omega_0\tau + \cos\omega_0(t-\tau) \cdot \sin\omega_0\tau]u(t-\tau)$$
$$= [\sin\omega_0(t-\tau) \cdot u(t-\tau)]\cos\omega_0\tau + [\cos\omega_0(t-\tau) \cdot u(t-\tau)]\sin\omega_0\tau$$

We know that, $\sin(\omega_0 t)u(t) \longleftrightarrow \frac{\omega_0}{s^2+\omega_0^2}$ and $\cos(\omega_0 t)u(t) \longleftrightarrow \frac{s}{s^2+\omega_0^2}$

Taking Laplace transform on both sides and using the time-shifting property,

$\Rightarrow$ $$X(s) = \frac{\omega_0 e^{-s\tau}}{s^2+\omega_0^2} \cdot \cos\omega_0\tau + \frac{se^{-s\tau}}{s^2+\omega_0^2} \cdot \sin\omega_0\tau$$

$\Rightarrow$ $$X(s) = \left[\frac{\omega_0 \cos\omega_0\tau + s\sin\omega_0\tau}{s^2+\omega_0^2}\right] \cdot e^{-s\tau}$$

12. (c)

We know, $$r(t) \longleftrightarrow \frac{1}{s^2}$$

Using the time-shifting property,

$$r(t-2) \longleftrightarrow \frac{e^{-2s}}{s^2}$$

Using the time-scaling property,

$$r\left(\frac{1}{3}t-2\right) \longleftrightarrow \frac{1}{\left|\frac{1}{3}\right|} \frac{e^{-2(3s)}}{(3s)^2}$$

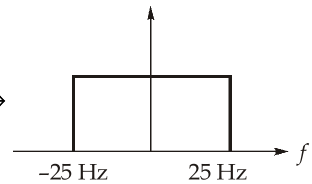
$\therefore$ $$G(s) = \frac{3e^{-6s}}{9s^2} = \frac{e^{-6s}}{3s^2}$$

$\Rightarrow$ $$G(2) = \frac{e^{-6(2)}}{3(2)^2} = \frac{e^{-12}}{12}$$

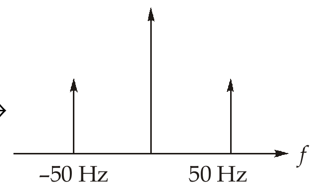
13. (c)

We have $$\text{sinc}(at) \longleftrightarrow \frac{1}{|a|} rect\left(\frac{f}{a}\right)$$

Therefore, $\text{sinc}(50t) \longleftrightarrow$



$\cos(100\pi t) \longleftrightarrow$



$$x_1(t) = 5\ \text{sinc}(50t) \cdot \cos(100\pi t)$$

The multiplication in time domain leads to convolution in frequency domain. Thus, $X_1(f)$ is convolution of two signals with spectrum extending from (–50 – 25) Hz to (50 + 25) Hz i.e. –75 Hz to 75 Hz.

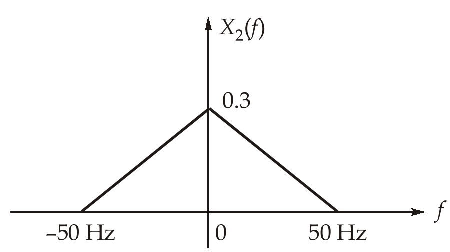
$\therefore$ Nyquist sampling rate for $x_1(t)$,

$$M = 2 \times 75$$
$$= 150 \text{ Hz}$$

Consider, $$x_2(t) = 15\ \text{sinc}^2(50t)$$

We have, $$A\text{sinc}^2(at) \longleftrightarrow \frac{A}{|a|} tri\left(\frac{f}{a}\right)$$

Taking Fourier transform, $$X_2(f) = \frac{15}{50} tri\left(\frac{f}{50}\right) = 0.3 tri\left(\frac{f}{50}\right)$$



$\therefore$ Nyquist sampling rate for $x_2(t)$,

$$N = 2 \times 50 = 100 \text{ Hz}$$

$$\Rightarrow \quad \frac{M}{N} = \frac{150}{100} = \frac{3}{2}$$

14. (c)

We know, $$u[n] \xleftrightarrow{DTFT} \pi\delta(\omega) + \frac{1}{1 - e^{-j\omega}};\ -\pi < \omega < \pi$$

$$\Rightarrow \quad U(e^{j\omega}) = \pi\delta(\omega) + \frac{1}{1 - e^{-j\omega}}$$

$$= \pi\delta(\omega) + \frac{1}{1-\cos\omega + j\sin\omega}$$

$$= \pi\delta(\omega) + \frac{(1-\cos\omega) - j\sin\omega}{(1-\cos\omega)^2 + \sin^2\omega}$$

$$= \pi\delta(\omega) + \frac{1-\cos\omega - j\sin\omega}{2(1-\cos\omega)}$$

$$= \pi\delta(\omega) + \frac{1}{2} - \frac{1}{2} j \frac{\sin\omega}{(1-\cos\omega)}$$

$$= \pi\delta(\omega) + \frac{1}{2} - \frac{1}{2} j \frac{2\sin\left(\frac{\omega}{2}\right)\cos\left(\frac{\omega}{2}\right)}{2\sin^2\left(\frac{\omega}{2}\right)}$$

$$= \pi\delta(\omega) + \frac{1}{2} - \frac{1}{2} j\cot\left(\frac{\omega}{2}\right)$$

$\therefore$
$$Re[U(e^{j\omega})] = \frac{1}{2} + \pi\delta(\omega)$$

$$Im[U(e^{j\omega})] = -\frac{1}{2}\cot\left(\frac{\omega}{2}\right)$$

15. (d)

