Image Generation

Bo Kang, Thomas Demeester, Tijl De Bie

Outline

- Introduction
- A Brief History
- Stable Diffusion Walk Through
- Demo: Train Your Own LoRA Model
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Introduction

- Generate or manipulate images with neural network models
- Applications
  - Unconditional generation
  - Text to image
  - Image to image
  - Inpainting
  - Many more…
- Tools
  - Commercial: OpenAI[1], Midjourney[2]
  - Open source: HF diffusers[3], SD Webui[4], ComfyUI[5], Kohya SS[6]

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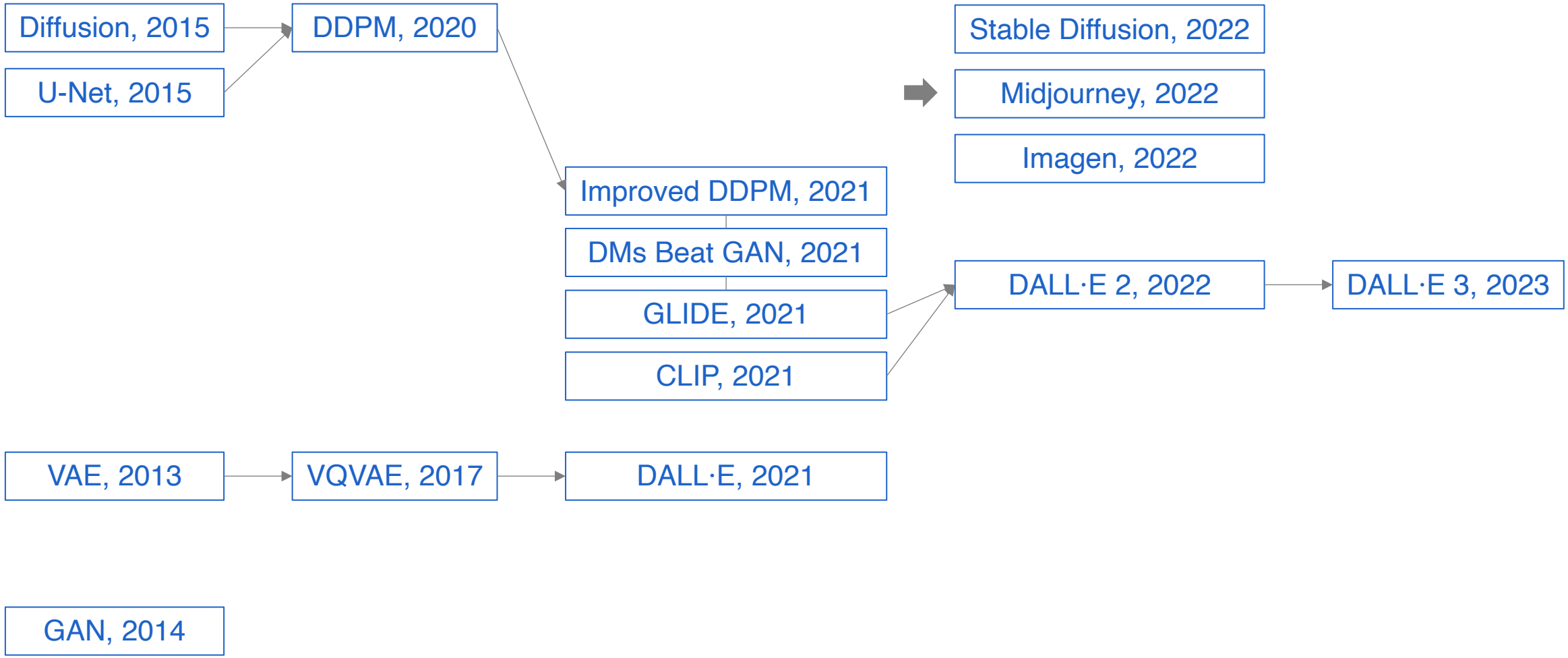
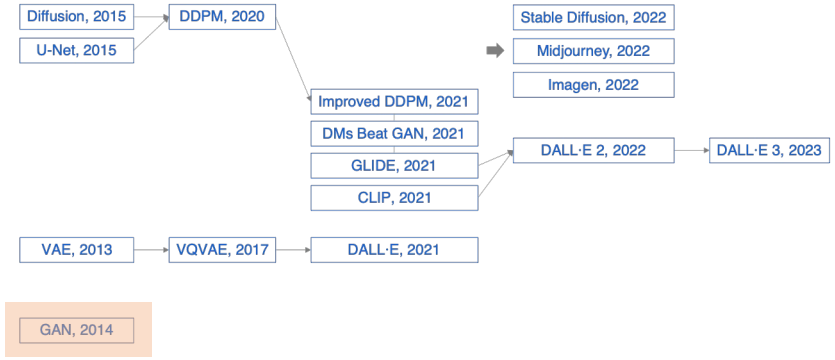
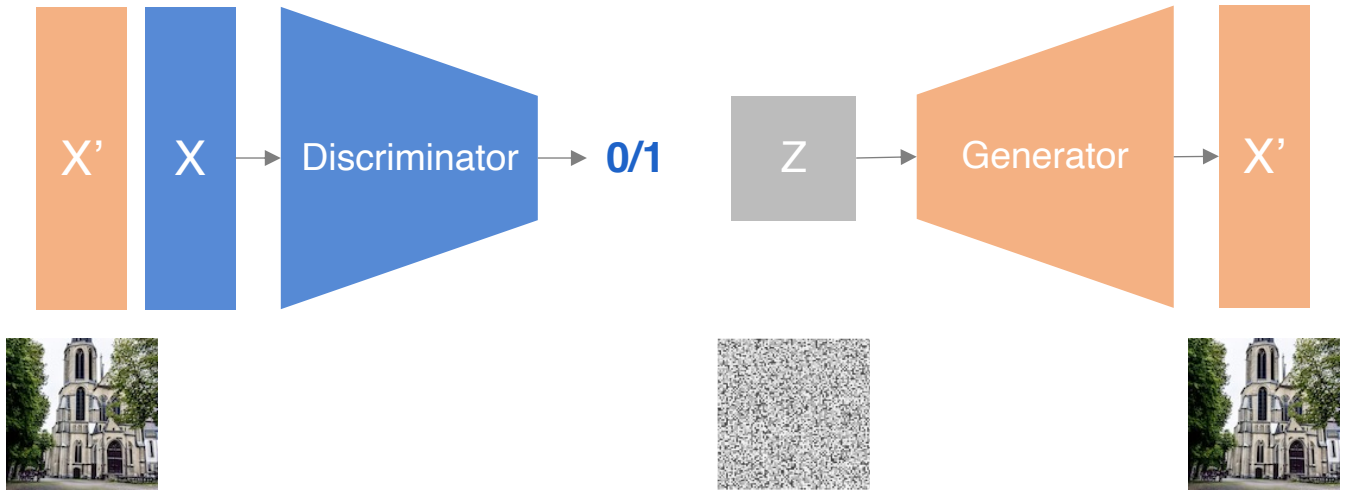
A Brief History



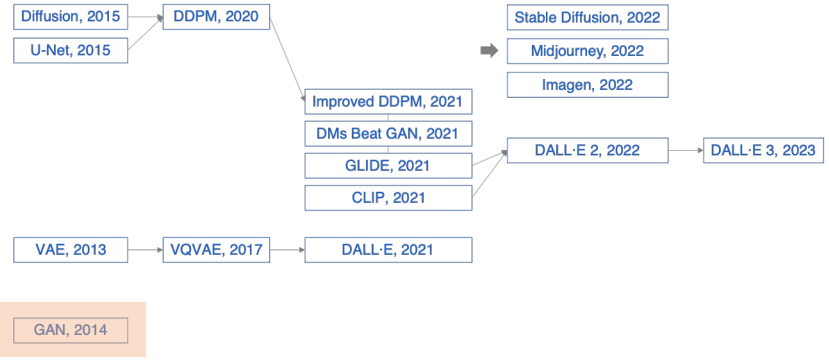
Generative Adversarial Network (GAN)



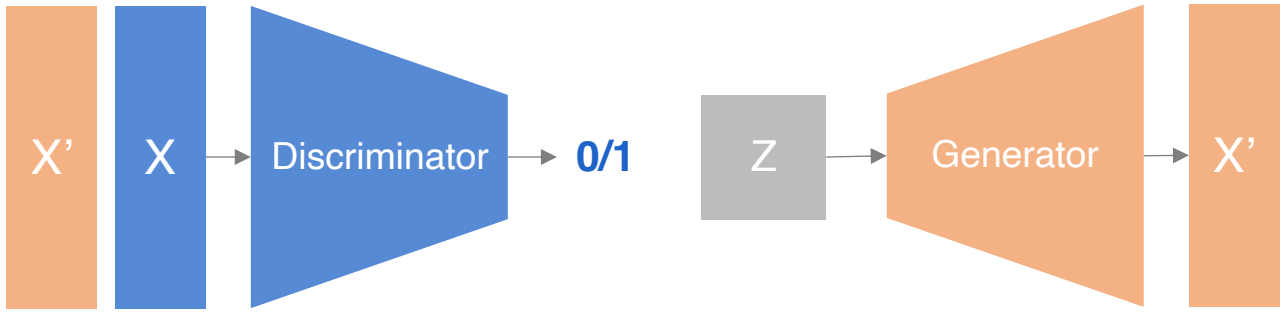
- Idea: train an image generator against a discriminator



