

# Deep Generative Models



Jan Čech

- Models generating samples from a given distribution
  - The target distribution can be very complex (high-dimensionality)
  - Synthesis of **realistic** samples of, e.g. Text, Code, Audio, Music, **Images**, Video, 3D models
- Unconditioned / Conditioned models
  - Conditioning by a “prompt” (text, image, semantic maps, ...)

- **Lecture outline**

1. Examples
2. Brief taxonomy and Evaluation
3. GANs
4. Diffusion Models

# Examples of Deep Generative Models



- Text

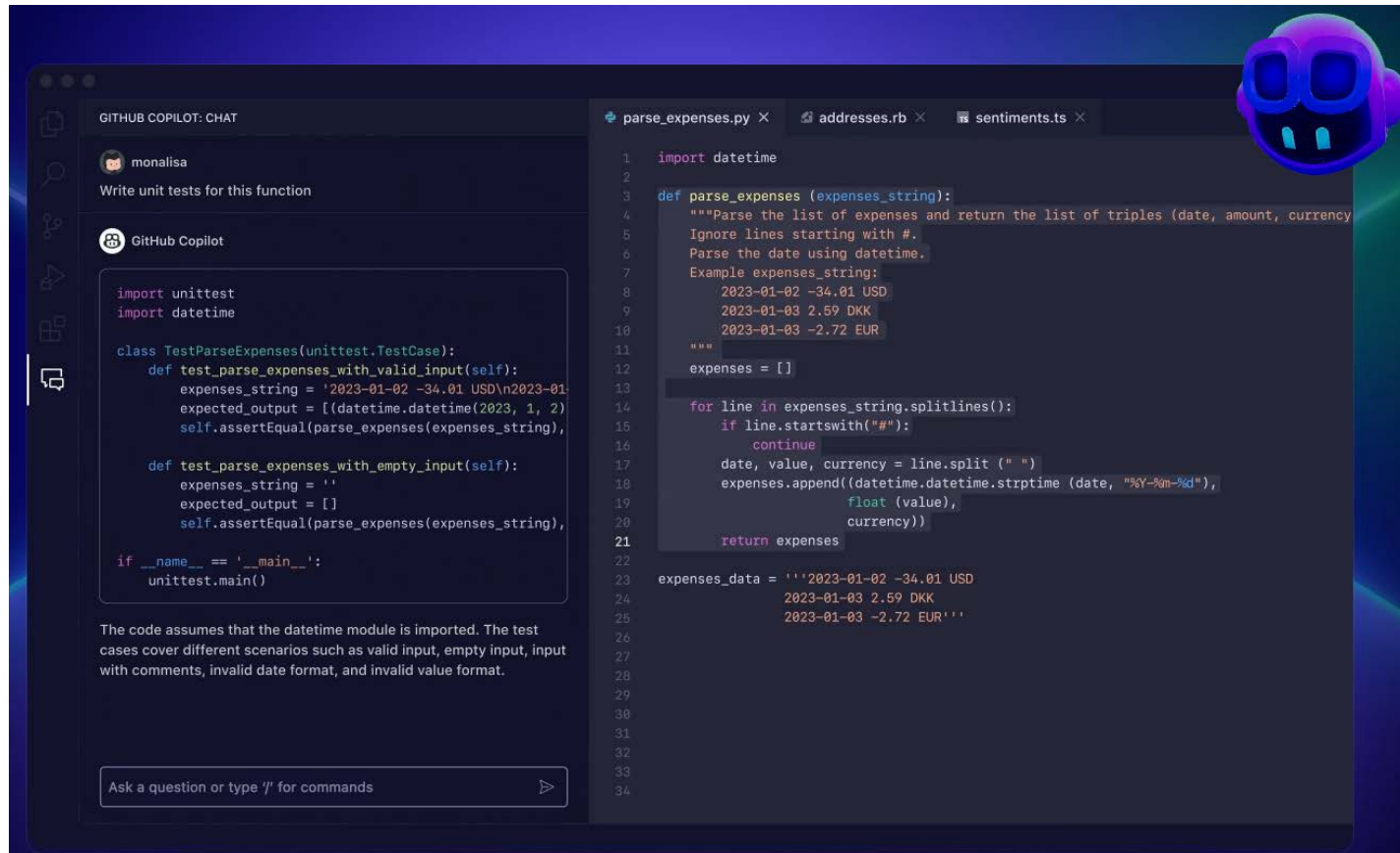
- Large Language models
- GPT 4 (OpenAI), Llama 3 (Meta) – open source



Meta

- Code

- GitHub Copilot – free for students



# Examples of Deep Generative Models

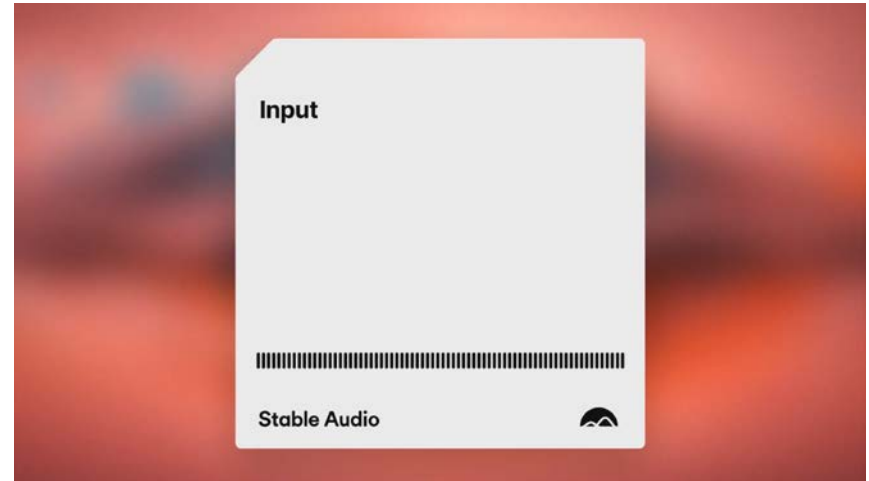


## ■ Audio

– text2audio, audio2audio

- [Stable Audio 2.0](#) (Stability.AI)

**Prompt:** Epic music, big drums, strong melody



- [Suno AI](#) (Cambridge, MA, USA)

**Lyrics:** Computer Vision Method course focuses on the following computer vision problems: finding correspondences between images using image features and their robust invariant descriptors, image retrieval, object detection and recognition, and visual tracking.

**Style:** Large symphonic orchestra and children chorus, epic melody.



– text2speech

- Many models, emotional speech, voice cloning, very realistic...

# Examples of Deep Generative Models

- Unconditioned Image generators
  - Generating photo-realistic samples from image distributions



(Images synthesized by a random sampling)

# Examples of Deep Generative Models

- Text2image
  - DALL-E (OpenAI), Imagen (Google), Midjourney, ...
  - **StableDiffusion** (Stability.AI, free open source)



panda mad scientist mixing sparkling chemicals, artstation



a propaganda poster depicting a cat dressed as french emperor napoleon holding a piece of cheese

# Examples of Deep Generative Models



8

- Video
  - Text2video, image2video, video2video
  - [SORA](#) (OpenAI)

A litter of golden retriever puppies playing in the snow. Their heads pop out of snow



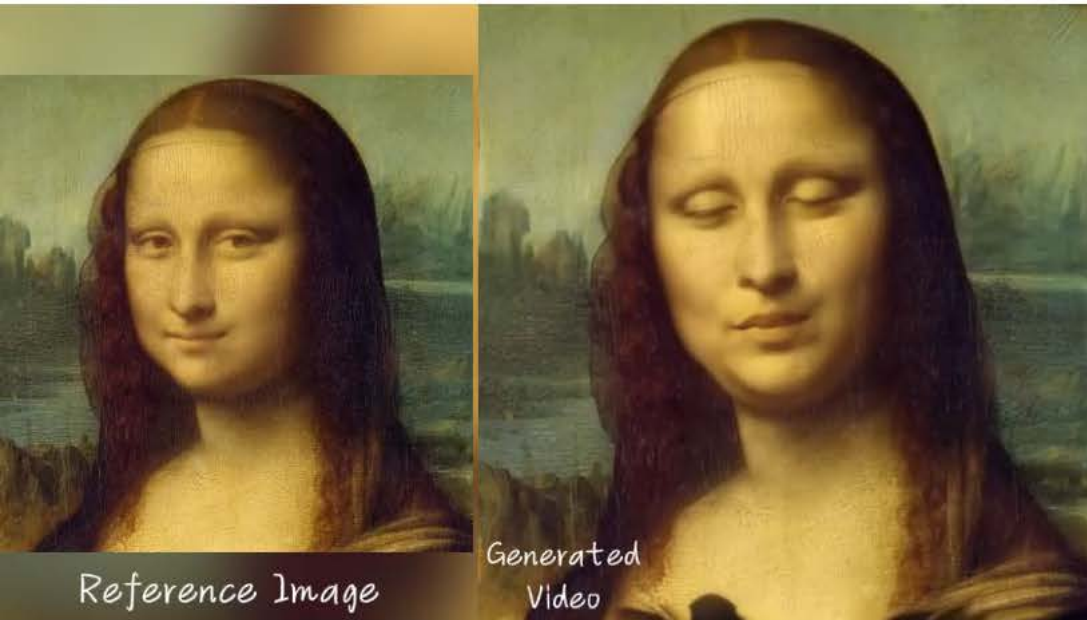
Beautiful, snowy Tokyo city is bustling. The camera moves through the bustling city street, following several people enjoying the beautiful snowy weather and shopping at nearby stalls. Gorgeous sakura petals are flying through the wind along with snowflakes





# Examples of Deep Generative Models

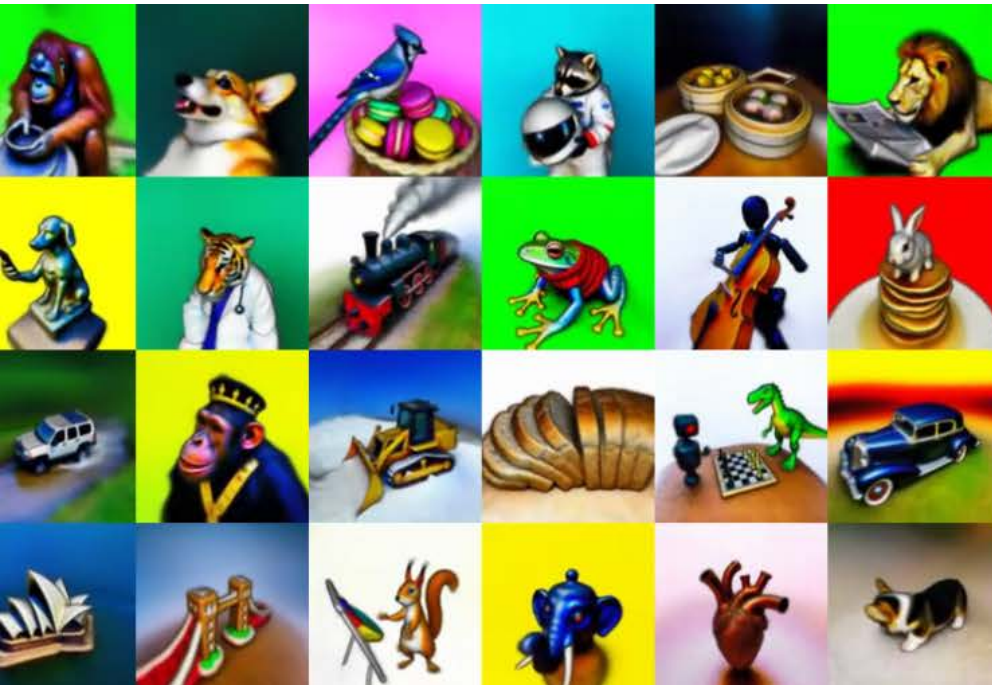
- “Talking/Singing head” (audio2video)
  - EMO: Emote Portrait Alive [[Tian-2024](#)] (Alibaba group)



# Examples of Deep Generative Models

- 3D models (audio2video)
  - Text to 3D [[DreamFusion-2022](#)] (Google), [[Magic3D-2023](#)] (NVIDIA)
  - Image to 3D (TripoSR, Stability.AI)

DreamFusion



TripoSR

Input images



TripoSR outputs



# Brief Taxonomy and Evaluation

# Taxonomy of Deep Generative Models



12

- Several approaches:
  1. Normalizing flow models [[Dinh-2017](#)]
  2. Autoregressive models [[Oord-2016](#)]
  3. Variational Autoencoders [[Kingma-2014](#)]
  4. **Generative Adversarial Networks (GANs)** [[Goodfellow-2014](#)]
  5. **Diffusion models** [[Sohl-Dickstein-2015](#), [Rombach-2022](#)]