Given,
$$x[n] = (0.5)^n u[n] + 2^n u[-n-1]$$

$\Rightarrow$
$$x[n] = (0.5)^n u[n] + \left(\frac{1}{2}\right)^{-n} u[-n-1]$$

Take Fourier transform on both sides,

$$X(e^{j\omega}) = \frac{1}{1-0.5e^{-j\omega}} + \frac{0.5e^{j\omega}}{1-0.5e^{j\omega}}$$ ... using the time-shifting and time reversal property of DTFT

$$= \frac{1}{1-0.5\cos\omega + j0.5\sin\omega} + \frac{0.5\cos\omega + j0.5\sin\omega}{1-0.5\cos\omega - j0.5\sin\omega}$$

$$= \frac{1-0.5\cos\omega - j0.5\sin\omega + 0.5\cos\omega + j0.5\sin\omega - (0.5)^2}{(1-0.5\cos\omega)^2 + (0.5\sin\omega)^2}$$

$$= \frac{1-(0.5)^2}{1+(0.5)^2 - \cos\omega} = \frac{0.75}{1.25-\cos\omega}$$

At $\omega = \frac{\pi}{3}$,

$$X\left(e^{j\frac{\pi}{3}}\right) = \frac{0.75}{1.25-\cos\frac{\pi}{3}} = \frac{0.75}{(1.25-0.5)} = \frac{0.75}{0.75} = 1$$

16. (c)

We know that

$$\cos\alpha = 1-\frac{\alpha^2}{2!}+\frac{\alpha^4}{4!}-\frac{\alpha^6}{6!}+\frac{\alpha^8}{8!}-......$$

$$= \sum_{k=0}^{\infty}\frac{(-1)^k}{(2k)!}\alpha^{2k}$$

Thus, $X(z) = \cos(z^{-2})$

$$= \sum_{k=0}^{\infty}\frac{(-1)^k}{(2k)!}(z^{-2})^{2k}; |z|>0$$

Taking inverse z-transform, we get

$$x[n] = \sum_{k=0}^{\infty}\frac{(-1)^k}{(2k)!}\delta[n-4k]$$

For $n = 4$, we get, $4 - 4k = 0 \Rightarrow k = 1$

Thus, $x[4] = \dfrac{(-1)}{(2)!} = -0.5$

17. (a)

Let $y[n] = nx[n] = (-5)^{-n}u[-n-1] = \left(\frac{-1}{5}\right)^n u[-n-1]$

Take z-transform on both sides,

$$Y(z) = \frac{-z}{z+\frac{1}{5}};\ \text{ROC}: |z| < \frac{1}{5}$$

Using the multiplication by 'n' property of z-transform, we have

$$y[n] = nx[n]$$

$$\Rightarrow \quad Y(z) = -z\frac{d}{dz}X(z)$$

We have, $-z\dfrac{d}{dz}X(z) = \dfrac{-z}{z+\frac{1}{5}}$

$$\Rightarrow \quad \frac{d}{dz}X(z) = \frac{1}{z+\frac{1}{5}}$$

$$\therefore \quad X(z) = \log\left(z+\frac{1}{5}\right);\ \text{ROC}: |z| < \frac{1}{5}$$

18. (a)

Window Type	Main Lobe Transition width (Approximate)
Rectangular	4π
Hamming	$\frac{8\pi}{N}$
Hanning	$\frac{8\pi}{N}$
Blackman	$\frac{12\pi}{N}$
Bartlett	$\frac{8\pi}{N}$

19. (c)

Let the flux at any instant be given by

$$\phi = \phi_m \sin \omega t \text{ Weber} \quad \text{...(i)}$$

The instantaneous emf induced in a coil of 'T' turns linked by this flux is given by Faraday's law as

$$e = -\frac{d}{dt}(\phi T) \text{ volt}$$

$$e = -\frac{Td\phi}{dt} = \frac{-Td(\phi_m \sin \omega t)}{dt} \text{Volt}$$

$$e = -T\omega\phi_m \cos \omega t \text{ volt}$$

$$e = E_m \sin\left(\omega t - \frac{\pi}{2}\right) \quad \text{...(ii)} \qquad (E_m = T\omega\phi_m)$$

On comparing (i) and (ii), we get

Induced emf lags driving magnetic flux by $\left(\frac{\pi}{2}\right)$ radian.

20. (a)

Method-I:

We have,

$$\text{Transformation ratio, } a = \frac{E_1}{E_2} = \frac{T_1}{T_2} = 5 \quad \text{...(i)}$$

Number of secondary turns, $T_2 = 10$ ...(ii)

Sum of primary voltage and secondary voltage,

$$E_1 + E_2 = 1200 \quad \text{...(iii)}$$

From equation (i), we get

$$E_1 = 5E_2 \quad \text{...(iv)}$$

And from equation (iii) and (iv), we get

$$5E_2 + E_2 = 1200$$

$$E_2 = 200 \text{ volt}$$

From equation (iii), $E_1 + E_2 = 1200$

$\Rightarrow$ $E_1 = 1200 - 200$

$E_1 = 1000$ Volt

Therefore, $E_1 - E_2 = 1000 - 200$

$= 800$ volt

Method-II:

We have,

$$\text{Transformation ratio, } a = \frac{E_1}{E_2} = \frac{T_1}{T_2} = 5 \quad \text{...(i)}$$

and sum of primary voltage and secondary voltage, $E_1 + E_2 = 1200$ V ...(ii)

Thus, $E_1 = 1200 - E_2$

Now, from equation (i),

$$\frac{1200 - E_2}{E_2} = 5$$

$$E_2 = 200 \text{ Volt}$$

From equation (ii),

$$E_1 = 1000 \text{ Volt}$$

Hence, $E_1 - E_2 = 1000 - 200 = 800$ Volt

21. **(d)**

An ideal transformer has the following properties:

- The primary and secondary winding resistance are negligible.
- The core has infinite permeability (μ) so that no mmf is required to establish the flux in the core.
- The leakage flux and leakage inductances are zero. The entire flux is confined to the core and links both windings.
- There are no losses due to resistance, hysteresis and eddy currents. Thus, efficiency is 100 percent.