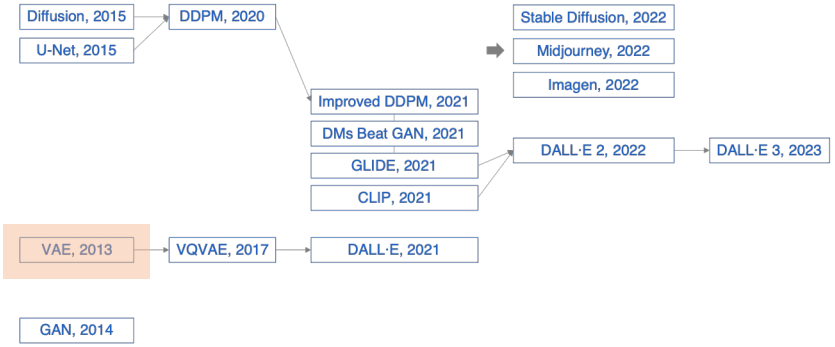
Generative Adversarial Network (GAN)



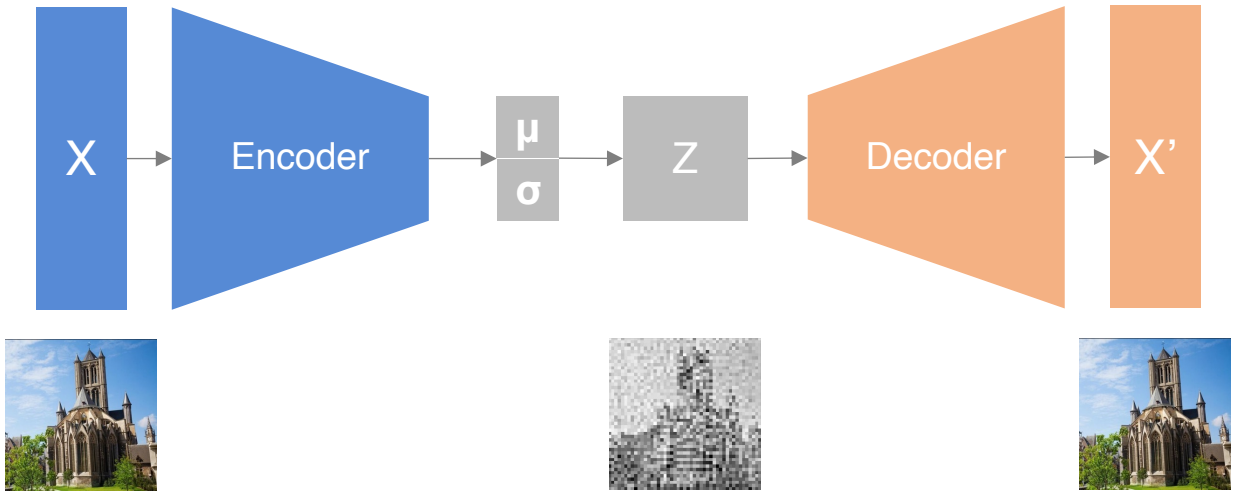
- Idea: train an image generator against a discriminator
- Pros
  - High fidelity
  - Many years of improving, easy to use
- Cons
  - Difficult to train
  - Low diversity
  - Mathematically less elegant



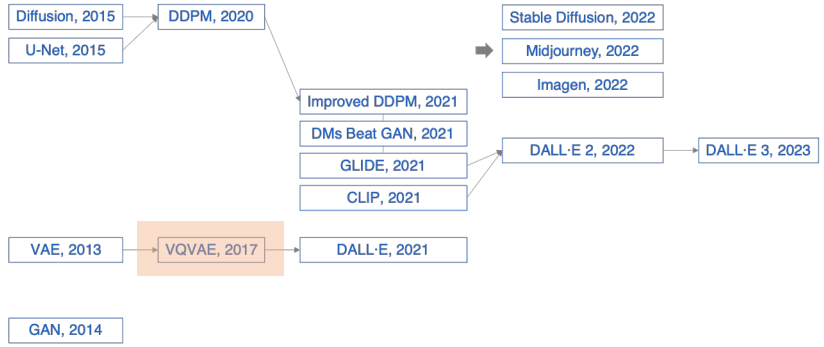
Variational Auto Encoder (VAE)



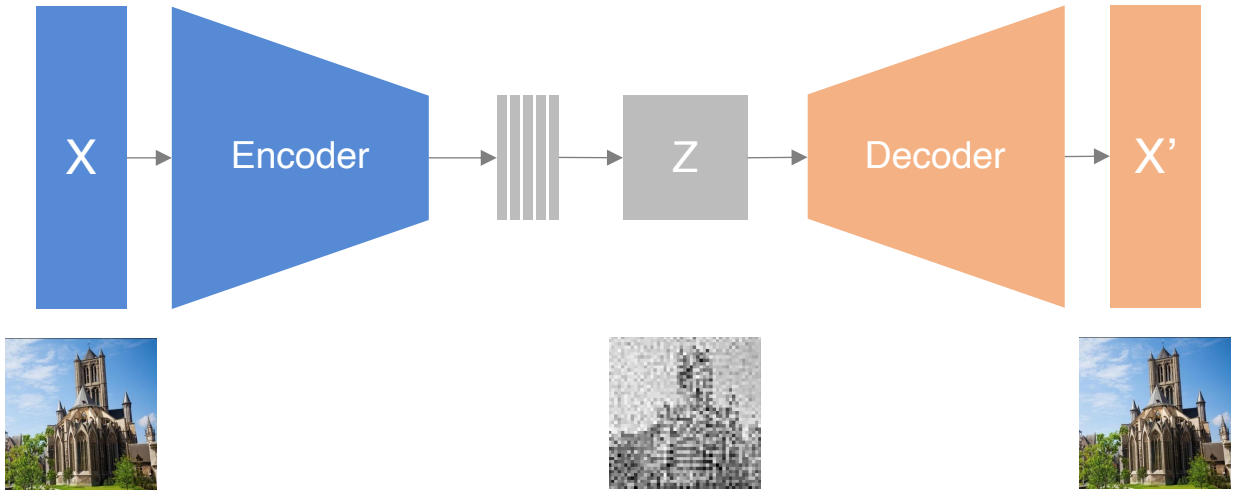
- Idea: learn a distribution over a latent space
- Pros
  - Better diversity
  - Principled probabilistic modeling
- Cons
  - Blurry outputs
  - Still difficult to train



Vector Quantized VAE (VQVAE)



- Idea: learns discrete latent space using vector quantization; learns an autoregressive model for generation
- Pros:
  - Better sample quality
  - More efficient representation
- Cons:
  - Generation needs extra model
  - Training unstableness