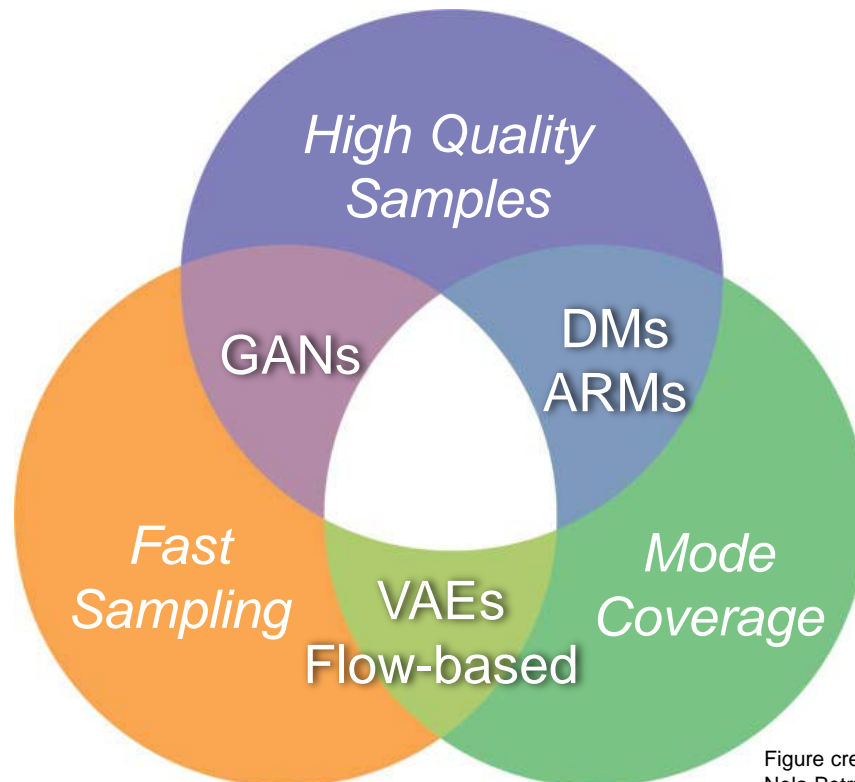


Figure credit  
Nela Petzelkova

# How to Measure Quality of Generative Models?



13

- Synthetic samples should resemble samples of real distribution in the sense of:
  - Fidelity (no obvious artifacts visible)
  - Diversity (enough variability respecting the original distribution)
    - Mode collapse – model always generate the same sample
- All methods use classifier trained on ImageNet (usually Inception v3)

## 1. Inception score (IS) [[Salimans-2016](#)]

- Only synthetic dataset
- Output softmax score  $p(y|x)$

$$\text{IS} = \exp\left(\mathbb{E}_x \text{KL}(p(y|x) || p(y))\right)$$

low entropy      high entropy

low entropy      high entropy

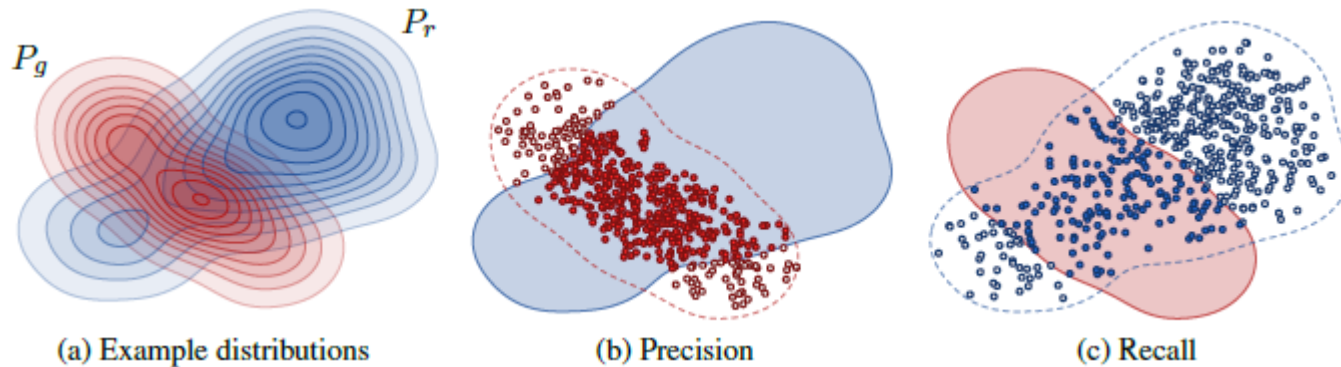
## 2. Fréchet Inception Distance (FID) [[Heusel-2017](#)]

- Two datasets – synthetic, real
- Each sample is “embedded” (features of the penultimate layer)
- Fit Gaussians  $(\mu, \Sigma), (\mu', \Sigma')$

$$\text{FID} = \|\mu - \mu'\|_2^2 + \text{Tr}\left(C + C' - 2(CC')^{1/2}\right)$$

# How to Measure Quality of Generative Models?

## 3. Precision-Recall for distributions [[Kynkäänniemi-2019](#)]



$$f(\phi, \Phi) = \begin{cases} 1, & \text{if } \|\phi - \phi'\|_2 \leq \|\phi' - \text{NN}_k(\phi', \Phi)\|_2 \text{ for at least one } \phi' \in \Phi \\ 0, & \text{otherwise,} \end{cases}$$

$$\text{precision}(\Phi_r, \Phi_g) = \frac{1}{|\Phi_g|} \sum_{\phi_g \in \Phi_g} f(\phi_g, \Phi_r)$$

$$\text{recall}(\Phi_r, \Phi_g) = \frac{1}{|\Phi_r|} \sum_{\phi_r \in \Phi_r} f(\phi_r, \Phi_g)$$

# Generative Adversarial Networks (GANs)

# Generative Adversarial Networks (GANs)



16

A photograph of Yann LeCun speaking at a conference. He is wearing glasses and a light-colored shirt, gesturing with his hands. The background is a dark stage with blue vertical lighting.

"Generative Adversarial Networks is the **most interesting idea in the last ten years** in machine learning."

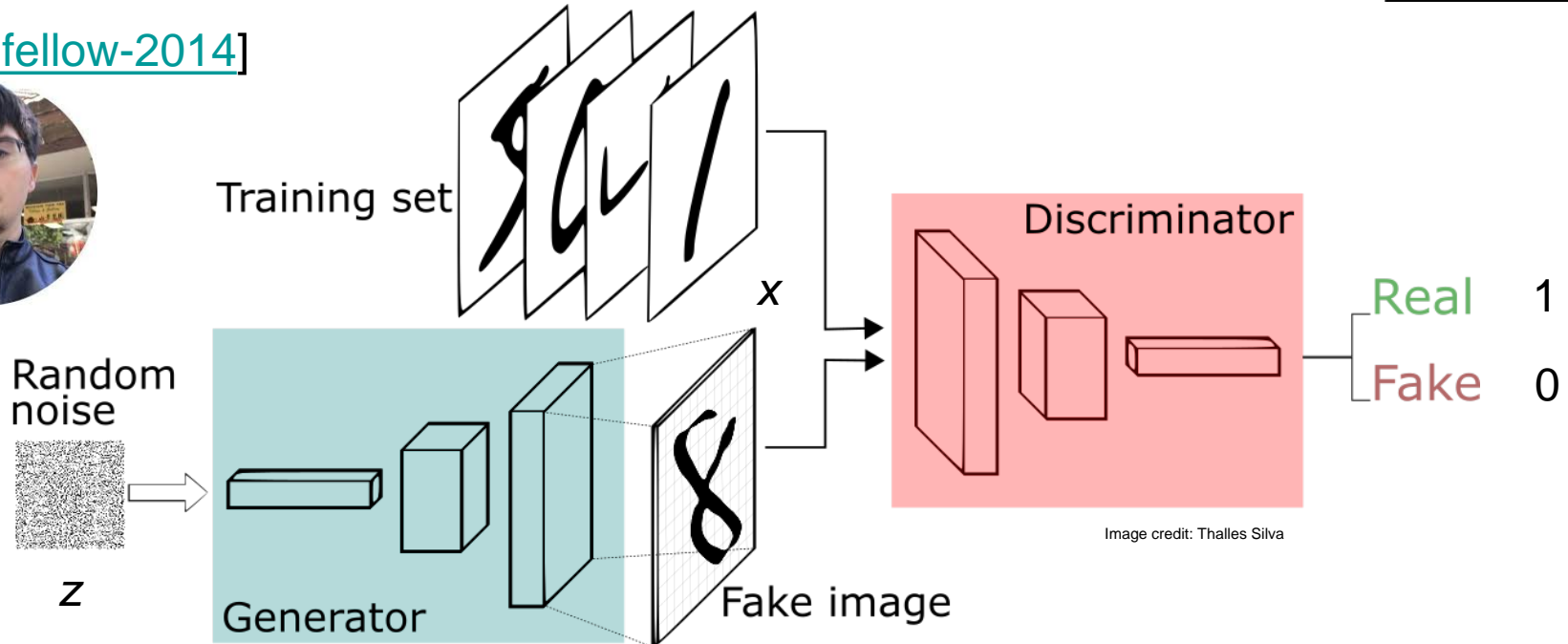
Yann LeCun, Director, Facebook AI



# Generative Adversarial Networks (GANs)



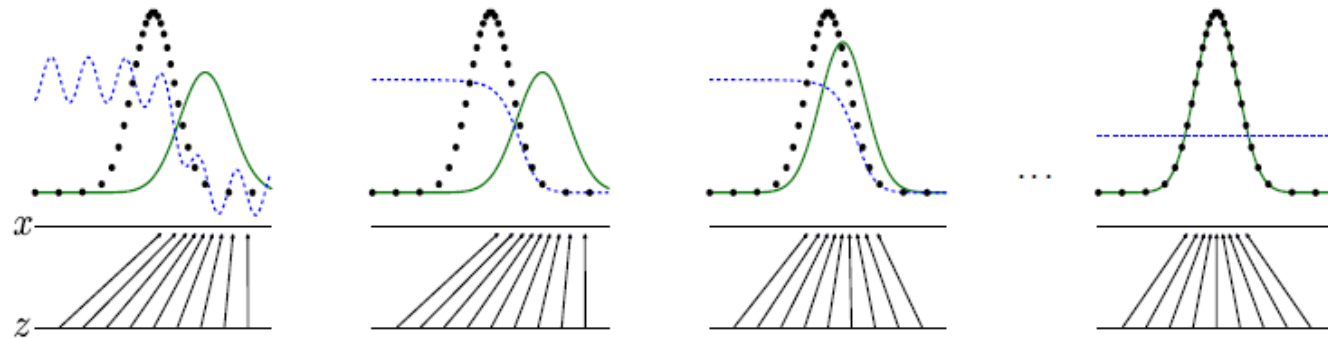
[Goodfellow-2014]



- Two networks: Generator  $G: N(0,1)^k \rightarrow X$ , Discriminator  $D: X \rightarrow [0,1]$
- Min max game between  $G$  and  $D$  when training
  - The discriminator tries to distinguish generated and real samples
  - The generator tries to fool the discriminator

$$\min_G \max_D V(D, G) = \mathbb{E}_{x \sim p_{\text{data}}(x)} [\log D(x)] + \mathbb{E}_{z \sim p_z(z)} [\log(1 - D(G(z)))]$$

# Generative Adversarial Networks (GANs)



- Seems to capture the image manifold
  - Smooth transitions when interpolating in the latent space