In an ideal transformer, $V_1 I_1 = V_2 I_2$. Hence, the kVA input of an ideal transformer is equal to the kVA output i.e., kVA is the same on both sides of the transformer.

A transformer is said to be on no load when the secondary winding is open-circuited. The secondary current is thus zero. When an alternating voltage is applied to the primary, a small current I_0 flows in the primary. The current I_0 is called the no-load current of the transformer. It is made up of two components I_μ and I_W. The component I_μ is called the magnetizing component. It magnetizes the core. In other words, it sets up a flux in the core and therefore I_μ is in phase with ϕ_M. The current I_μ is also called reactive, or wattless component of no-load current.

The component I_W is responsible for the hysteresis and eddy-current losses in the core and the negligible I^2R loss in the primary winding. The current I_W is called the active component or wattful component of no-load current. It is in the phase with the applied voltage V_1. The no-load current I_0 is small of the order of 3 to 5 percent of the rated current of the primary.

22. (b)

A rotating electric machine has two main parts, stator and rotor, separated by the air gap. The stator of the machine doesn't move and normally is the outer frame of the machine. The rotor is free to move and normally is the inner part of the machine.

Both stator and rotor are made of ferromagnetic materials. Slots are cut on the inner periphery of the stator and the outer periphery of the rotor. Conductors are placed in the slots of the stator or rotor. They are interconnected to form windings. The winding in which voltage is induced is called the armature winding. The winding through which current is passed to produce the main flux is called the field winding. The field winding is placed on the stator whereas the armature winding is placed on the rotor.

23. (d)

We know that,

emf generated is given as $E = \dfrac{N\phi z}{60} \times \dfrac{P}{A}$

For lap winding, $A = P$

$$E_1 = \frac{N\phi z}{60} = \frac{N \times 0.060 \times 500}{60} \quad \text{...(i)}$$

For wave winding, $A = 2$

$$E_2 = \frac{N\phi z}{60} \times \frac{P}{2}$$

$$= \frac{N \times 0.060 \times 500}{60} \times \frac{8}{2} \quad \text{...(ii)}$$

Now,

According to question, $E_2 - E_1 = 1500$ V

From equation (i) and (ii),

$$\left(\frac{N \times 0.060 \times 500}{60} \times 4\right) - \left(\frac{N \times 0.060 \times 500}{60}\right) = 1500$$

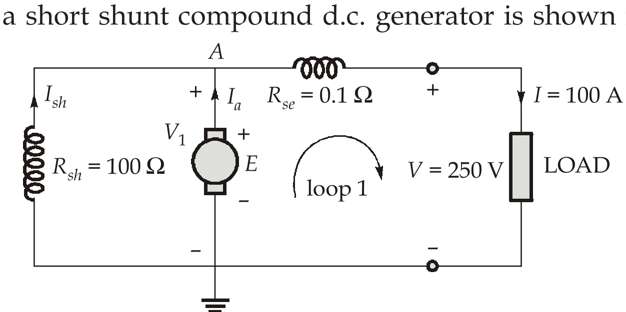
$$2N - \frac{N}{2} = 1500$$

$$\frac{3N}{2} = 1500$$

$$N = 1000 \text{ rpm}$$

24. (b)

The circuit diagram of a short shunt compound d.c. generator is shown in fig. below:



On applying KVL in loop 1, we get

$$-E + (2 \times 0.5) + (I_a \times 0.1) + (0.1 \times 100) + 250 = 0$$

$$E = 250 + 1 + 0.1I_a + 10$$

$$E = 261 + 0.1I_a \quad \text{...(i)}$$

Now, on applying KVL in loop 1 again, we get

$$V_1 = (0.1 \times 100) + 250$$

$$V_1 = 260 \text{ Volt} \quad \text{...(ii)}$$

$$\therefore \quad I_{sh} = \frac{V_1}{R_{sh}}$$

$$I_{sh} = \frac{260}{100} = 2.6\text{A}$$

Thus, $I_a = I + I_{sh}$

$$I_a = 100 + 2.6$$

$$I_a = 102.6 \text{ A}$$

Hence, $E = 0.1I_a + 261$

$$E = (0.1 \times 102.6) + 261$$

$$E = 271.26 \text{ Volt}$$

25. (d)

The power factor of a synchronous motor depends on the excitation:

- under excited: lagging power factor
- over-excited: leading power factor
- normal excitation: unity power factor

26. (a)

For successful parallel operation of synchronous generators, they must have the same voltage, frequency and phase sequence. Power factor is not a necessary condition.

27. (a)

Pressurized Water Reactor: This is the most widely used type of nuclear reactor worldwide, where water is kept under pressure to prevent it from boiling and used to transfer heat to a secondary system. PWRs use ordinary water as both coolant and moderator and thus, is most comonly used in commercial nuclear power plants.

28. (d)

- Lead-acid batteries have a lower energy density compared to lithium-ion batteries.
- Lead-acid batteries typically have a lower cycle life compared to lithium-ion batteries, meaning they need to be replaced more frequently.
- Lead-acid batteries are less expensive upfront but have higher operational costs over time due to their lower efficiency, shorter cycle life, and greater need for maintenance. Lithium-ion batteries, though more expensive initially, offer reduced long-term costs due to lower maintenance needs and longer operational life.

Hence, only statement 2 and 3 are correct.

