Topic 24: Diffusion Models

1. Introduction to Diffusion Models: https://www.assemblyai.com/blog/diffusion-models-for-machine-learning-introduction
2. Lilian Weng's deep but readable overview of diffusion principles and advances: https://lilianweng.github.io/posts/2021-07-11-diffusion-models/
3. Demystifying Diffusion Models: narrative explanation and code from scratch: https://goyalpramod.github.io/blogs/demysitifying_diffusion_models/
4. Stable Diffusion Got Supercharged: Two Minute Papers summary of ControlNet advancements: https://www.youtube.com/watch?v=1RvZWHtFXuY
5. Hugging Face tutorial with full annotated implementation: https://huggingface.co/blog/annotated-diffusion
6. Latent Diffusion Models: the research behind Stable Diffusion using latent space denoising: https://arxiv.org/abs/2112.10752
7. Google AI blog on Imagen and Parti architectures: https://blog.google/technology/research/how-ai-creates-photorealistic-images-from-text/
8. DreamBooth: Fine-Tuning Text-to-Image Diffusion Models (official Google project for personalizing models): https://dreambooth.github.io/
9. Tutorial for custom fine-tuning with DreamBooth: https://huggingface.co/blog/dreambooth
10. Beginner-friendly guide with setup and token tips for DreamBooth: https://www.analyticsvidhya.com/blog/2023/09/dreambooth-stable-diffusion-for-custom-images/
11. ControlNet (adding Control to Stable Diffusion's Image Generation): Explains sketch/pose/edge guidance: https://blog.segmind.com/what-is-stable-diffusion-controlnet/
12. T2I-Adapter: What is it and How Does it Work? A lightweight alternative to ControlNet for guided diffusion: https://medium.com/@utkarsh135/t2i-adapter-what-is-it-and-how-does-it-work-23b42ba99ed4
13. Identifying AI-Generated Images with SynthID: DeepMind's watermarking for diffusion-generated content: https://deepmind.google/discover/blog/identifying-ai-generated-images-with-synthid/
14. OpenAI paper showing diffusion outperforming GANs: https://arxiv.org/abs/2105.05233
15. NVIDIA's EDM paper improving diffusion efficiency: https://arxiv.org/abs/2206.00364
16. Stable Diffusion XL: https://arxiv.org/abs/2307.01952
17. LoRA for Stable Diffusion: lightweight personalization technique using low-rank adapters: https://stable-diffusion-art.com/lora/
18. Runway Gen-2: diffusion-powered text-to-video system for short video generation: https://research.runwayml.com/gen2
19. DALL·E 3: OpenAI's latest image generation model with integrated prompt understanding and filters: https://openai.com/dall-e-3
20. Consistency Models: faster one-step generation alternative to diffusion processes: https://arxiv.org/abs/2303.01469

21. Hugging Face Diffusers Library (open-source toolkit for building, training, and using diffusion models): https://github.com/huggingface/diffusers