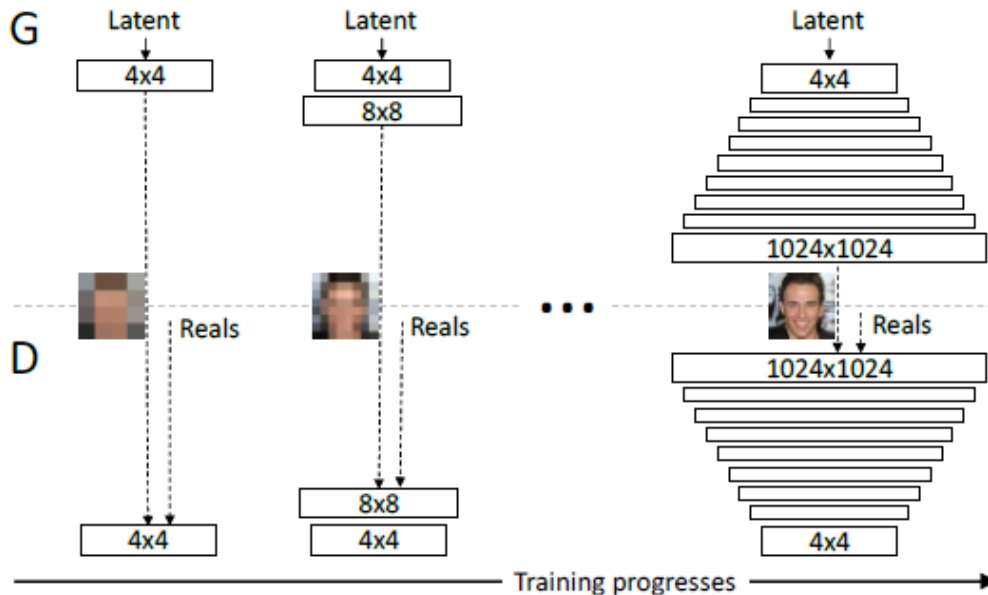
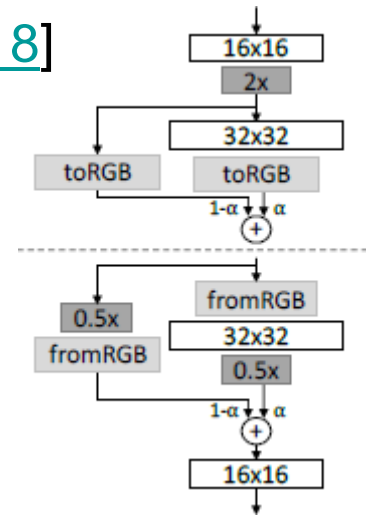


- However:
  - The training is fragile (alternating optimization), mode collapse
  - Did not work well for high-resolution (until recently)

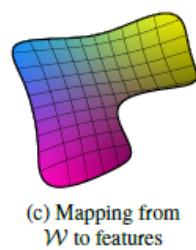
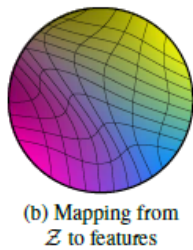
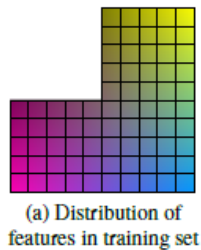
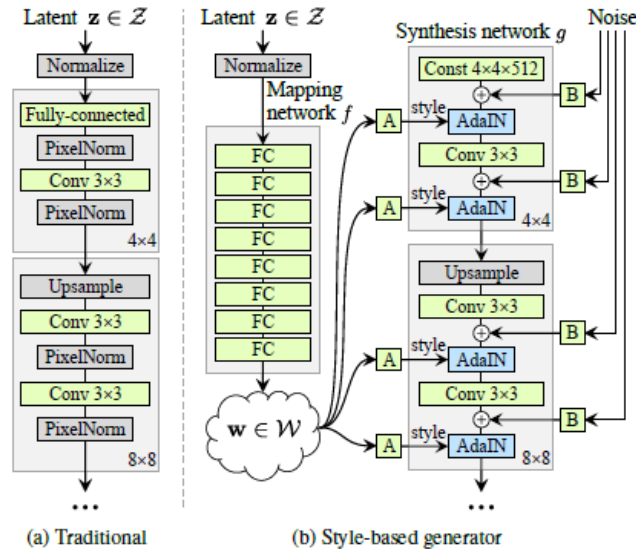
# High resolution GANs



- Synthesis of 1024x1024 face images [[Nvidia-ProGAN-2018](#)]
- Trained from CelebA-HQ dataset 30k images
- Progressive training
  - Complete GAN for low-resolution (4x4)
  - Upsample, concatenate with res-net connections
  - Train everything end-to-end



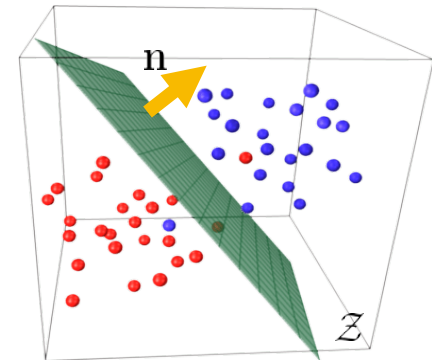
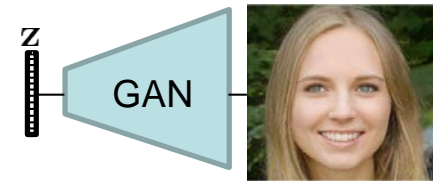
■ StyleGAN [[Karras-2019](#)] (Nvidia)



– Multi-layer style transfer, training from 70k Flickr dataset, “[hyper-realistic](#)”

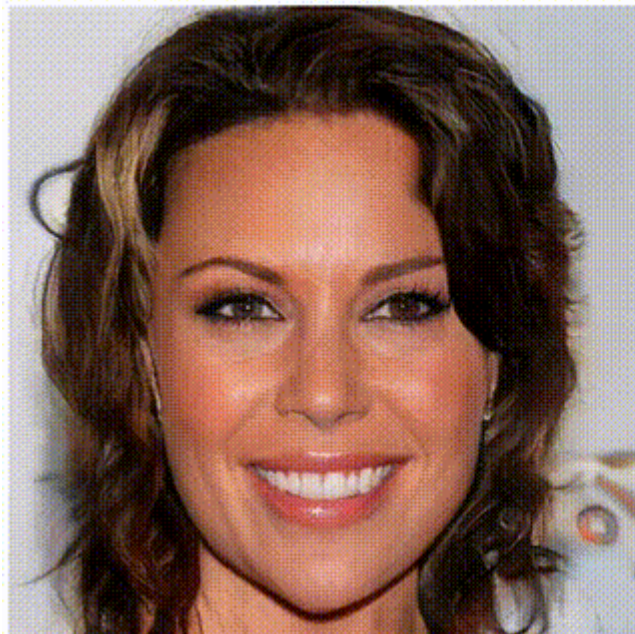
■ Follow-up paper [[Nvidia-2020](#), [Nvidia-2021](#), [Nvidia-2022](#)]

# GAN – latent space manipulation



- Every  $z$  from input distribution gives a realistic image
- Finding semantic direction in the latent vector space
  - Train a linear binary classifier on labeled set  $(z_i, y_i)$
  - Normal of the discriminative hyperplane is the semantic direction
- Semantic Editing / “Manipulation”  $z = z_0 + \alpha n$

INSTRUCTION: press +/- to adjust feature, toggle feature name to lock the feature



random face		
Male	Age	Skin_Tone
- +	- +	- +
Bangs	Hairline	Bald
- +	- +	- +
Big_Nose	Pointy_Nose	Makeup
- +	- +	- +
Smiling	Mouth_Open	Wavy_Hair
- +	- +	- +
Beard	Goatee	Sideburns
- +	- +	- +
Blond_Hair	Black_Hair	Gray_Hair
- +	- +	- +
Eyeglasses	Earrings	Necktie
- +	- +	- +

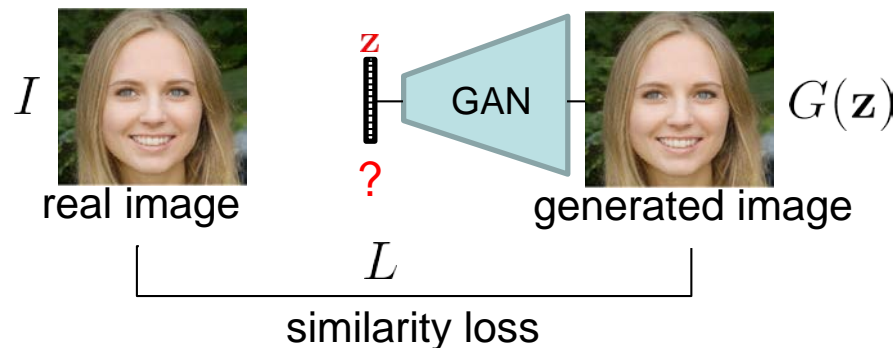


[demo]

[Abdal-SIGGRAF-2021]

# GAN Inversion (projection)

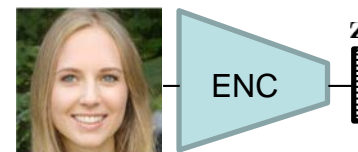
- Given an input real image  $I$ , find the latent code that generate the image



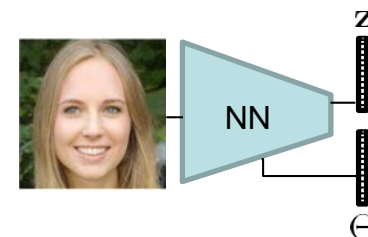
- Many approaches:

- Direct optimization
  - Works, but slow (~10s)
- Encoders [[Alaluf-2022](#), [Tov-2021](#)]
  - Very fast, but less accurate
- Pivotal tuning [[Roich-2021](#)]
  - Optimizes the GAN model parameters
- Hyper-networks [[Alaluf-2022](#)]
  - NN which adjust GAN model parameters
- Piece-wise inversion [[Šubrtová-2022](#)]
  - More degrees of freedom, very accurate, but slow

$$\min_{\mathbf{z}} L(I, G(\mathbf{z}))$$

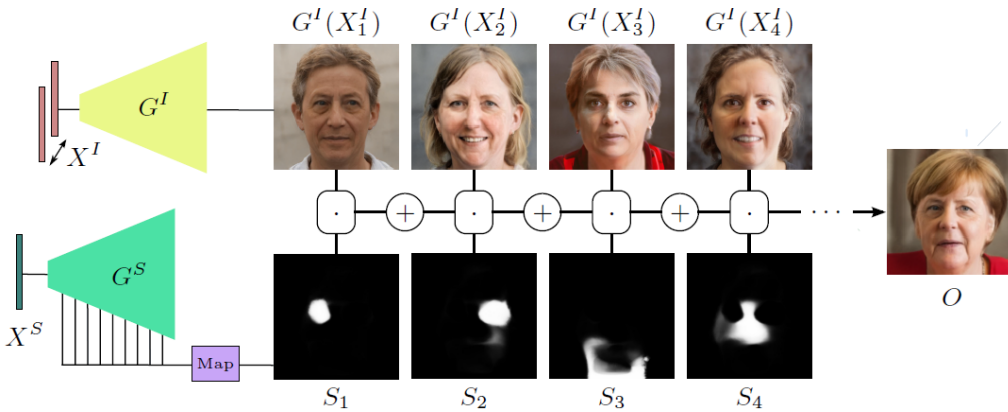


$$\min_{\mathbf{z}, \Theta} L(I, G(\mathbf{z}, \Theta))$$



# GAN Inversion (projection)

- Chunky GAN – piecewise image inversion [[Šubrtová-2022](#)]



$$O(X^I, S) = \sum_{i=1}^n G^I(X_i^I) \cdot S_i.$$

$$\min_{X^S, X^I} \mathcal{L}_{\text{LPIPS}}(I, \sum_{i=1}^n G^I(X_i^I) \cdot G^S(X^S)_i) + \lambda_{\text{reg}} \sum_{i=1}^n \|X_i^I - X_\mu^I\|_2^2.$$