29. (a)

Advantages of wound rotor are

- High starting torque and low starting current.
- Additional resistance can be connected in the rotor circuit to control speed.

Advantages of the squirrel cage rotor are as follows:

- Robust construction and cheaper.
- The absence of brushes reduces the risk of sparking.
- Lesser maintenance.
- Higher efficiency and higher power factor.

Therefore, statement 2 is not correct.

30. (d)

We have,

Motor speed, $N = 1755$ rpm

rotor frequency, $f_r = 1.5$ Hz

We know that, $N = (1 - s)N_s$ ...(i) where s = slip and N_s = synchronous speed

and $f_r = sf_s$ ...(ii)

where $f_s = \frac{N_s P}{120} = \frac{4N_s}{120}$

$$f_s = \frac{N_s}{30} \quad \text{...(iii)}$$

From (ii) and (iii), we get

$$f_r = 1.5 = \frac{sN_s}{30} \text{ i.e., } sN_s = 45 \quad \text{...(iv)}$$

Now, from (i) and (iv), we get

$$N = N_s - sN_s = 1755 \text{ rpm}$$

$\Rightarrow$ $N_s = N + sN_s$

$$N_s = 1755 + 45 = 1800 \text{ rpm}$$

$\therefore$ $N_s = \frac{120P}{f_s}$

$$\frac{120 \times 4}{f_s} = 1800$$

$$f_s = 60 \text{ Hz}$$

From (ii), we get $1.5 = s \times 60$

$$s = \frac{1.5}{60} = 0.025$$

Thus, option (d) is correct.

31. (c)

At standstill condition i.e. when rotor is at rest ($N_r = 0$)

$$s = \frac{N_s - N_r}{N_s} = \frac{N_s - 0}{N_s} = 1$$

Thus, option (c) is correct.

32. (a)

According to given data,

Maximum torque, $\tau_{max} = \tau_{fl}$, torque at full load

We know that

$$\frac{\tau_{max}}{\tau_{fl}} = \frac{s_{fl}^2 + s_m^2}{2s_{fl}\, s_m}$$

where, s_{fl} = slip at full-load torque; s_m = slip at maximum torque.

$$1 = \frac{s_{fl}^2 + s_m^2}{2s_{fl}\, s_m}$$

$$s_{fl}^2 + s_m^2 = 2s_{fl}\, s_m$$

$$(s_{fl}^2 - s_m)^2 = 0$$

$$s_{fl} = s_m$$

Given, $s_{fl} = 4\%$

Thus, $s_m = \frac{R_2}{X_2} = \frac{4}{100}$

$$R_2 = 0.04X_2$$

From the given options, the above relation is satisfied by

Rotor impedance = $(0.01 + 0.25j)\ \Omega$

Thus, option (a) is correct.

33. (d)

- Short Circuit Test of Transformer is performed on HV side and the supply voltage is so adjusted that rated current flows through the shorted secondary. It is, thus, used to determine the impedance and losses of the transformer.
- The supply voltage required for the flow of rated current in the shorted secondary is around 2-12% of rated voltage. Thus, the core loss is negligible in the short circuit test because the applied voltage is low and does not magnetize the core significantly.
- The magnetizing reactance is determined from the open circuit test, not the short circuit test.

34. (c)

We have,Power rating of the motor,

$$P_{rated} = 5 \text{ kW}$$

Supply voltage, $V = 400$ V

frequency, $f = 50$ Hz

slip, $s = 4\% = 0.04$

Efficiency, $\eta = 0.85$

We know that,

$$\eta = \frac{P_{output}}{P_{input}} = 0.85$$

$$P_{\text{input}} = \frac{5}{0.85} = 5.88 \text{ kW}$$

As, $$P_{\text{rotor loss}} = s.\, P_{\text{input}}$$
$$= 0.04 \times 5.88 = 0.23 \text{ kW}$$

35. (b)

- In induction motor, rotor is not supplied with DC. The stator generates the rotating magnetic field, and the rotor current is induced from this field.
  Whereas, in synchronous motor, the rotor gets DC excitation.
- In squirrel-cage induction motor, the rotor is not externally energized by AC. The voltage/current is induced by the stator's rotating magnetic field.
- In a synchronous motor, the rotor and stator both rotates at the same speed, with the rotor being excited by a DC supply, and the stator generating a rotating magnetic field using an AC supply.

Therefore, option (b) is correct.

36. (a)

Using impulse invariant technique, analog frequencies separated by multiples of $2\pi/T_s$ map to the same digital frequency. Thus, it is limited to design of Low Pass and Band Pass filters only.

37. (d)

Transformer is a static device which consist of two or more stationary electric circuit interlinked by a common Magnetic circuit for the purpose of transferring electrical energy between them. Since, there is no relative motion between the coils, the frequency of induced voltage in coil 2 is exactly the same as the frequency of the applied voltage to coil 1.

38. (a)

Heavy water (D_2O) is an effective moderator as it slows down neutrons without absorbing them excessively. Its higher density compared to normal water contributes to this property.

Section B : Analog and Digital Communication Systems-1

39. (b)

Given data, $$f_m = 3 \text{ kHz}$$
$$\Delta f = 6 \text{ kHz}$$

For FM signal, $$\Delta f = k_f A_m$$

Given, $$A'_m = 4 A_m$$

$\therefore$ $$\Delta f' = 4k_f A_m = 4 \times 6 = 24 \text{ kHz}$$

Using Carson's rule, $$\text{B.W (new)} = (\Delta f' + f'_m)2 = \left(24 + \frac{3}{2}\right)2 = 51 \text{ kHz}$$

40. (d)

- A PLL locks onto the phase and frequency of an incoming signal making it effective for tracking.
- The loop filter bandwidth directly impacts how quickly the PLL can respond to changes (tracking speed) and reject high-frequency noise.