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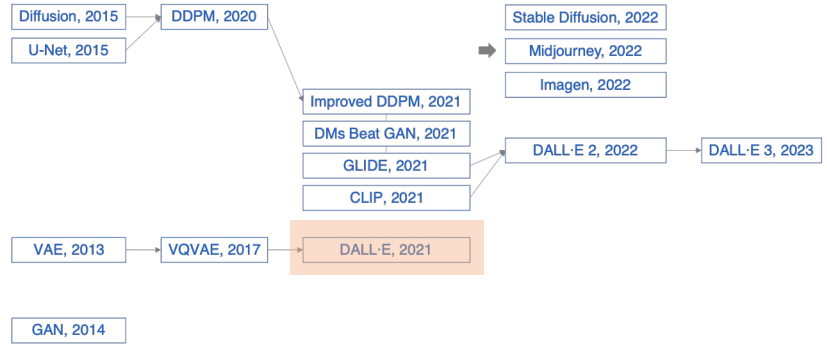
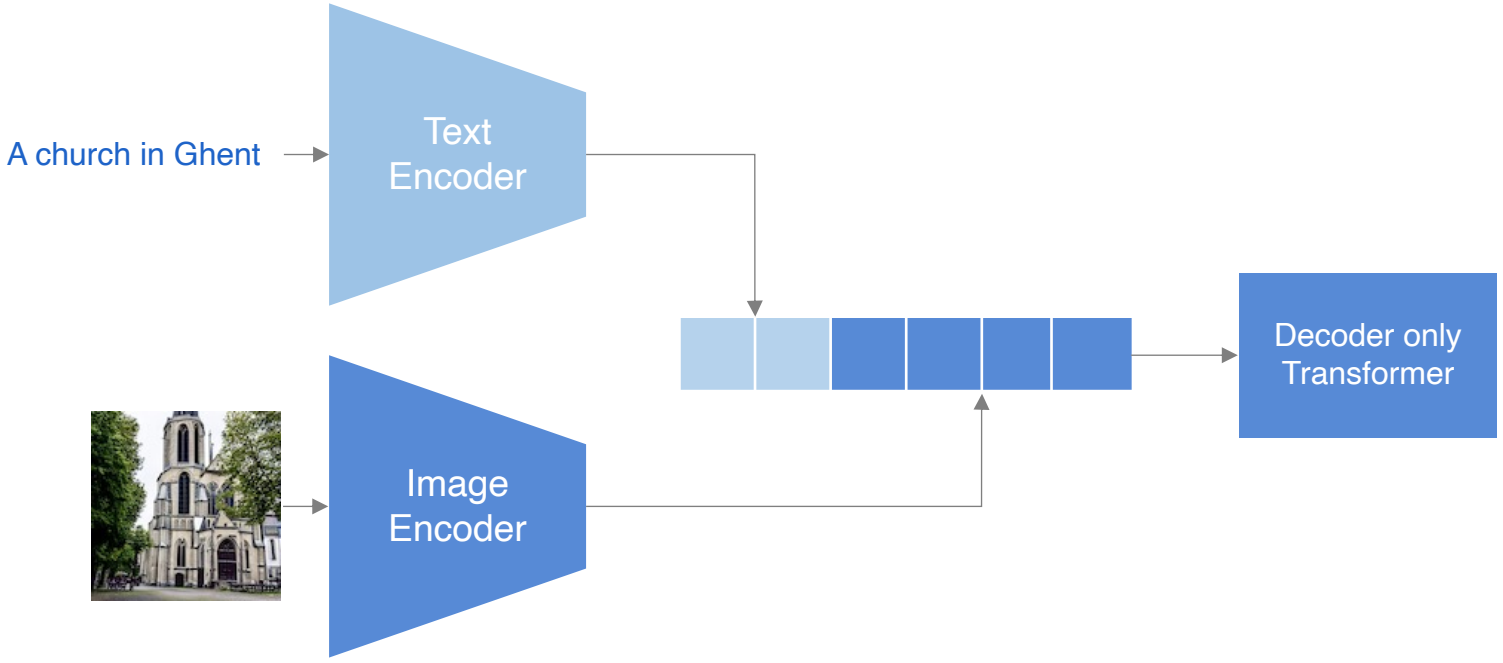
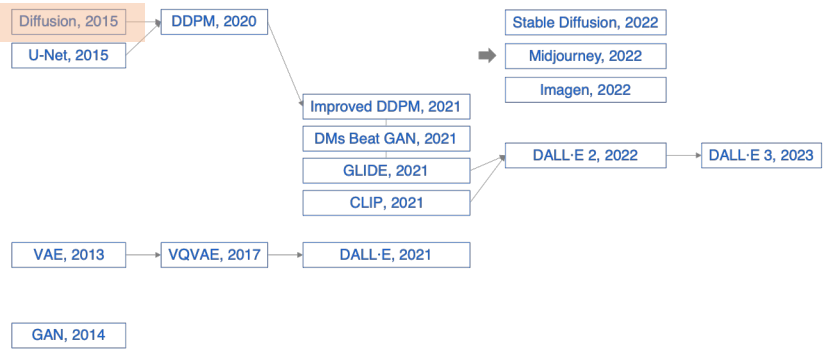


- Idea: VQVAE with text guidance and GPT style autoregressive latent representation generation



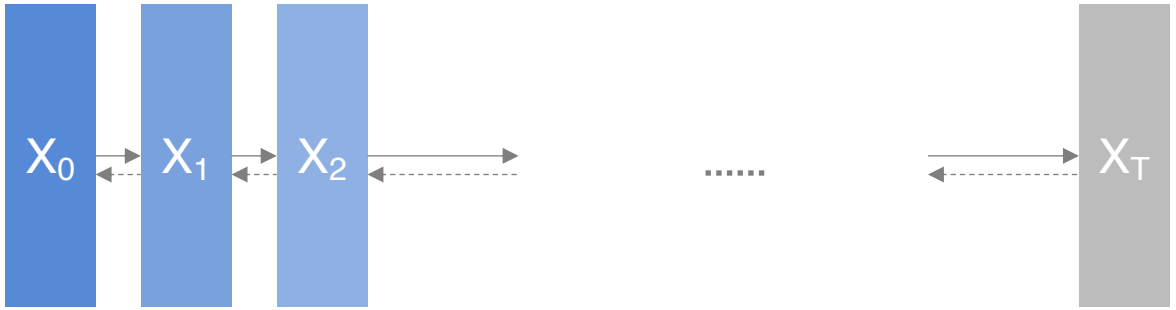
Diffusion model



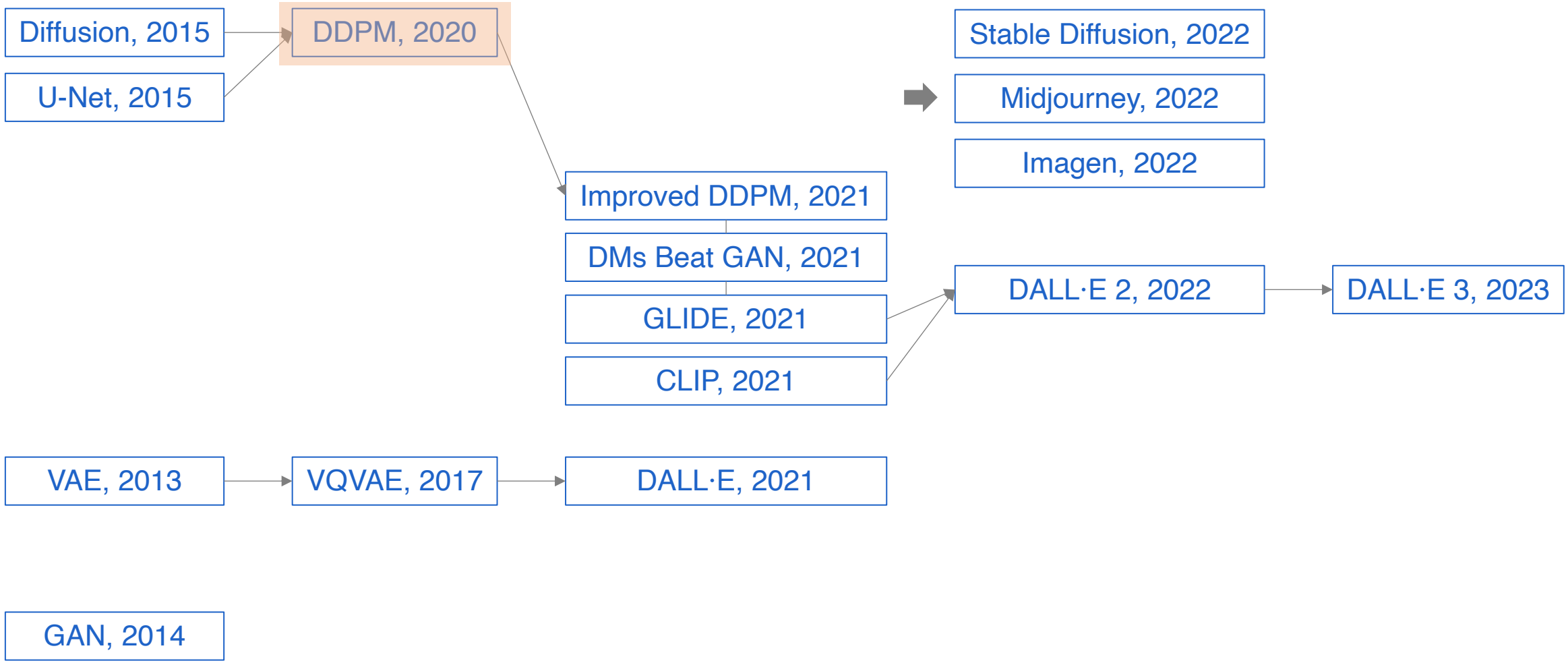
- Idea: gradually transform a distribution of random noise into a complex image data distribution through a reverse diffusion process