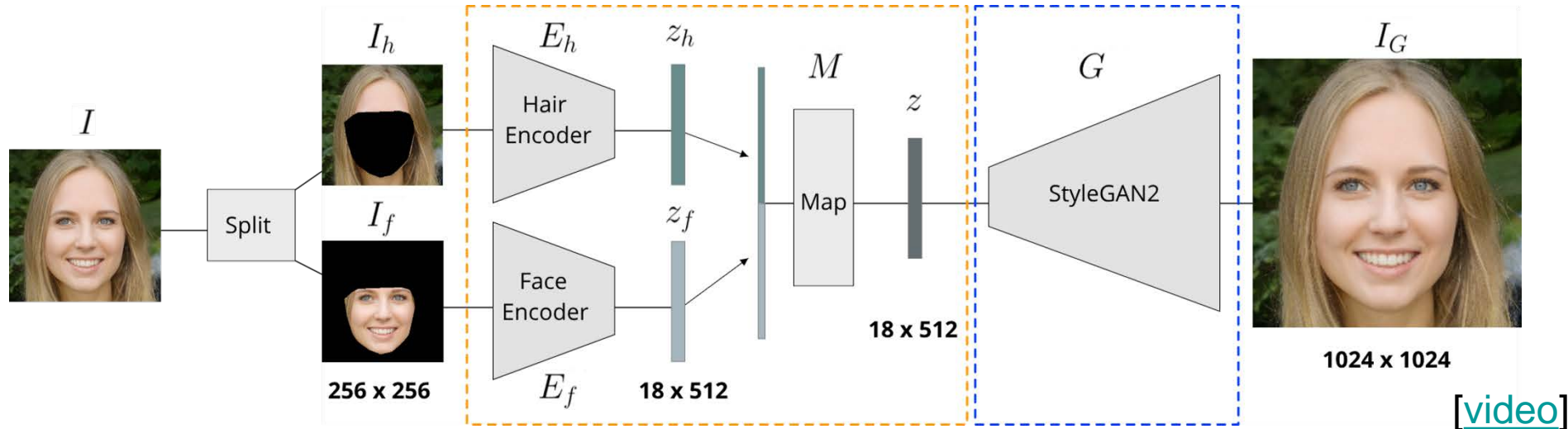
- Partial inversion, Interactive editing

## ChunkyGAN: Real Image Inversion via Segments

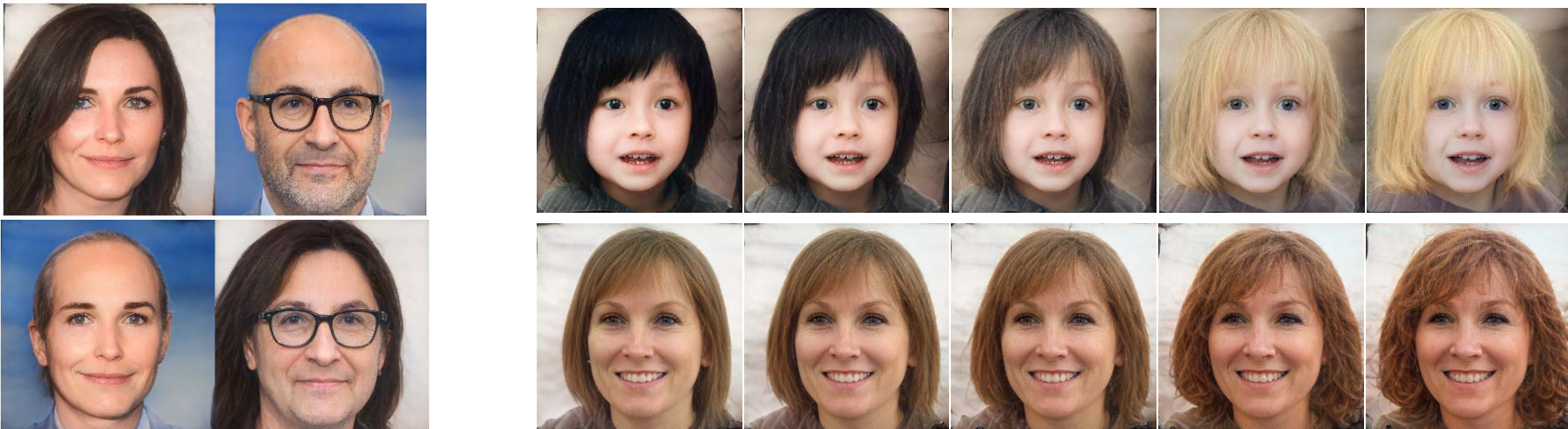


# Hairstyle Transfer using StyleGAN

- Fully automatic hairstyle transfer, unaligned portraits [[Šubrtová-FG-2021](#)]



- Basic idea: Train two encoders (Hair, face) + fixed StyleGAN decoder
- Hairstyle interpolation, Editing in hairstyle latent space





# Text-based Image Manipulation

- StyleCLIP [[Patashnik-2021](#)]
  - Text-Driven Manipulation of StyleGAN Imagery
  - Latent code manipulation driven by [CLIP](#) text-image similarity



Input

“Beyonce”

“A woman  
without makeup”“Elsa from  
Frozen”

Input

“A man with a  
beard”

“A blonde man”

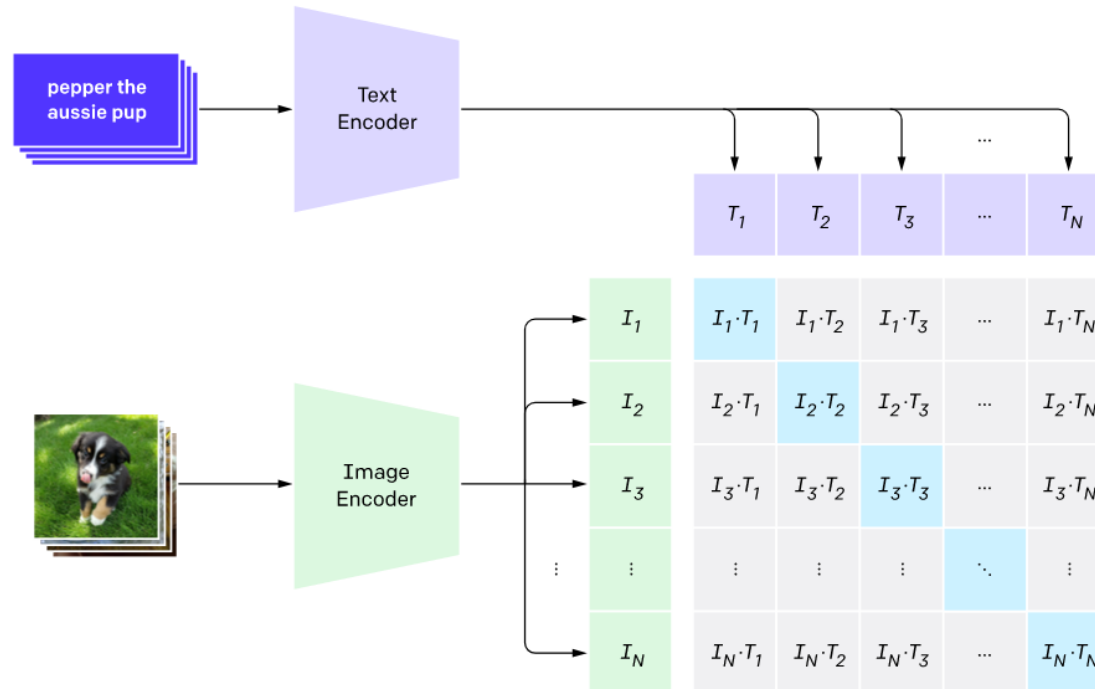
“Donald Trump”

$$\arg \min_{w \in \mathcal{W}^+} D_{\text{CLIP}}(G(w), t) + \lambda_{\text{L2}} \|w - w_s\|_2 + \lambda_{\text{ID}} \mathcal{L}_{\text{ID}}(w)$$

# CLIP – Connecting Text and Images (recap)



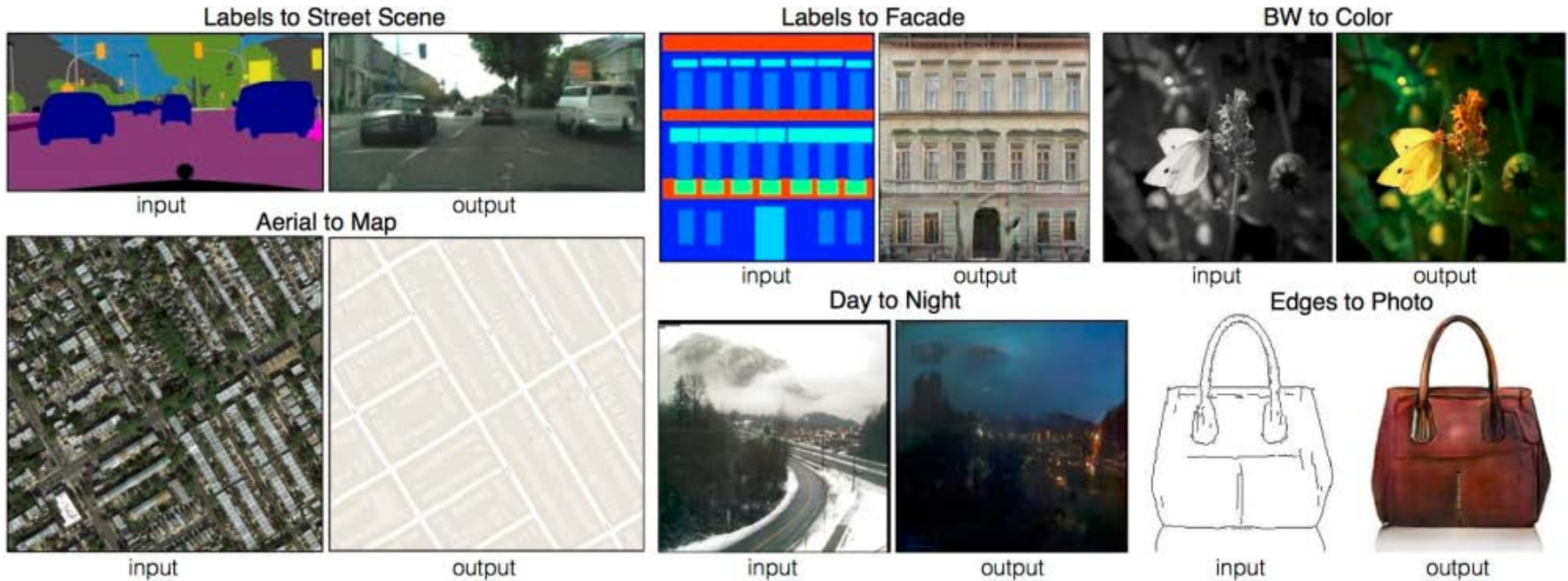
- CLIP [[Radford-2021](#)] by OpenAI
  - “*Contrastive Language–Image Pre-training*”
  - Learn joint text-image embedding => Text-image (cosine) similarity
  - Learned from 400M WebImageText (WIT) dataset



- Zero-shot prediction (on par with Resnet on ImageNET benchmark)
  - Loop over ImageNET-classes:  $\max \text{CLIP}(E_T(\text{"A photo of a <class>"}), E_I(I))$
- Trained [model](#) publicly available

# Image to Image Translation

- Transfer image between domains [[Isola-2017](#)]

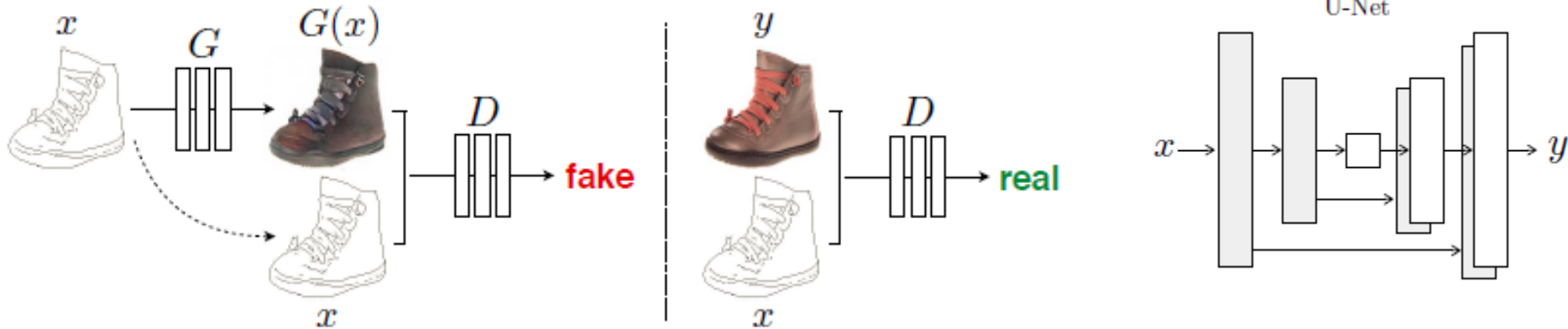


- Many applications [[pix2pix](#)], Super-resolution [[Šubrtová-2018](#)]



# Image to Image Translation

- Combines fully convolutional net training with (conditional) GAN



$$G^* = \arg \min_G \max_D \mathcal{L}_{cGAN}(G, D) + \lambda \mathcal{L}_{L1}(G)$$