- The VCO's frequency varies depending on the control voltage, allowing the PLL to adjust to match the frequency and phase of the incoming signal.
- Higher loop gain can widen the lock range and reduce pull-in time, improving the PLL's ability to quickly lock onto the signal.

41. (b)

The inverse fourier transform of the power spectral density gives the auto-correlation function. We get,

$$R_x(\tau) = F^{-1}\left[\frac{S_0}{2}\right] = \frac{S_0}{2}\delta(\tau)$$

The auto-correlation function at the output is given by

$$R_y(\tau) = R_x(\tau) * h(\tau) * h(-\tau)$$

Since, $P = R(0) = \frac{S_0}{2T^2}[T-0] = \frac{S_0}{2T} = 10S_0$

So, $$T = \frac{1}{20}$$

42. (d)

For a stationary random process:

(a) The mean of $X(t)$ is constant (independent of time t)

(b) The autocorrelation function $R_X(\tau)$ depends only on the time difference τ, not the absolute time.

Since the random process $X(t)$ has zero mean and auto-correlation function, $R_X(\tau) = e^{-|\tau|}$; depends only on τ, $X(t)$ is a stationary random process.

- The Fourier transform of auto-correlation function $R_X(\tau)$ gives the power spectral density $S_X(f)$. Thus,

$$R_X(\tau) = e^{-|\tau|} \xleftrightarrow{F.T} S_X(f) \frac{2}{1+(2\pi f)^2} = \frac{2}{1+4\pi^2 f^2}$$

- The variance of $X(t)$ is defined as $\sigma_X^2 = R_X(0)$ where $R_X(0)$ is the autocorrelation function evaluated at $\tau = 0$.

$$R_X(\tau) = e^{-|\tau|}$$

$$R_X(0) = e^0 = 1;\ \text{Thus } \sigma_X^2 = 1$$

43. (c)

- In TV broadcasting (e.g. analog systems like NTSC or PAL), VSB is used to transmit the video signal efficiently. The video signal has a wide bandwidth, so using standard AM would require a large bandwidth (twice the baseband signal bandwidth) SSB, despite having less bandwidth is not used for TV broadcasting due to receiver complexity.
  - VSB reduces the bandwidth by transmitting one full sideband and a small portion (vestige) of the other sideband, along with the carrier.
  - This saves bandwidth compared to conventional AM while still allowing for simpler demodulation

- The defining feature of VSB is that it transmits:
  - One full sideband.
  - A portion (or vestige) of the other sideband.
  - The carrier, for demodulation purpose.

  This design helps retain the advantages of both SSB (bandwidth saving) and standard AM (ease of demodulation).
- Since VSB signals always contain a portion of the second sideband, the bandwidth of VSB is greater than SSB but less than AM.
- In VSB demodulation, the receiver uses a filter to remove any excessive components of the vestigial sideband that may interfere with signal reconstruction.

44. (d)

- The bandwidth requirements for PM and FM are dependent on the modulation index: In general, PM and FM signals require approximately the same bandwidth when properly designed, but PM does not inherently require more bandwidth than FM for the same signal.
- Phase modulation (PM) is the basis of digital phase-shift keying (PSK) technique. In PSK, the phase of the carrier is discretely changed based on the transmitted data bits (e.g. BPSK, QPSK). PM and PSK are closely related because PSK can be considered as a discrete form of PM.
- PM is a non-linear modulation scheme because the modulating signal varies the phase of the carrier, creating a non-linear relationship between the modulating and modulated signal. In contrast, linear modulation schemes (e.g., AM) involve linear variations of the amplitude maintaining a linear relationship. Thus, PM is not a linear modulation scheme.

  The PM signal is given as

  $$S_{PM}(t) = A_c\cos(2\pi f_c t + K_p m(t))$$

  and the FM signal as,

  $$S_{FM}(t) = A_c \cos\left(2\pi f_c t + 2\pi K_f \int_{-\infty}^{t} m(t)dt\right)$$

  Thus, integrating the modulating signal and then using a phase modulator gives a FM signal.

45. (b)

The frequency multiplier ratio, n is

$$n = \frac{(\Delta f)_{WBFM}}{(\Delta f)_{NBFM}} = \frac{12}{0.1} = 120$$

The carrier frequency at the output of multiplier is

$$f_c' = nf_c = 120 \times 120 \text{ kHz} = 14400 \times 10^3 \text{ Hz}$$
$$= 14.4 \text{ MHz}$$

To shift the carrier frequency between 80 MHz to 100 MHz, the range of local oscillator frequencies are:

(80 – 14.4) MHz to (100 + 14.4) MHz

i.e., 65.6 MHz to 114.4 MHz

46. (c)

- The time constant RC determines how quickly the capacitor discharges.
  - For proper operation of the envelope detector, RC must allow the capacitor to follow the variations of the modulating signal (envelope), not the carrier waveform.
  - $RC << \dfrac{1}{2\pi f_c}$ ensures that the capacitor does not hold charge for periods comparable to the carrier frequency f_c.
- $RC > \dfrac{1}{2\pi f_m}$

  The time constant RC must be large enough for the capacitor to follow the envelope of the modulating signal without excessive discharge during the intervals between peaks of the modulating signals.
  - $RC > \dfrac{1}{2\pi f_m}$ ensures that the capacitor discharge slowly enough to maintain the envelope shape.
  - If RC is smaller than this limit, distortion will occur as the capacitor discharges too quickly.
- The capacitor must discharge at a rate proportional to the modulating signal frequency f_m, not the carrier frequency f_c.
- A proper value of RC ensures smooth detection of the envelope without distortion.