- Pros:
  - New paradigm
  - Mathematically principled

- Cons:
  - Generation is still not good enough
  - Slow

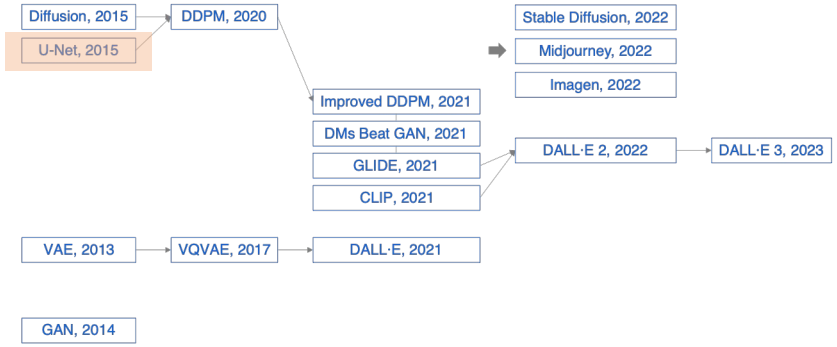


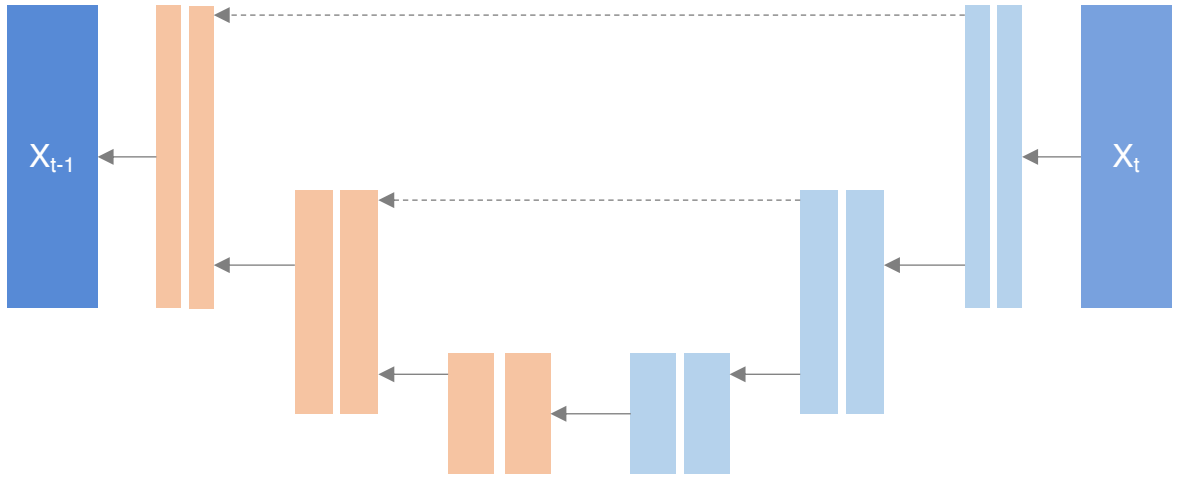
Diffusion model Improvements



U-Net

- Idea:
  - CNN based encoder decoder architecture
  - Originally used to predict segmentation of an image
  - here predicts noise in the reverse diffusion process





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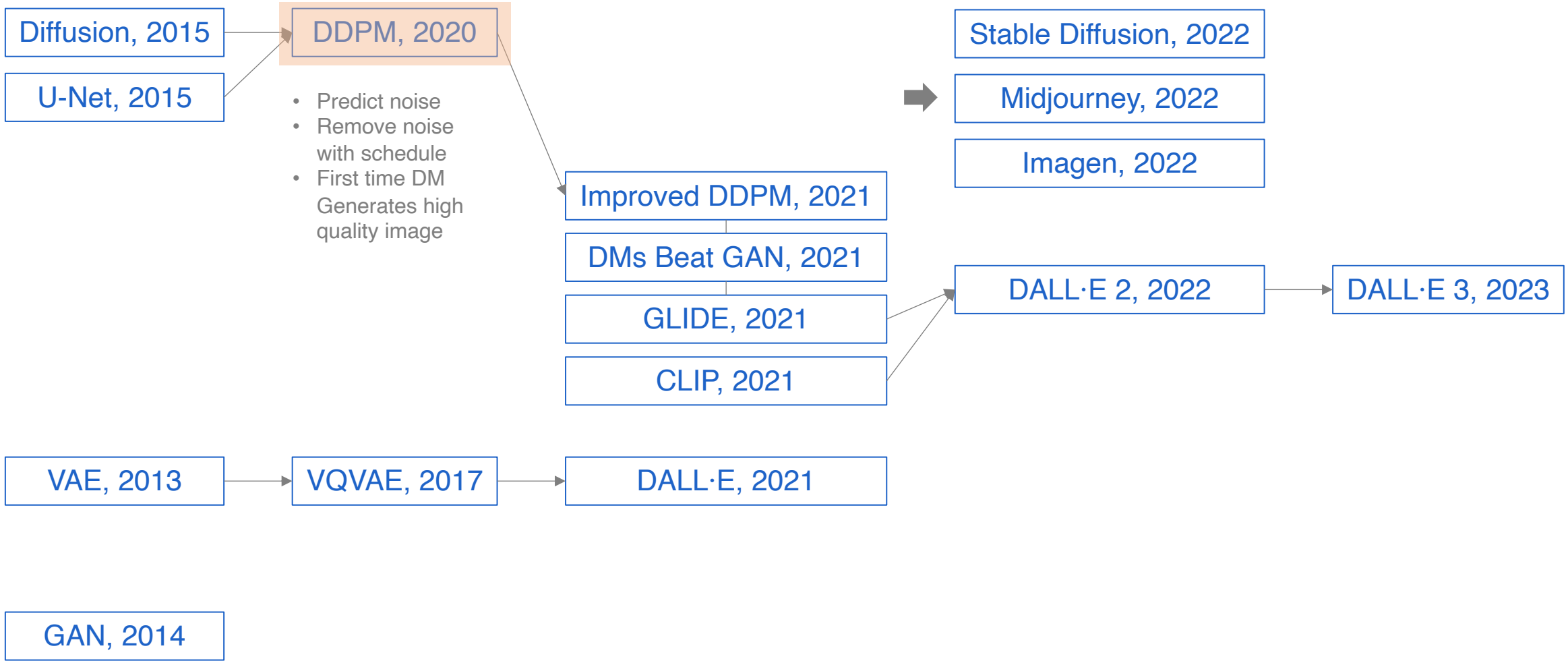

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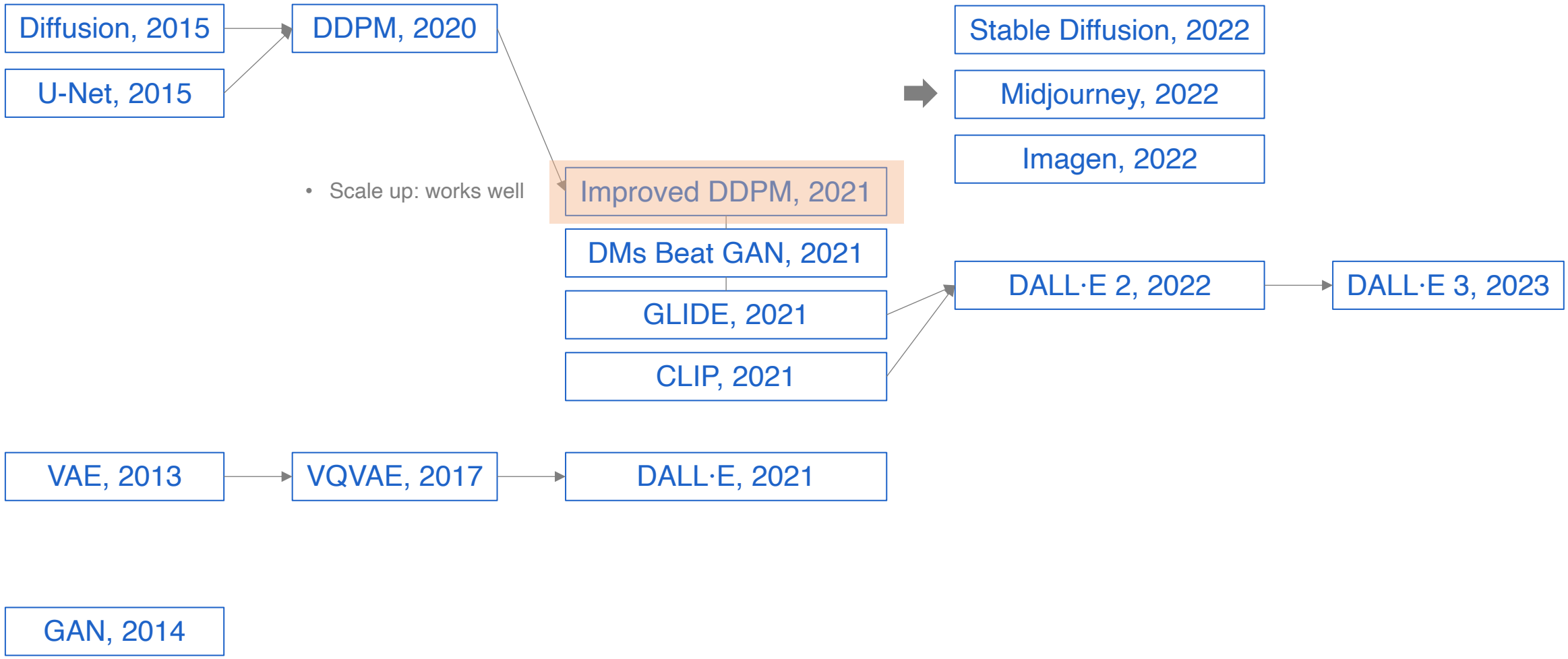