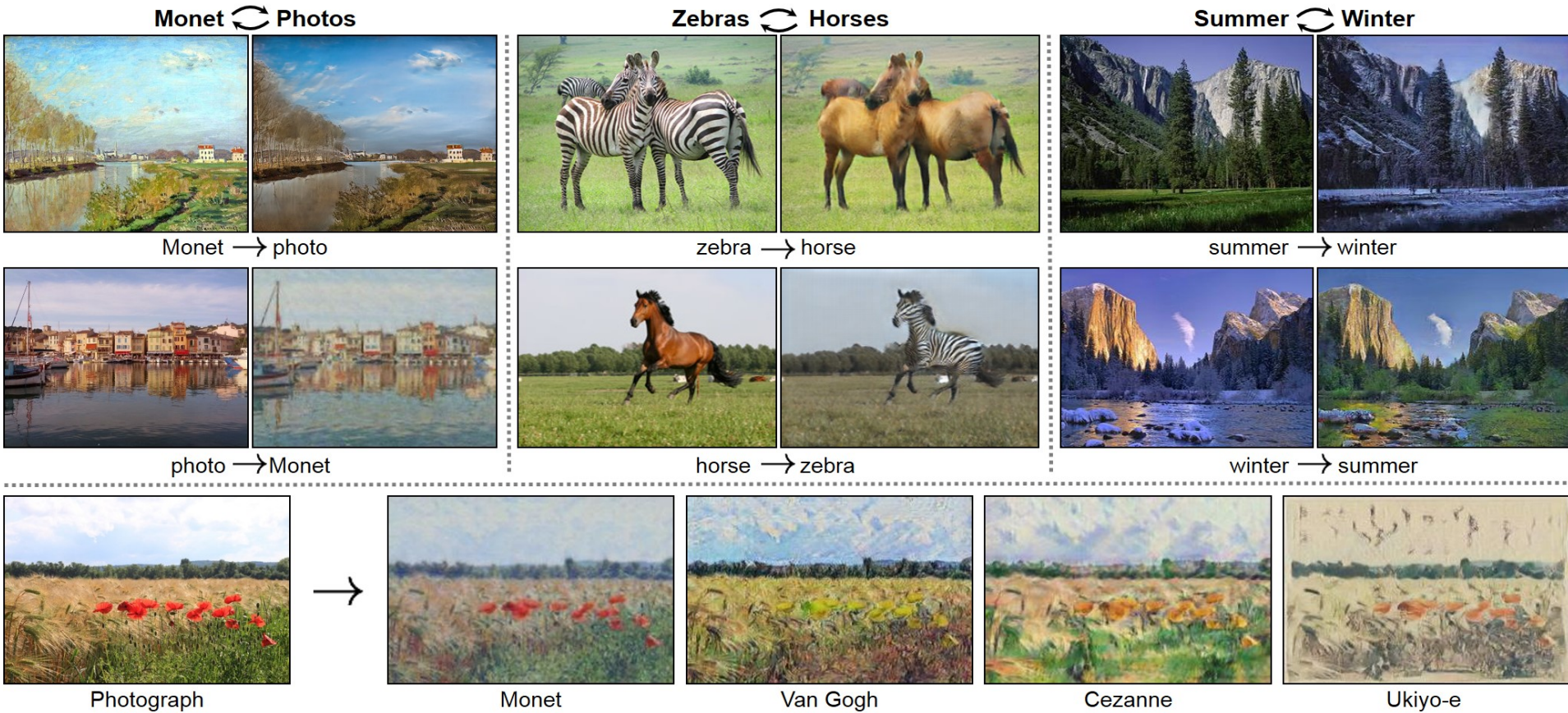
$$\mathcal{L}_{L1}(G) = \mathbb{E}_{x,y,z} [\|y - G(x, z)\|_1]$$

$$\mathcal{L}_{cGAN}(G, D) = \mathbb{E}_{x,y} [\log D(x, y)] + \mathbb{E}_{x,z} [\log(1 - D(x, G(x, z)))]$$

- Difficulties with imposing variability (only via dropout when testing)
- Training needs pixel-to-pixel source and target image correspondences

# Cycle GAN

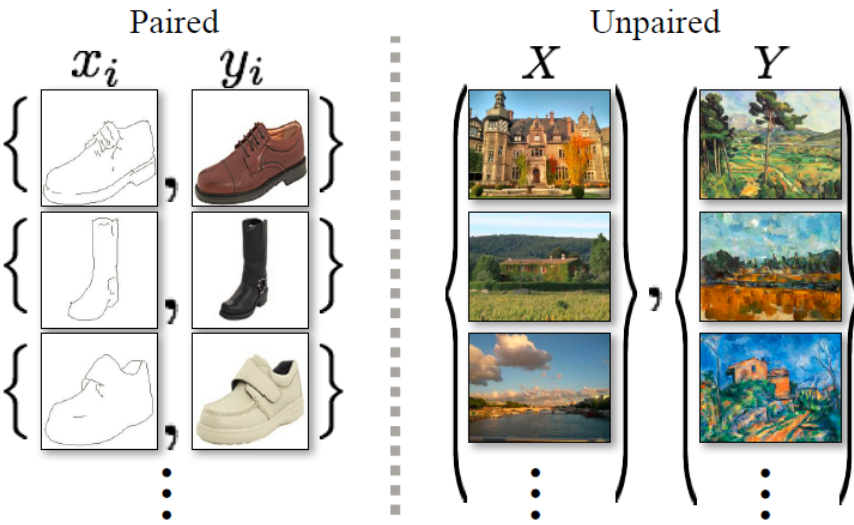
- Translating without pix-to-pix correspondences [[Zhu-2017](#)]



# Cycle GAN



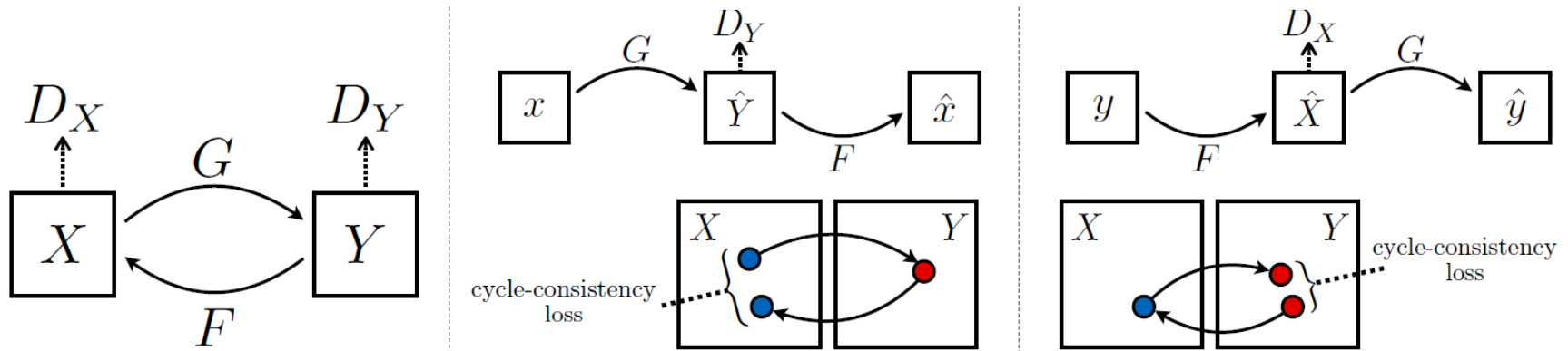
- Unpaired set of images to train the translation



$$\mathcal{L}(G, F, D_X, D_Y) = \mathcal{L}_{GAN}(G, D_Y, X, Y) + \mathcal{L}_{GAN}(F, D_X, Y, X) + \lambda \mathcal{L}_{cyc}(G, F),$$

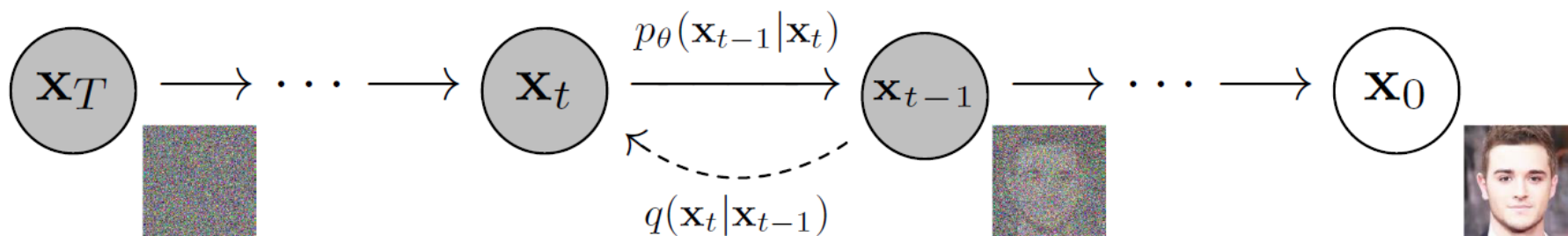
$$\mathcal{L}_{cyc}(G, F) = \mathbb{E}_{x \sim p_{data}(x)} [\|F(G(x)) - x\|_1] + \mathbb{E}_{y \sim p_{data}(y)} [\|G(F(y)) - y\|_1]$$

- Cycle consistency



# Diffusion Models

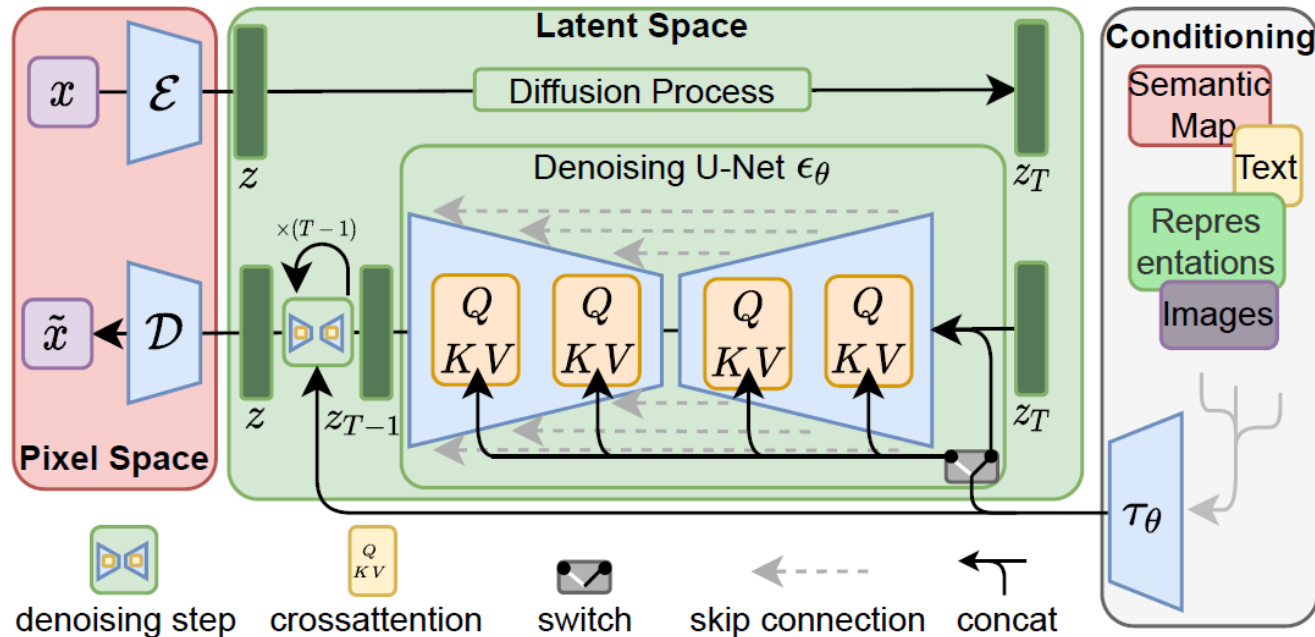
- Inspiration from thermodynamics [[Sohl-Dickstein-2015](#)]
- Denoising diffusion probabilistic models [[Ho-2020](#)]
- Main principle:
  - Forward diffusion: progressively destroy the data by injecting noise
    - Gaussian Noise
  - Reverse diffusion: learn to reverse the process to sample generation
    - denoising U-NET



$$L_{DM} = \mathbb{E}_{x, \epsilon \sim \mathcal{N}(0,1), t} \left[ \|\epsilon - \epsilon_\theta(x_t, t)\|_2^2 \right]$$



- Stable Diffusion [[Rombach-2022](#)]
  - Latent diffusion model (diffusion/denoising runs in latent space)
  - Encoder/decoder pixel-latent space learnt offline
  - Conditioning by cross-attention (Text, Semantic maps, ...)