47. (d)

f_s varies from 550 kHz to 1900 kHz

$$IF = 450 \text{ kHz}$$

Superheterodyne receivers uses "down-conversion" where the incoming radio frequency (f_s) is mixed with a local oscillator (f_l) frequency to create a intermediate frequency (IF) such that

$$f_l = f_s + IF$$

We have,

$$f_{l\max} = f_{s\max} + IF = 1900 + 450 = 2350 \text{ Hz}$$

$$f_{l\min} = f_{s\min} + IF = 550 + 450 = 1000 \text{ Hz}$$

$$\therefore \quad \frac{f_{l\max}}{f_{l\min}} = \frac{2350}{1000} = 2.35$$

48. (c)

Given: Carrier power, $P_C = 1000$ W

Power in each sideband, $P_{USB} = P_{LSB} = \dfrac{P_C\mu^2}{4} = 230$ W

where μ is the modulation index.

$$\therefore \quad P_{LSB} = P_{USB} = \frac{P_C\mu^2}{4}$$

$$230 = \frac{1000\,\mu^2}{4}$$

$$\mu^2 = \frac{230 \times 4}{1000}$$

$$\mu = \frac{2}{10}\sqrt{23} = 0.959$$

$$\%\ \text{modulation} = 95.9\%$$

49. (c)

We have,

$$m(t) = 2\cos[(4\pi \times 10^3 t)]$$

$$c(t) = 5\cos[2\pi \times 10^6 t]$$

Hence, $f_m = 2 \times 10^3$ Hz and $A_m = 2$

$f_c = 1 \times 10^6$ Hz and $A_c = 5$

It is given that $\Delta f = 4 \times (\text{BW})_{AM} = 4 \times 2f_m$

$$\Delta f = 4 \times 2 \times 2 \times 10^3$$

$$= 16 \text{ kHz}$$

$\therefore$ Modulation Index, $\beta = \frac{\Delta f}{f_m} = \frac{16 \text{ kHz}}{2 \text{ kHz}} = 8$

The FM signal can be written as

$$S(t) = A_c \sum_{n=-\infty}^{\infty} J_n(\beta)\cos(2\pi f_c + nf_m)t$$

The value of n for the term $\cos[2\pi\,(1008 \times 10^3 t)]$ can be obtained as

$$f_c + nf_m = 1008 \times 10^3$$

$$(10^6) + (n \times 2 \times 10^3) = 1008 \times 10^3$$

$$n = \frac{8}{2} = 4$$

Hence, the coefficient of the term.

$$\cos[2\pi\,(1008 \times 10^3 t)] = A_c J_n(\beta) = 5J_4(8)$$

50. (c)

Let n be the number of bits per sample

$$n = \lceil \log_2 L \rceil$$

where $\lceil \log_2 L \rceil$ indicates the next higher integer to be taken if $\log_2 L$ is not an integer value. Since f_s samples are taken per second, and each encoded into n bits, thus, nf_s binary pulses must be transmitted per second. Thus, time duration of 1-bit,

$$\tau = \frac{1}{nf_s} = \frac{T_s}{n} = \frac{T_s}{\lceil \log_2 L \rceil}$$

where T_s is the Nyquist interval.

51. (a)

Given, $R_{XX}(\tau) = Ae^{(-\alpha|\tau|)}$

The second moment of the random variable Y is

$$E[Y^2] = E[(X(5) - X(2))^2]$$

$$= E[X^2(5) + X^2(2) - 2X(5)X(2)]$$

$$= E[X^2(5)] + E[X^2(2)] - 2E[X(5)X(2)]$$

$$= R_{XX}(0) + R_{XX}(0) - 2R_{XX}(3)$$
$$E[Y^2] = A + A - 2Ae^{-3\alpha}$$
$$= 2A(1 - e^{-3\alpha})$$

52. (d)

On passing the input process $X(t)$ with PSD $S_{XX}(\omega)$ through an LTI system with impulse response $H(\omega)$, the PSD of output $Y(t)$ is given by

$$Y(t) = X(t + T) - X(t - T) \quad \text{...(i)}$$
$$S_{YY}(\omega) = S_{XX}(\omega)\,|H(\omega)|^2$$

Taking Fourier Transform of equation (i),

$$Y(\omega) = e^{j\omega T} X(\omega) - e^{-j\omega T} X(\omega)$$
$$Y(\omega) = [e^{j\omega T} - e^{-j\omega T}]X(\omega)$$
$$H(\omega) = 2j \sin \omega T$$

Thus,
$$S_{YY}(\omega) = |2j\sin(\omega T)|^2 S_{XX}(\omega)$$
$$S_{YY}(\omega) = 4 \sin^2\omega T\, S_{XX}(\omega)$$

53. (c)

The noise voltage at the input of the RF amplifier is given by

$$V_n = \sqrt{4kTBR}$$

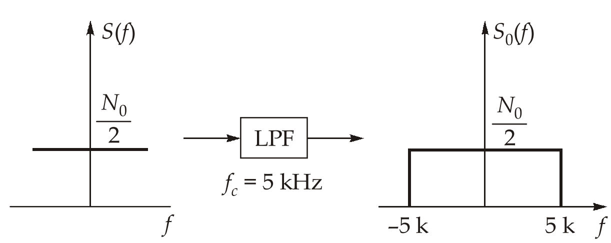
where
$$k = 1.38 \times 10^{-23} \text{ Joules/Kelvin}$$
$$T = 290°\text{K}$$
$$B = 5 \text{ MHz}$$
$$R = 250 + 320 = 570\ \Omega$$

$$V_n = \sqrt{4 \times 1.38 \times 10^{-23} \times 290 \times 5 \times 10^6 \times 570}$$
$$V_n = \sqrt{45.622 \times 10^{-12}} = 6.75 \times 10^{-6} = 6.75\ \mu\text{V}$$