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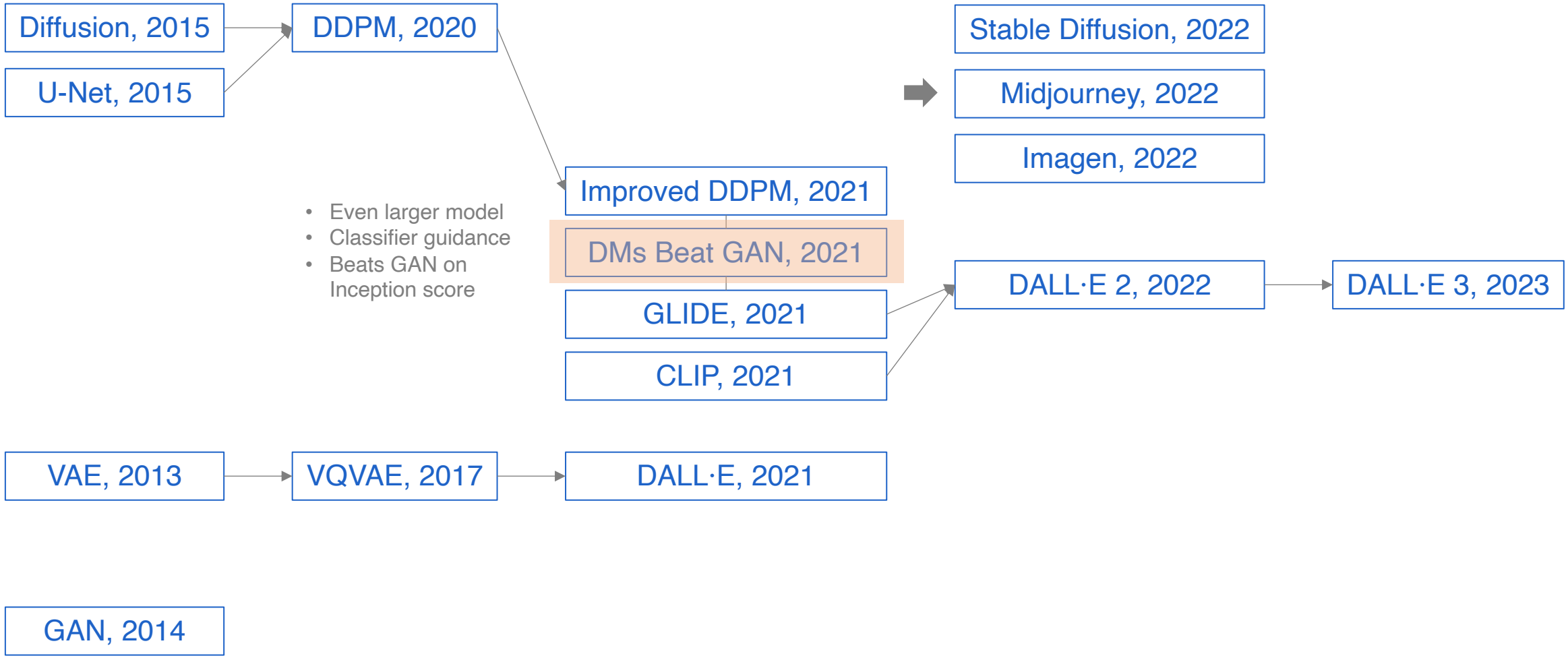

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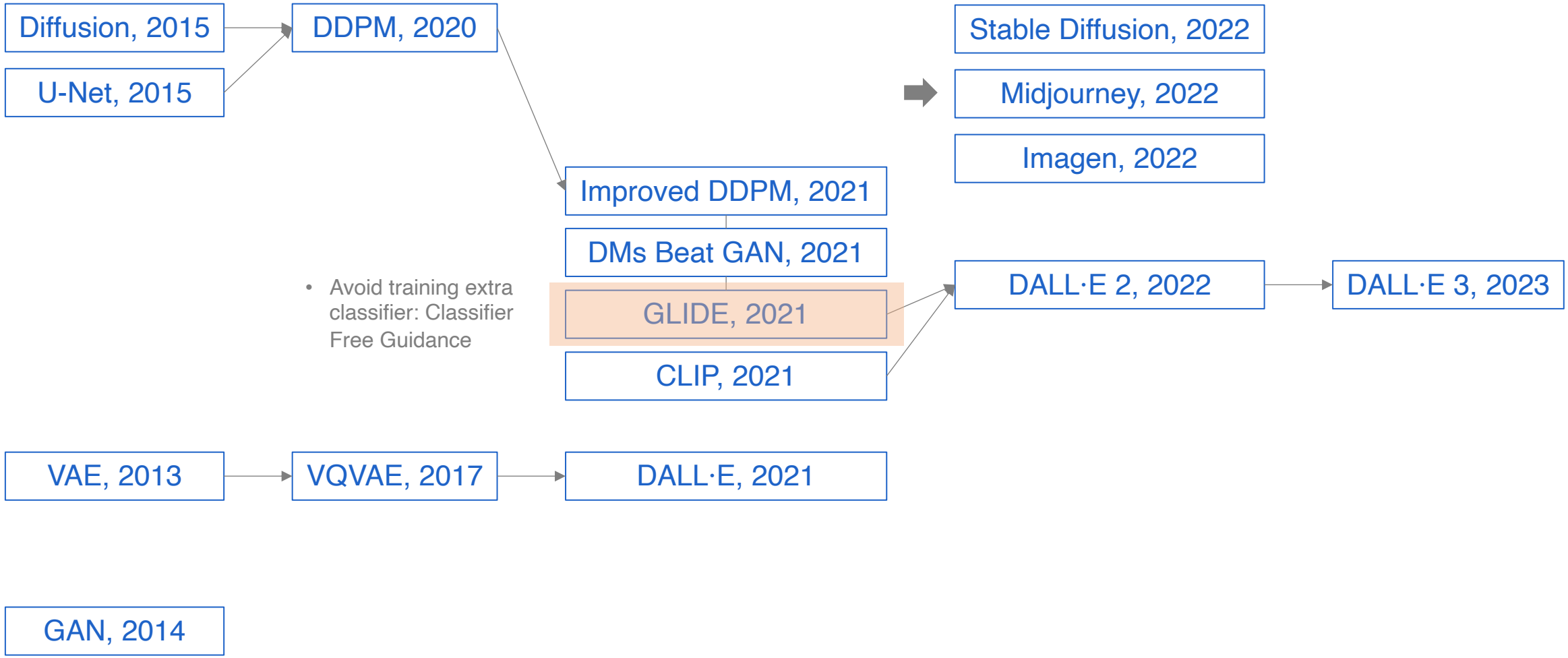