$$L_{LDM} := \mathbb{E}_{\mathcal{E}(x), y, \epsilon \sim \mathcal{N}(0,1), t} \left[ \|\epsilon - \epsilon_\theta(z_t, t, \tau_\theta(y))\|_2^2 \right]$$

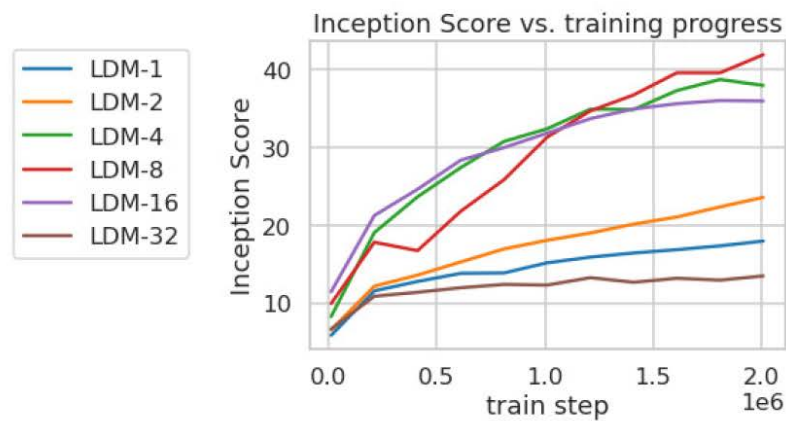
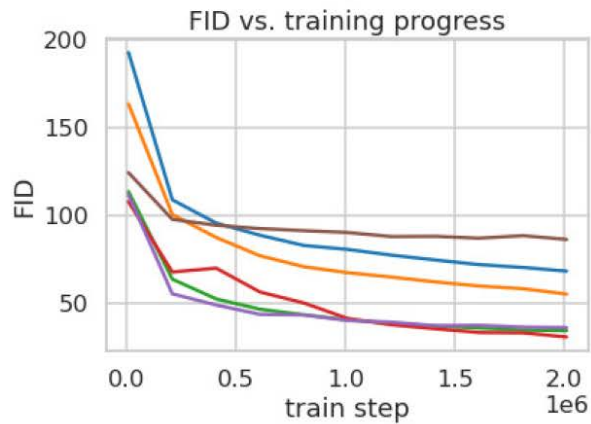
# Stable Diffusion



- Unconditioned generation (256x256)



- Effect of the latent “compression”



LDM-1  
(pixel level)

LDM-8  
(default)

# Stable Diffusion

- Text-to-Image
  - Trained on LAION dataset - 1.45B
  - CLIP embedding for text

'A street sign that reads "Latent Diffusion"'

'A zombie in the style of Picasso'

'An image of an animal half mouse half octopus'

'An illustration of a slightly conscious neural network'

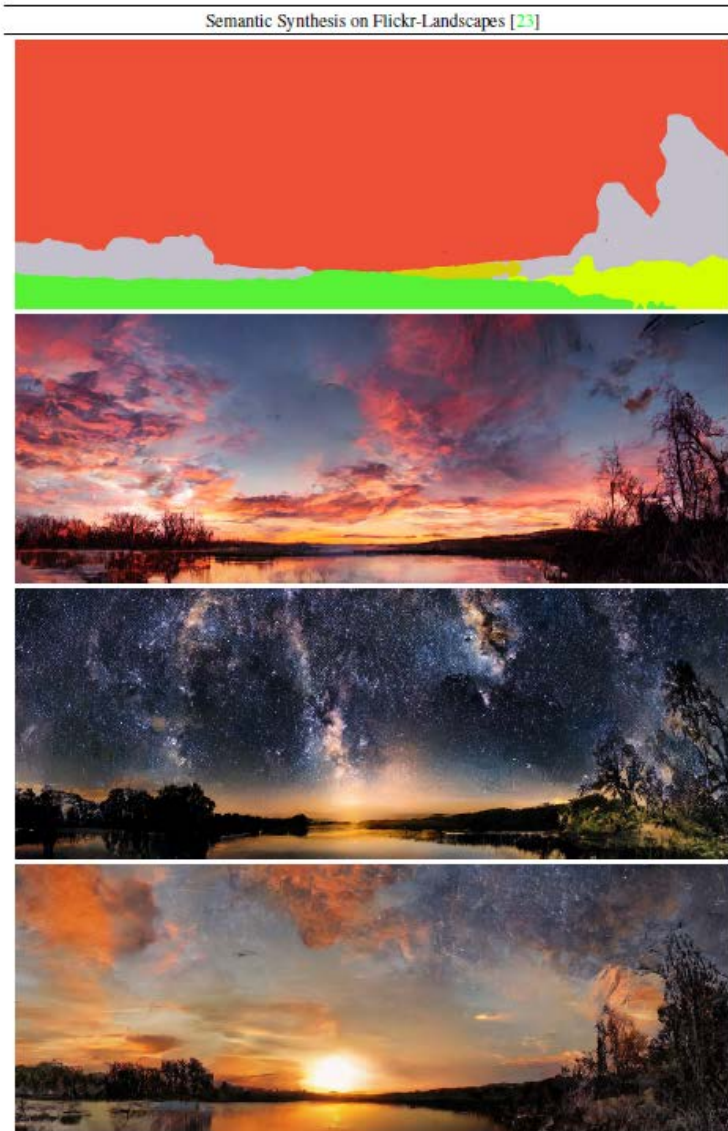
'A painting of a squirrel eating a burger'

'A watercolor painting of a chair that looks like an octopus'

'A shirt with the inscription: "I love generative models!"'



- Semantic maps



- Inpainting



# Classifier-Free Guidance (CFG)



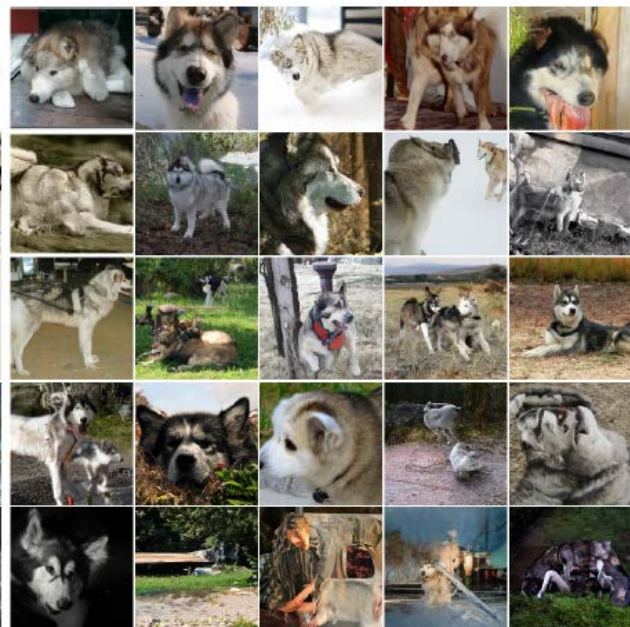
37

- Classifier-free diffusion guidance [[Ho-2022](#)]
- Controls adherence with the input prompt
  - Trade-off mode coverage and sample fidelity
- Training: Diffusion model is trained both conditional and unconditional
  - Unconditional sample (null prompt) is given in certain probability  $p_{\text{uncond}}$  (e.g. 0.1/0.2/0.5)
- Inference:
  - Sample a weighted combination of conditioned and unconditional denoising model in each step

$$\tilde{\epsilon}_{\theta}(\mathbf{z}_{\lambda}, \mathbf{c}) = (1 + w)\epsilon_{\theta}(\mathbf{z}_{\lambda}, \mathbf{c}) - w\epsilon_{\theta}(\mathbf{z}_{\lambda})$$

# Classifier-Free Guidance (CFG)

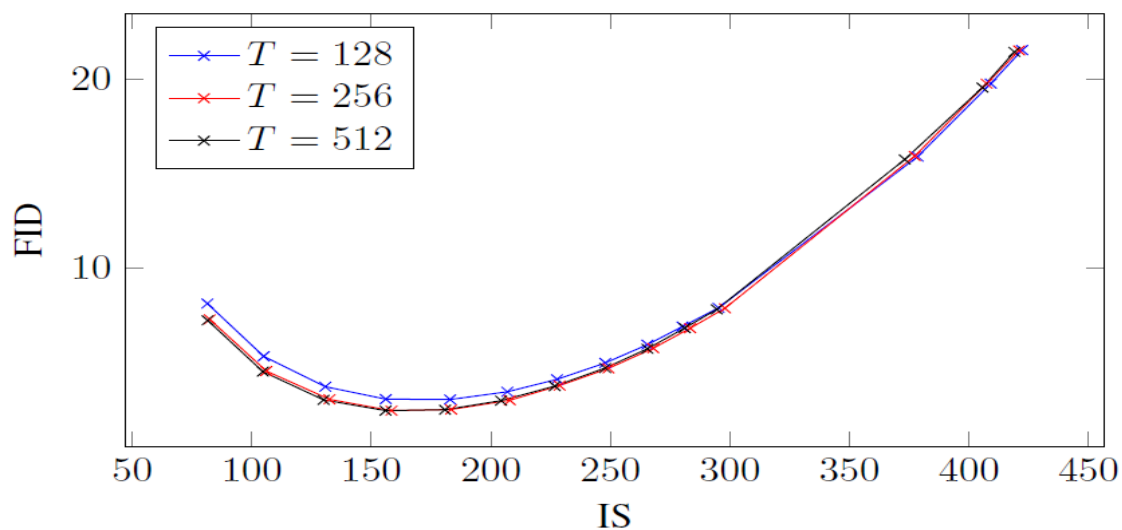
$w = 0$



$w = 1$



$w = 4$



# Text2image using Textual Inversion

- An image is worth one word... [[Gal-2022](#)]
- Text to image with custom objects



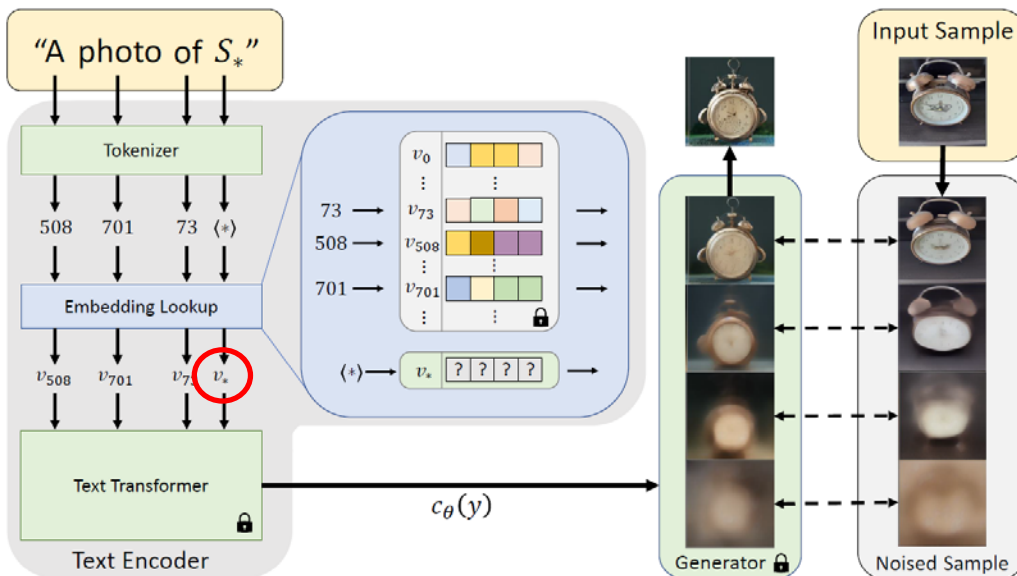
Input samples  $\xrightarrow{\text{invert}}$  “ $S_*$ ”

“Painting of two  $S_*$  fishing on a boat”

“A  $S_*$  backpack”

“Banksy art of  $S_*$ ”

“A  $S_*$  themed lunchbox”