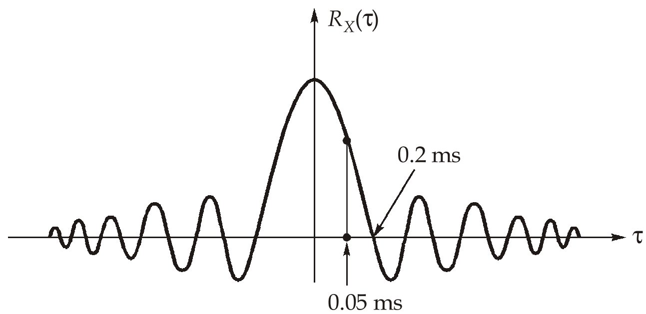
54. (c)

- Flicker noise is more prone at low frequencies, hence it is also known as $\frac{1}{f}$ noise. Transit time noise comes into picture at high frequencies (in micro wave frequency range).
- Shot Noise and Thermal Noise are white noises with a flat spectral density across a wide range of frequencies.

55. (a)



The inverse fourier transform of power spectral density gives the auto-correlation function. The auto-correlation function of output signal is as below,



$R_X(\tau) \neq 0$ at τ = 0.05 msec. So, the samples obtained are correlated.

56. (c)

Frequency detector can be modelled as a differentiator.

$$S_X(f) = \frac{N_0}{2} \rightarrow \boxed{H(f) = j2\pi f} \rightarrow S_Y(f) = S_X(f)|H(f)|^2$$

$$S_Y(f) = \frac{N_0}{2} 4\pi^2 f^2$$

Hences, the output power spectral density is parabolic and cannot be flat, gaussian and raised-cosine.

57. (a)

Thermal noise also known as white noise has a flat power spectral density across frequencies. The thermal noise power is given by

$$P_n = kTB$$

where B is the bandwidth. As B increases, P_n also increases.

Section C : Electronic Devices & Circuits-2 + Analog Circuits Topics-2

58. (c)

The vacuum level refers to the energy of a free stationary electron that is outside of any material.

59. (d)

For the MOS capacitor in inversion,

Width of depletion region in the substrate,

$$W = \left[\frac{2\epsilon_s \phi_s}{qN_a}\right]^{\frac{1}{2}}$$

where,

$$\phi_s = 2\phi_f = 2\left[\frac{1}{2} \times 4 \times 10^{-6} \times 10^4\right]$$

$$= 2\left[\frac{1}{2} \times 0.04\right] = 40 \text{ mV}$$

$$N_a = \frac{2\epsilon_s \phi_s}{q \times W^2} = \frac{2 \times 11.7 \epsilon_0 \times 40 \times 10^{-3}}{1.6 \times 10^{-19} \times 16 \times 10^{-12}} = 3.23 \times 10^{18}/\text{cm}^3$$

60. (a)

Given, $V_{DS} = 0.7$ V
$r_0 = 50$ kΩ

We know that, for MOSFET in saturation, the drain current,

$$I_D = K_n'(V_{GS} - V_T)^2(1 + \lambda V_{DS})$$
$$= I_{Dsat}(1 + \lambda V_{DS}) \quad \text{...(i)}$$

We have, $$\frac{\partial I_D}{\partial V_{DS}} = \frac{1}{r_0}$$

From equation (i), $$\frac{\partial I_D}{\partial V_{DS}} = \frac{1}{r_0} = I_{Dsat} \times \lambda$$

∴ $$\lambda = \frac{1}{r_0 \times I_{Dsat}}$$
$$= \frac{1}{50 \times 10^3 \times 2 \times 10^{-3}}$$
$$= \frac{1}{100} = 0.01 \text{ V}^{-1}$$

61. (b)

We have, Resistance, $$R = \frac{\rho l}{A}$$
$$R \propto \rho \propto \frac{1}{\sigma} \propto \frac{1}{n}$$, where n is the doping concentration

If $n\uparrow$ $R\downarrow$
$n\downarrow$ $R\uparrow$

In a bipolar junction transistor, the base region is lightly doped, the number of free charge carriers are less and hence, the resistance is high and as the emitter is the most highly doped, its resistance is low.

62. (b)

$$V_T = V_{t0} + \gamma\left[\sqrt{V_{SB} + 2\phi_f} - \sqrt{2\phi_f}\right]$$
$$= 0.8 + 0.4\left[\sqrt{3 + 0.7} - \sqrt{0.7}\right]$$
$$V_T = 1.23 \text{ V}$$

63. (a)

Given, $I_{DSS} = 12$ mA
$V_p = -7$ V
$V_{GS} = -3.5$ V

∴ $$I_D = I_{DSS}\left[1 - \frac{V_{GS}}{V_P}\right]^2 = 12 \text{ mA}\left[1 - \frac{3.5}{7}\right]^2 = 3 \text{ mA}$$

64. (b)

The Early effect is the phenomenon where an increase in reverse voltage across the base-collector junction of a BJT causes a reduction in the effective base width. This reduction in the effective base width causes less recombination of carriers in the base region which results in an increase in collector current. Thus, statements 2, 3 and 4 are correct.