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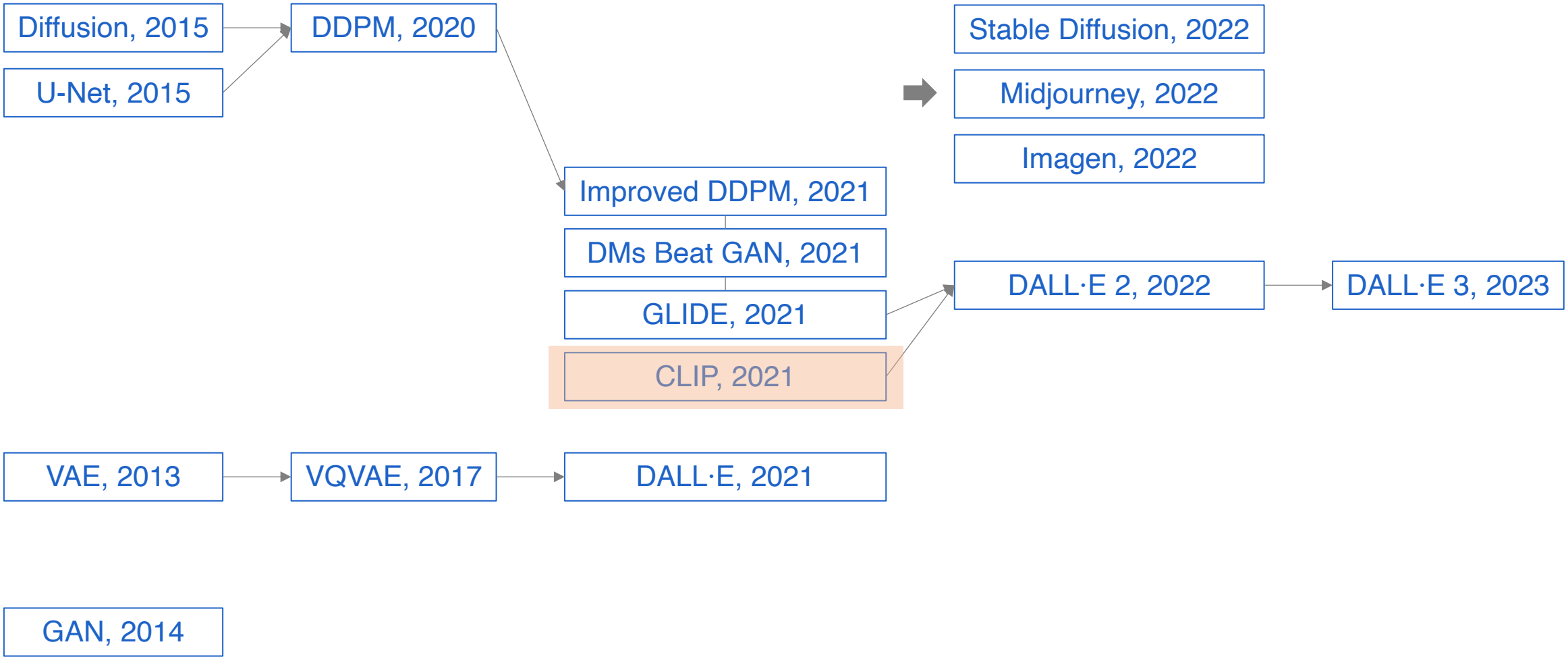
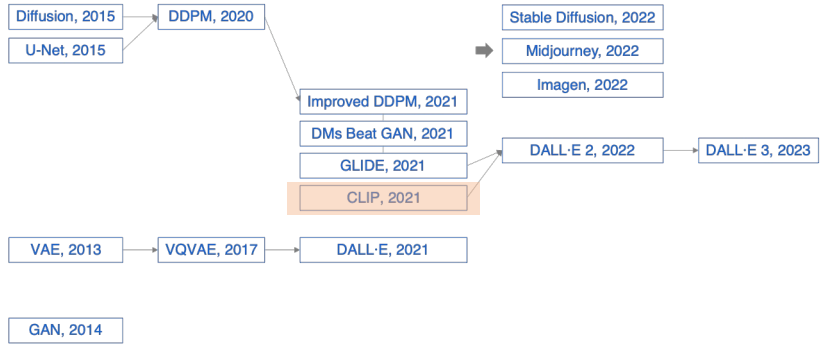
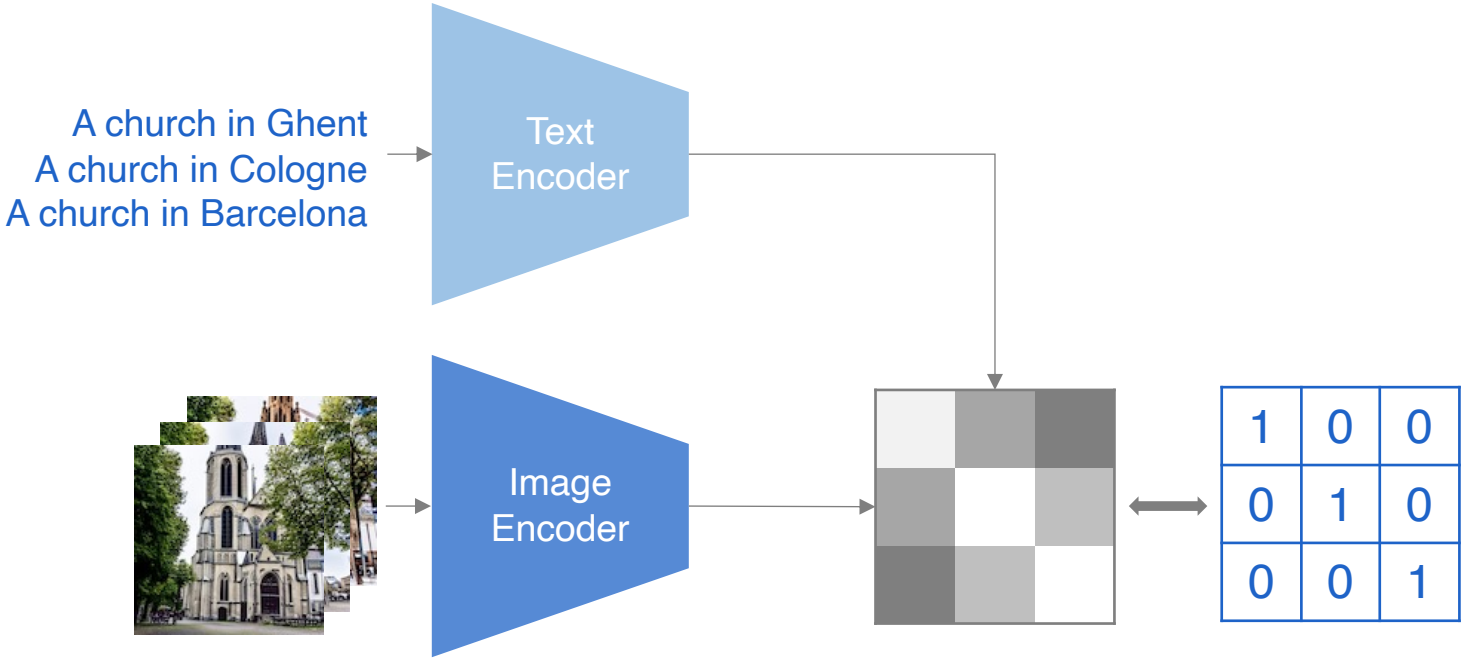


CLIP: Contrastive Language-Image Pre-training



- Idea: learns image embeddings that matches relevant text embeddings



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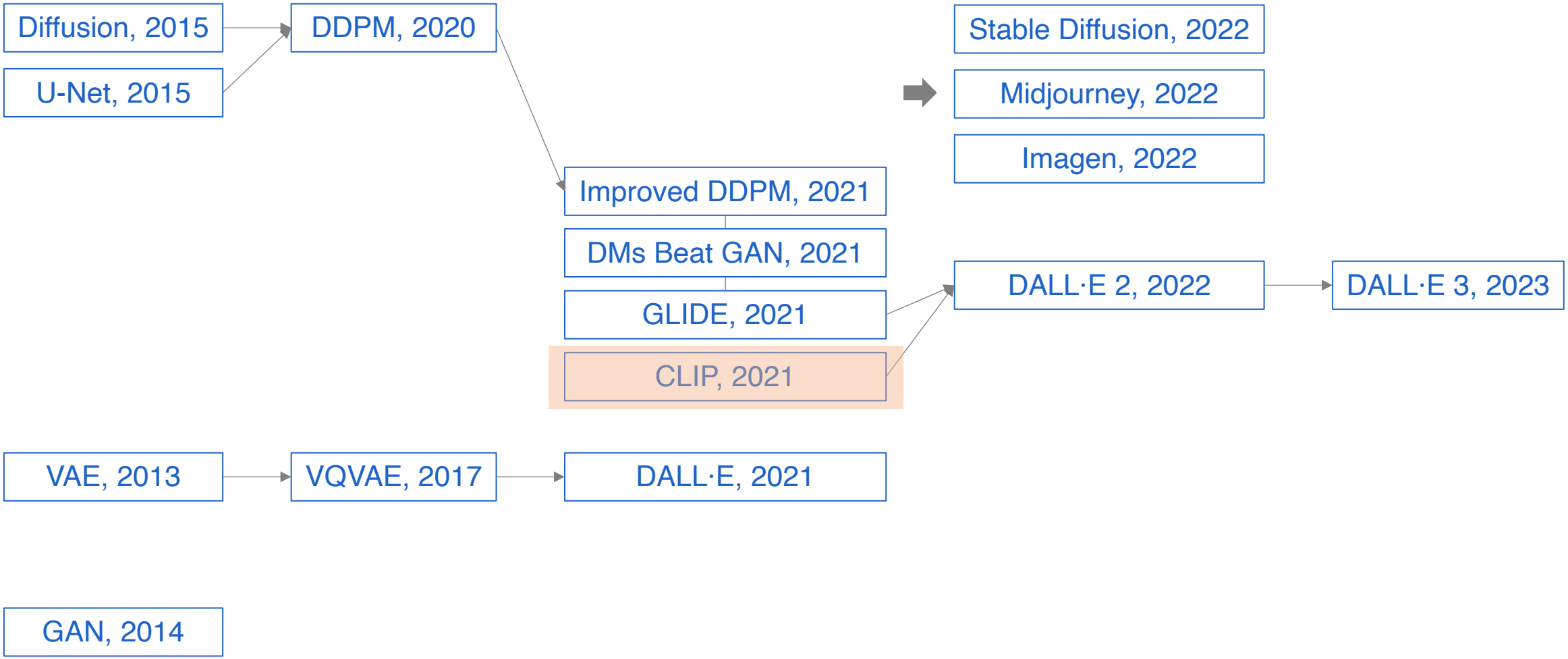