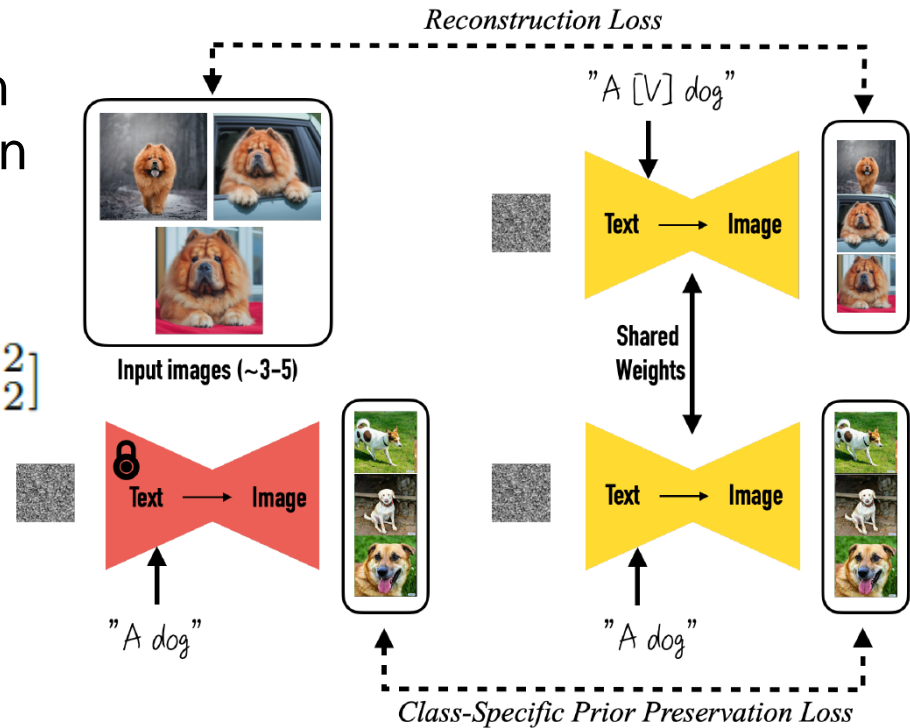
$$v_* = \arg \min_v \mathbb{E}_{z \sim \mathcal{E}(x), y, \epsilon \sim \mathcal{N}(0,1), t} \left[ \|\epsilon - \epsilon_\theta(z_t, t, c_\theta(y))\|_2^2 \right]$$

# DreamBooth

- DreamBooth [Ruiz-2022] (Google)
- **Fine-Tuning** Text-to-Image Diffusion Models for Subject-Driven Generation

$$\mathbb{E}_{\mathbf{x}, \mathbf{c}, \epsilon, \epsilon', t} [\omega_t \|\hat{\mathbf{x}}_{\theta}(\alpha_t \mathbf{x} + \sigma_t \epsilon, \mathbf{c}) - \mathbf{x}\|_2^2 + \lambda \omega_{t'} \|\hat{\mathbf{x}}_{\theta}(\alpha_{t'} \mathbf{x}_{pr} + \sigma_{t'} \epsilon', \mathbf{c}_{pr}) - \mathbf{x}_{pr}\|_2^2]$$

– about 5 mins on A100 GPU



Input images



in the Acropolis



swimming



sleeping



in a doghouse



in a bucket



getting a haircut



# Fine-tuning Diffusion Models



- LoRA (Low-Rank Adaptation) [[Hu-2021](#)] (Microsoft)
  - General light-weight adaptation of any models (including LLMs)
    - Faster training and storing
  - Many models available (e.g., [CivitAI](#), [PromptHero](#))

$$W' = W + BA$$

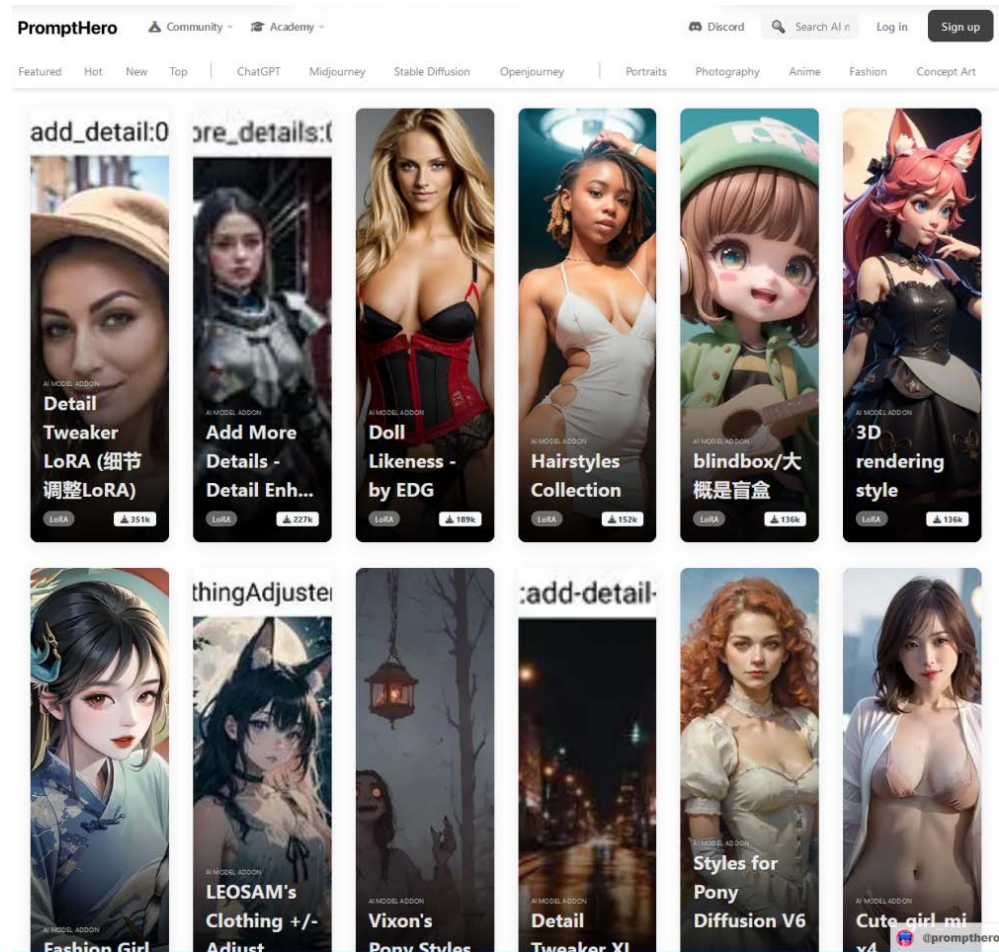
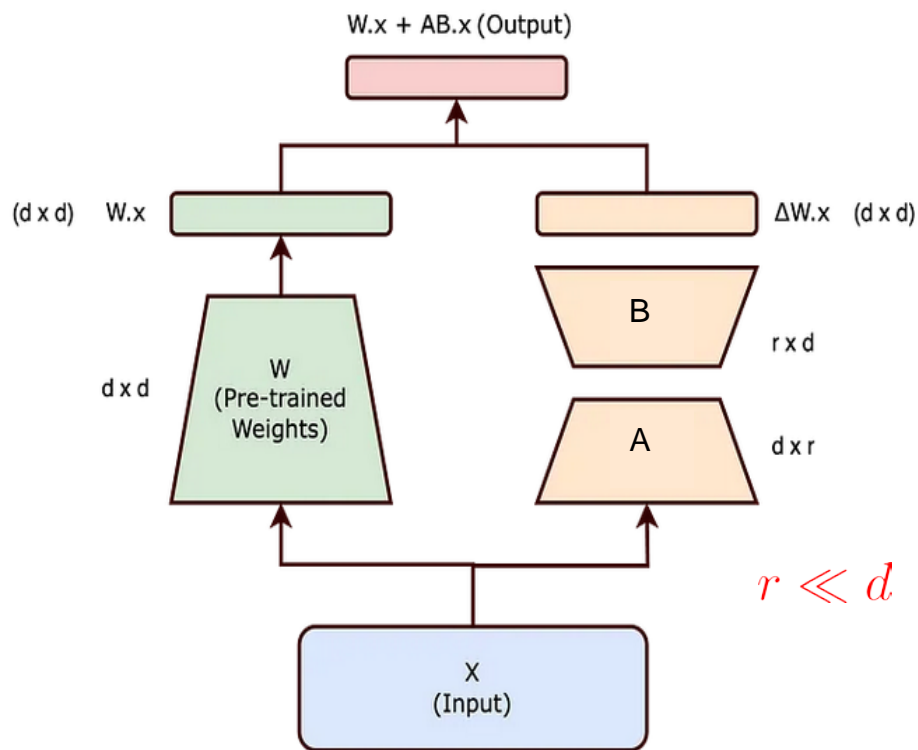


Figure credit Bhavin Jawade

## Adding Conditional Control to Text-to-Image Diffusion Models [[Zhang-2023](#)]



Input Canny edge



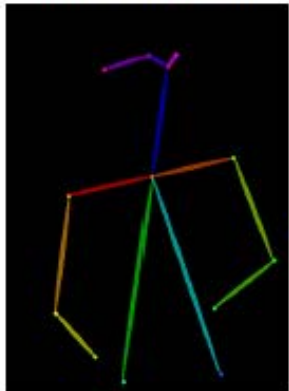
Default



"masterpiece of fairy tale, giant deer, golden antlers"



"..., quaint city Galic"



Input human pose



Default



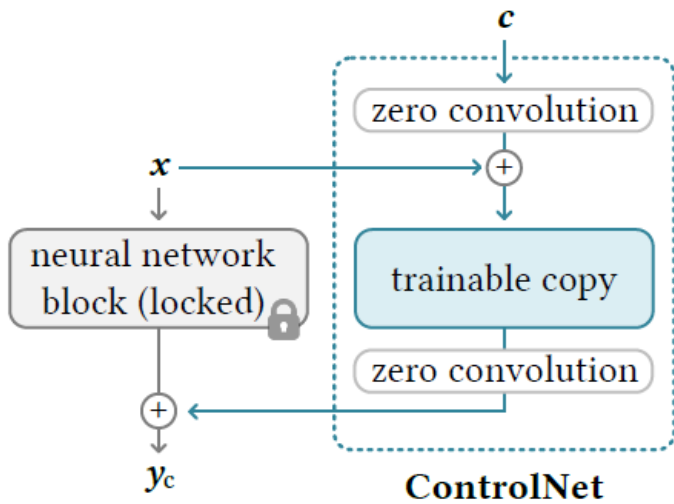
"chef in kitchen"



"Lincoln statue"

# ControlNet

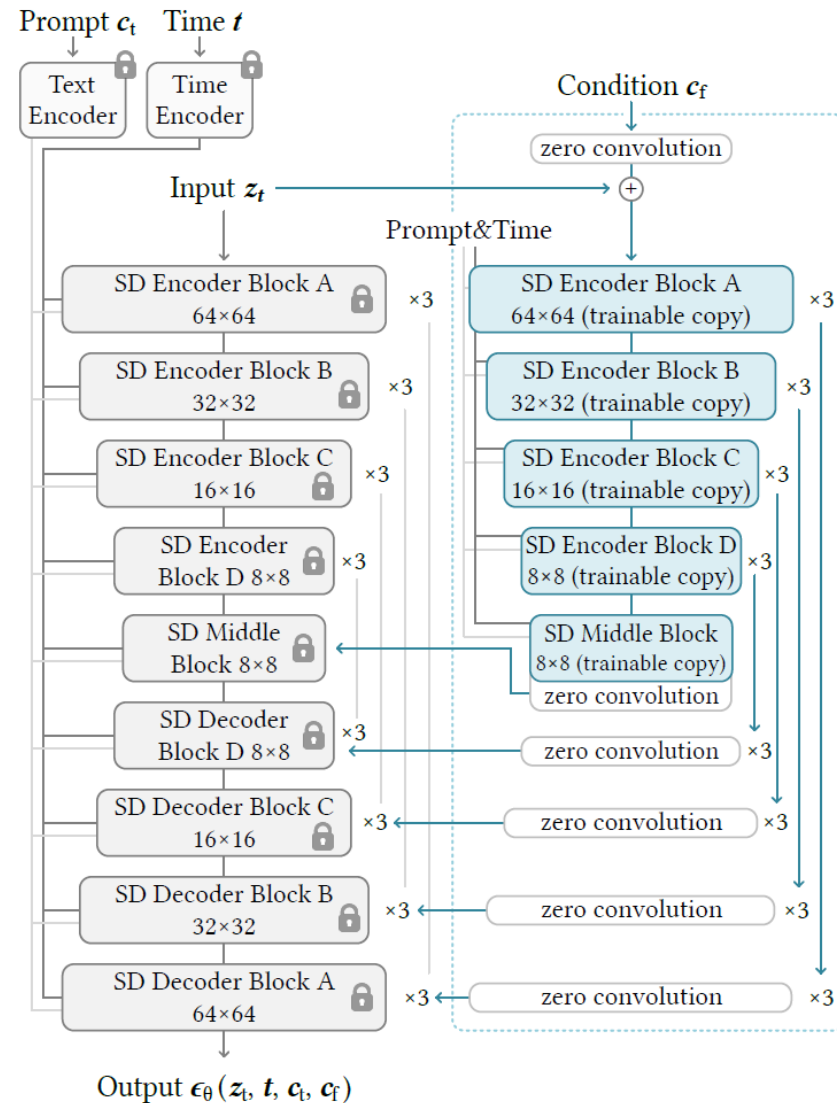
- Fine-tuning with a trainable copy
  - Zero convolutions: 1x1 convolutions with weights initialized to zeros