65. (c)

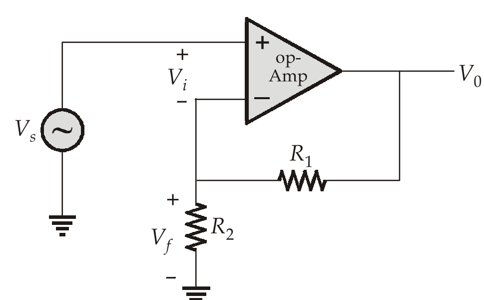
- The threshold voltage is the minimum gate-to-source voltage required to create a conductive channel between the source and drain terminals, enabling current flow. If V_T is less, channel in MOSFET is formed quickly for conduction.
- The Threshold voltage of MOSFET is given by

$$V_T = V_{FB} + 2\phi_F + \frac{t_{ox}\sqrt{4\varepsilon_s q N_a \phi_F}}{\varepsilon_{ox}}$$

  Hence, V_T can be reduced by reducing the oxide layer thickness (t_{ox}) or by reducing the substrate doping concentration (N_a).
- V_T can be reduced by suitable ion implantation. Like in NMOS, positive donor ions are implanted to reduce repulsion to the electrons present in the channel.

66. (c)

We have,



As, $$V_f = \frac{V_0 \times R_2}{R_1 + R_2}$$

$$\frac{V_f}{V_0} = \text{feedback factor, } \beta = \frac{R_2}{R_1 + R_2} = \frac{0.25 \times 10^3}{(1.5 + 0.25) \times 10^3} = \frac{1}{7}$$

Now, the gain of the amplifier with negative feedback is

$$A_f = \frac{A}{1 + \beta A}$$

As, $\beta A >>> 1$, then

$$A_f = \frac{1}{\beta} \approx 7$$

67. (d)

We have,

Amplifier gain without feedback,

$$A_{OL} = 100$$

$$\text{Desensitivity, } D = 1 + \beta A_{OL} = 2$$

- $$\text{Harmonic distortion, } D_{Hf}\% = \frac{D_H}{1+\beta A}$$

$$2\% = \frac{D_H}{2}$$

$$D_H = 4\%$$

- Amplifier gain with feedback,

$$A_{CL} = \frac{A_{OL}}{1+\beta A} = \frac{100}{2} = 50$$

- $$\begin{aligned}\text{Amount of feedback} &= -20 \log_{10}(1 + \beta A) \\ &= -20 \log_{10} 2 \\ &= -6.02 \text{ dB}\end{aligned}$$

68. (d)

Procedure to find type of feedback:

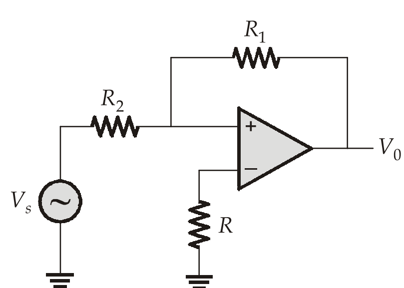
Step-I: Identify feedback element or feedback network in the given circuit.

Step-II: If feedback element is directly connected to output node, then it indicates voltage sampling otherwise current sampling.

Step-III: If feedback element is directly connected to input node, then it indicates shunt mixing otherwise series mixing.

Now,

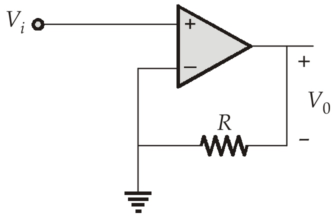
P.



- Feedback element-R_1.
- R_1 is directly connected to output, hence voltage sampling.
- R_1 is directly connected to input, hence shunt mixing.

Therefore, P-voltage shunt feedback.

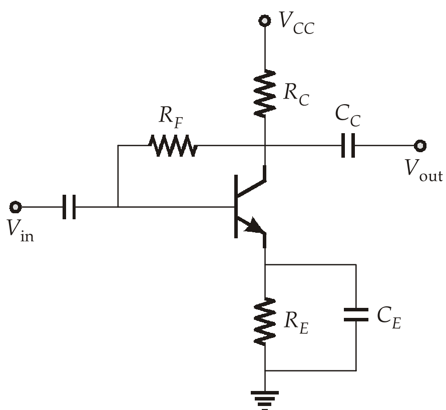
Q.



- Feedback element-R.
- R is directly connected to output, hence voltage sampling.
- R is not directly connected to input, hence series mixing.

Therefore, Q-voltage series feedback.

R.



- Feedback element-R_F.
- R_F is directly connected to output ⇒ Voltage sampling.
- R_F is directly connected to input ⇒ Shunt mixing.

Therefore, R-voltage shunt feedback.

69. (d)

We know that,

Feedback Topology	Input Impedance	Output Impedance
Voltage shunt	Decreases	Decreases
Voltage series	Increases	Decreases
Current shunt	Decreases	Increases
Current series	Increases	Increases

70. (a)

- When an op-amp input is in the volts range, its bandwidth is limited by the slew rate, which determines how quickly the output can change, impacting high-frequency signal fidelity.
- When the input signal is in the millivolt range, the op-amp operates in the small-signal region. The bandwidth of the op-amp is defined by the unity gain bandwidth (also known as the gain-bandwidth product), which is a constant.

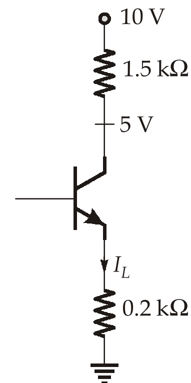
Thus, both the statements are correct.

71. (a)

72. (d)

As op-amp is ideal, it implies $V_- = V_+ = 5$ V due to the virtual short.

Now,



$$I_C \approx I_L = \frac{10-5}{1.5} = \frac{50}{15} = \frac{10}{3} \text{ mA}$$

73. (b)

74. (b)

Both the statements are individually correct but statement (II) is not the correct explanation of statement (I).

75. (a)

As quiescent point of a class A power amplifier is at the centre of the load line, the power dissipation of transistor is maximum since the transistor is biased to conduct for both positive and negative input cycles. Thus, efficiency of class A amplifier is low.

❍❍❍❍