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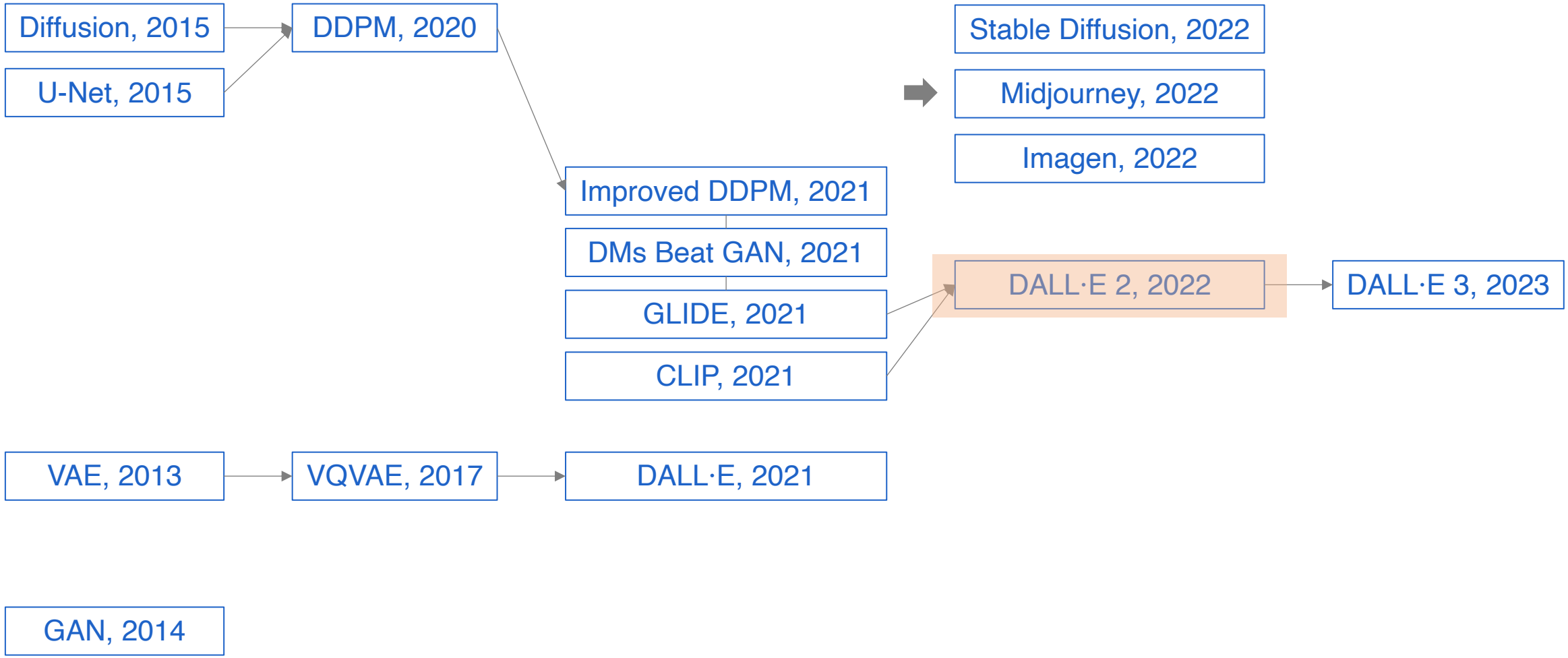

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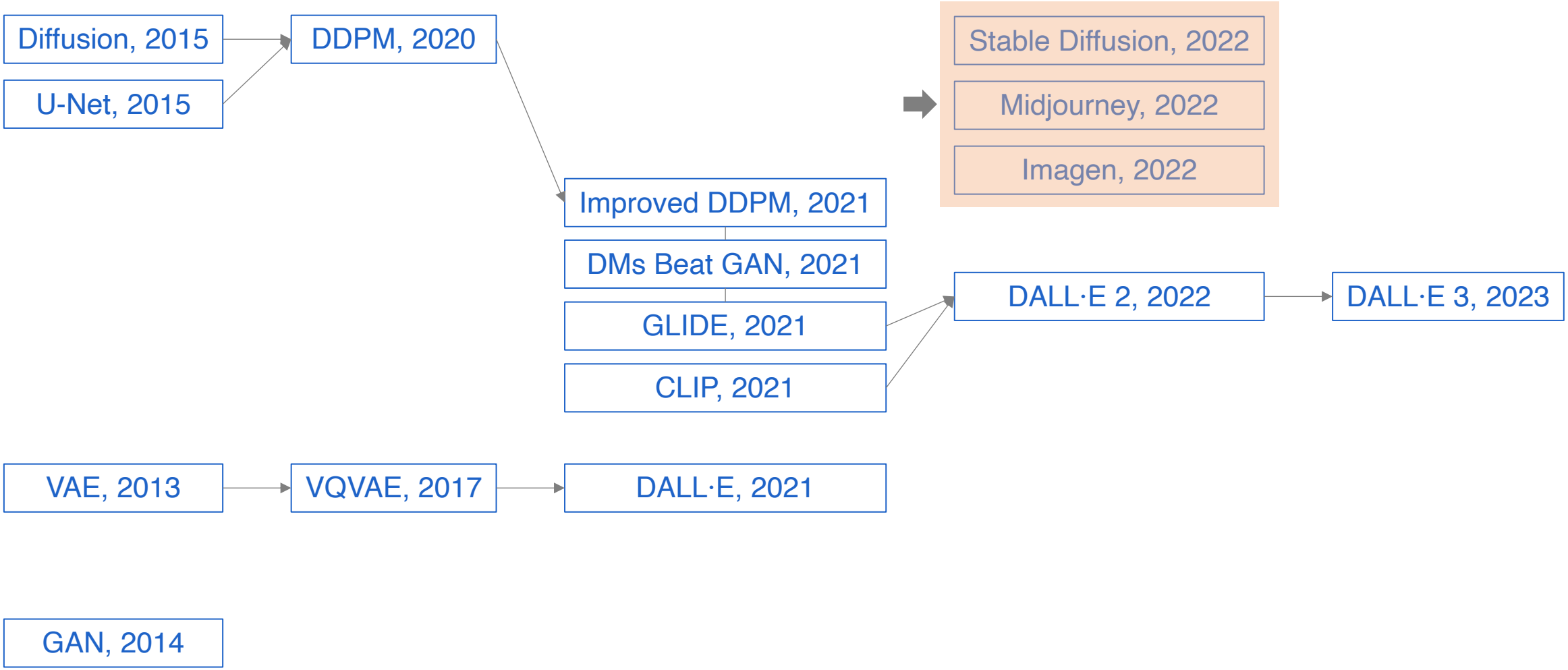
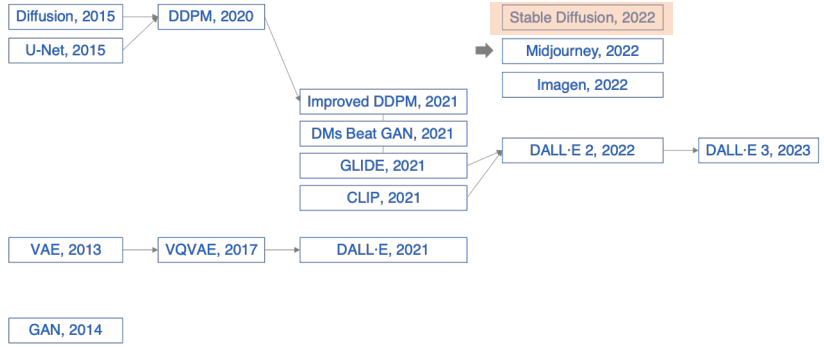
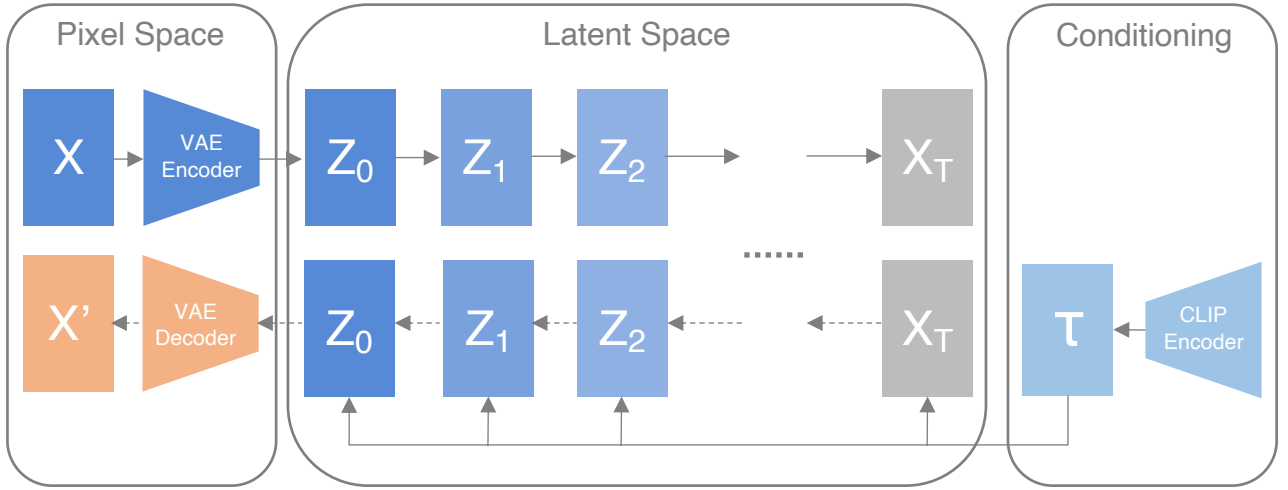


Stable Diffusion

- Idea: apply diffusion process in latent space





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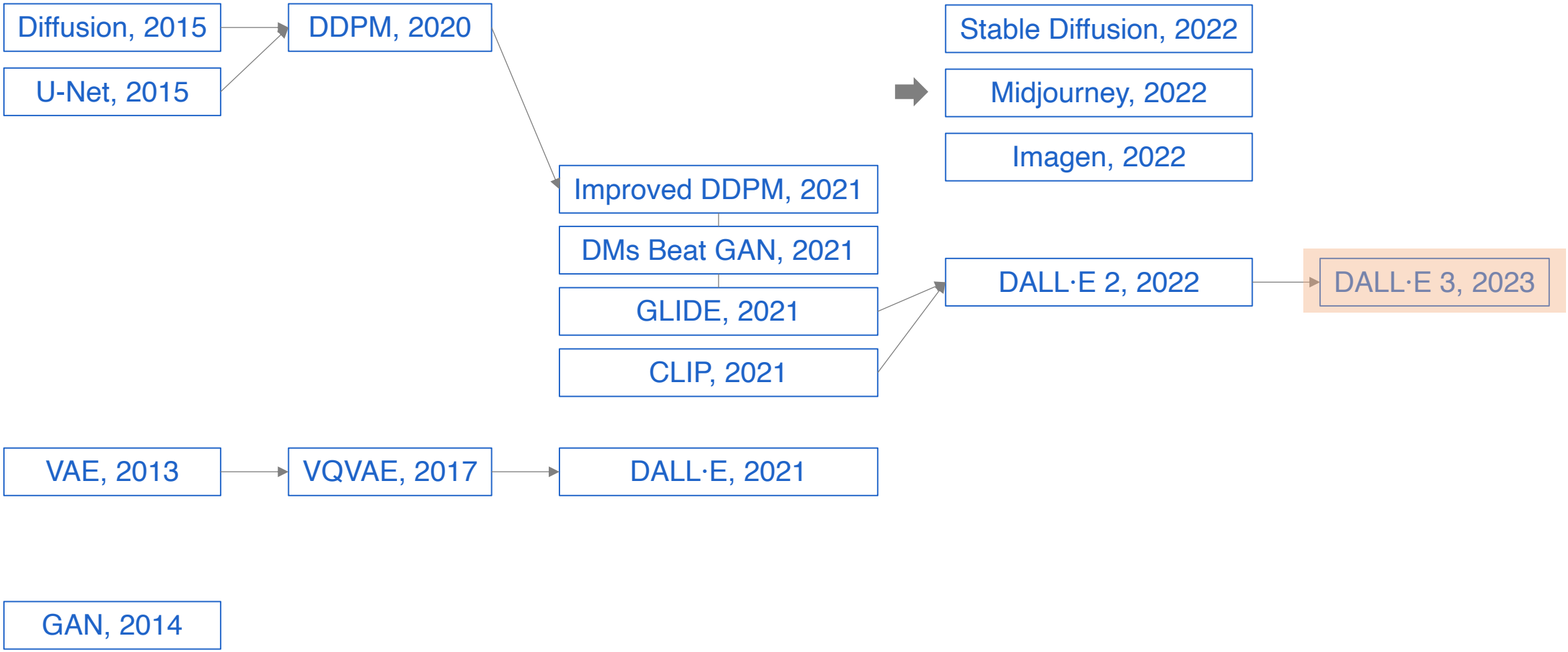

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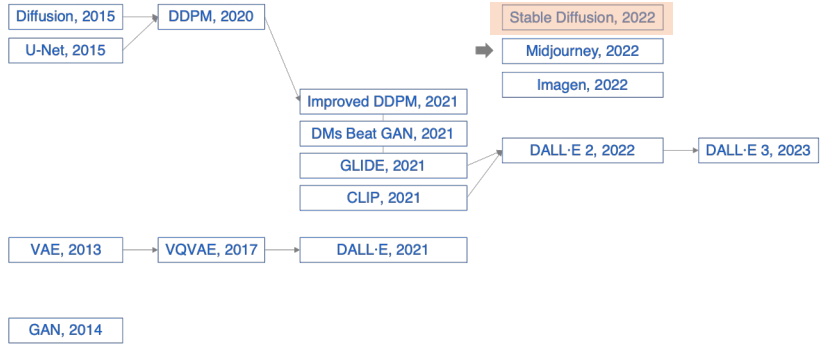
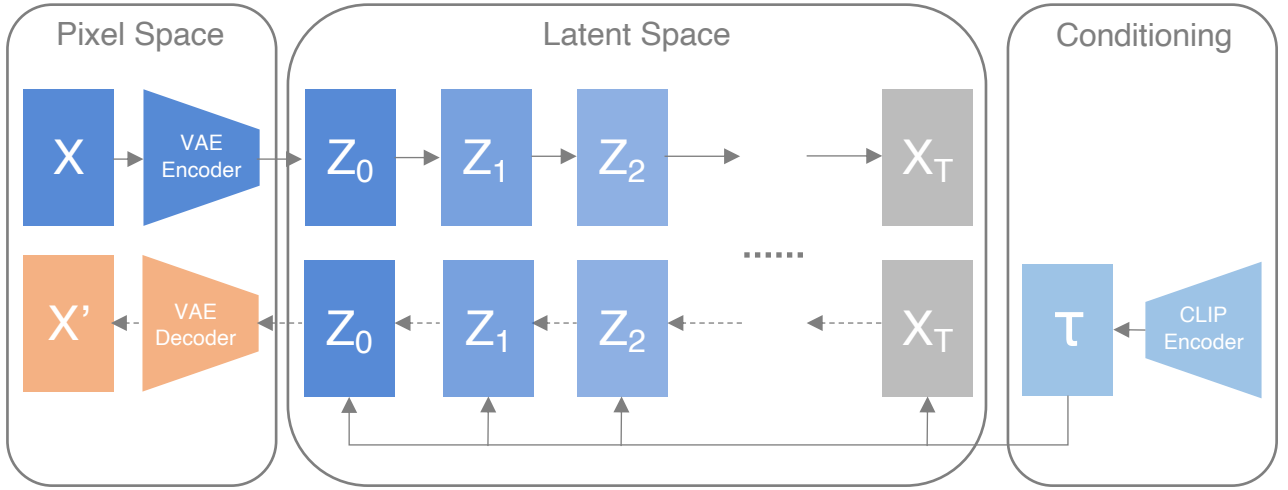


- Idea: apply diffusion process in latent space

- Sub modules
  - Encoder
  - CLIP
  - Scheduler
  - U-Net



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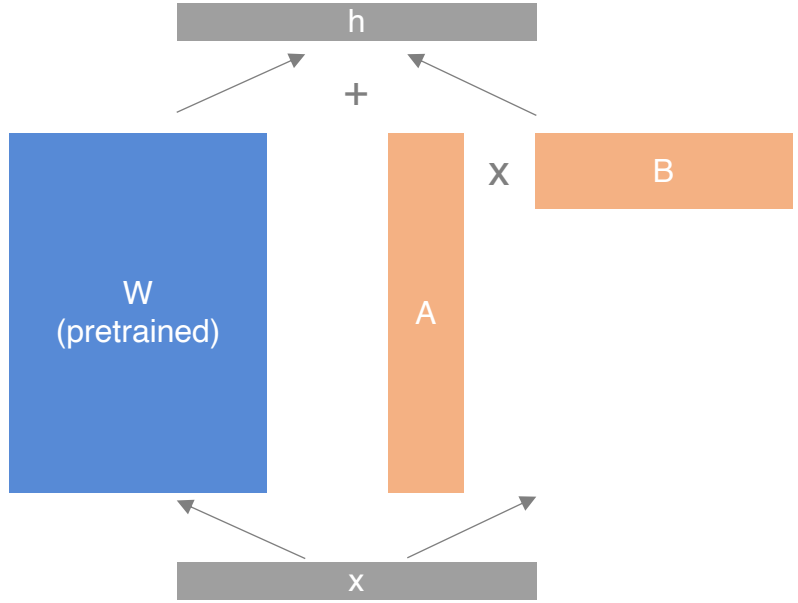
Demo: LoRA Training

- Idea: finetune stable diffusion model by adapt the model weights using extra low rank parameter matrices
- Tools: Huggingface, Kohya_SS
- Training data preparation
- Train
- Usage



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References

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