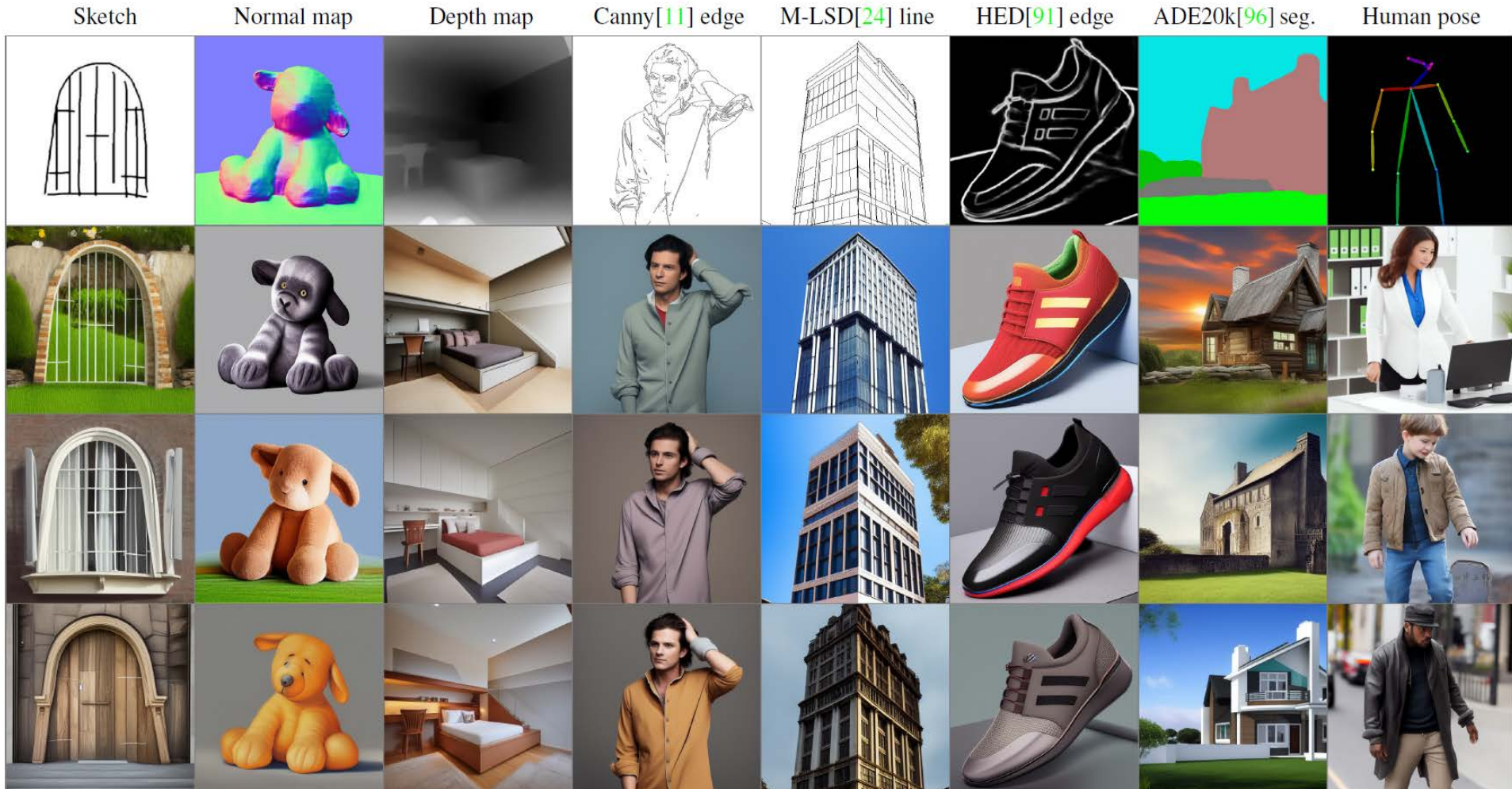
Extra cond.

$$\mathcal{L} = \mathbb{E}_{z_0, t, c_t, c_f, \epsilon \sim \mathcal{N}(0,1)} \left[ \|\epsilon - \epsilon_\theta(z_t, t, c_t, c_f)\|_2^2 \right]$$

- 50% of text prompt  $c_t$  randomly replaced with empty string



- Models trained for multiple conditioning



- [Demo](#), [code and models](#) available

# Prompt2Prompt



## ■ Prompt-to-Prompt image editing [Hertz-2022]



“The boulevards are crowded today.”  
↓ ↓ ↓ ↓ ↓ ↓ ↓ ↓



“Photo of a cat riding on a bicycle.”  
~~car~~



“Landscape with a house near a river  
and a rainbow in the background?”



“My fluffy bunny doll.”  
↑ ↑ ↑ ↑ ↑



“a cake with decorations.”  
jelly beans



“Children drawing of a castle next to a river.”

fixed random seed



“lemon cake.”



“apple cake.”



“lego cake.”

## – Generated image editing

- Stress/Weaken words, Changing words, Adding new phrases
- Fixing random seed does not help (layout changes)

fixed random seed and attention maps



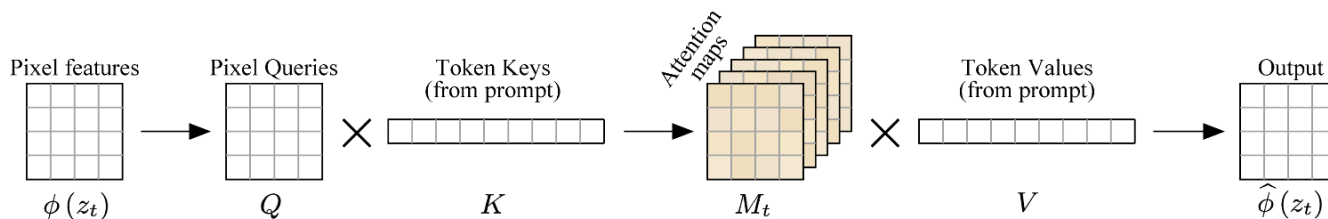
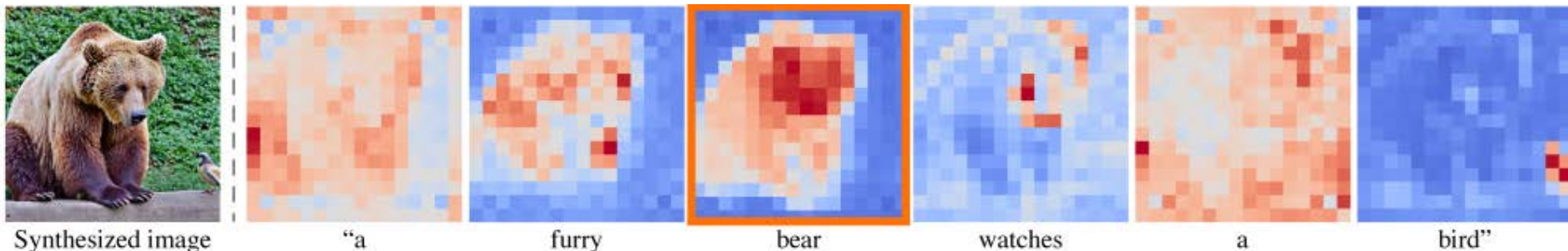
“apple cake.”



“lego cake.”

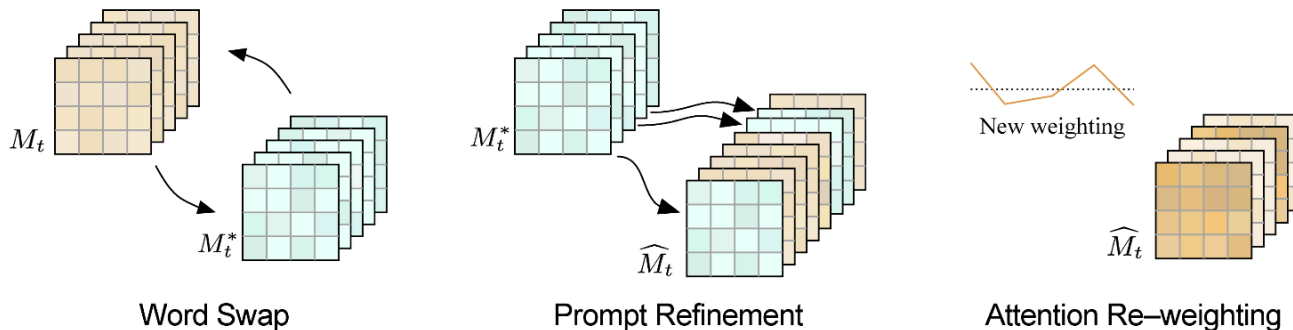
# Prompt2Prompt

- Controlling the Cross-Attention (text x image)
  - Attention maps are responsible for spatial layout
  - Keep existing attention maps, change/add new word maps, reweight



Text to Image Cross Attention

Cross Attention Control



# Instruct Pix2Pix

- Instruct Pix2Pix [[Brooks-2023](#)]
  - Textual editing or real images

*"Swap sunflowers with roses"*



*"Add fireworks to the sky"*



*"Replace the fruits with cake"*



*"What would it look like if it were snowing?"*



*"Turn it into a still from a western"*



*"Make his jacket out of leather"*



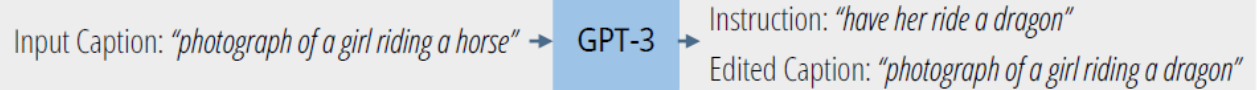
# Instruct Pix2Pix



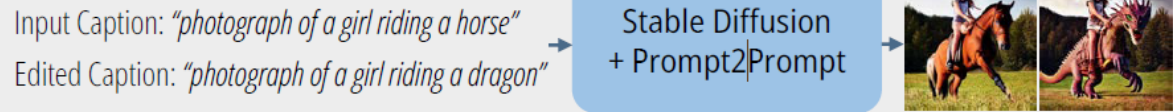
- New model trained from synthetic captions (Chat GPT) and images edited by Prompt-to-Prompt

## Training Data Generation

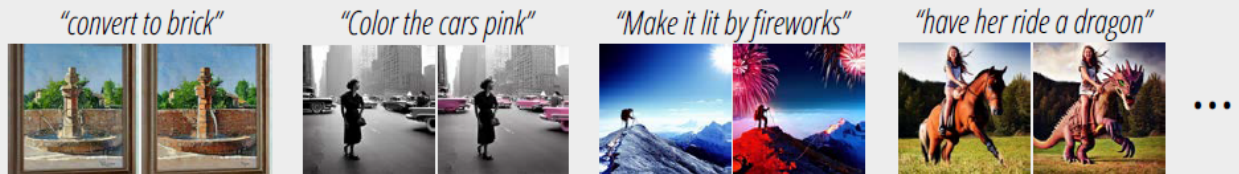
(a) Generate text edits:



(b) Generate paired images:



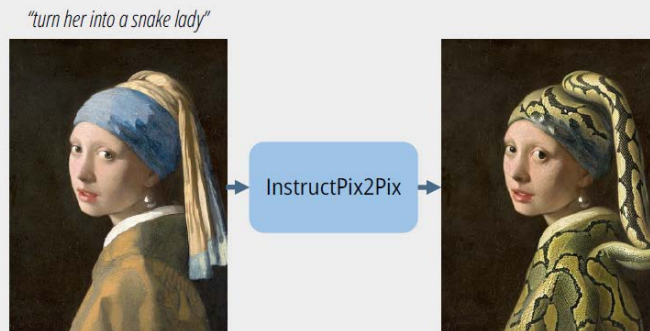
(c) Generated training examples:



450k training examples

## Instruction-following Diffusion Model

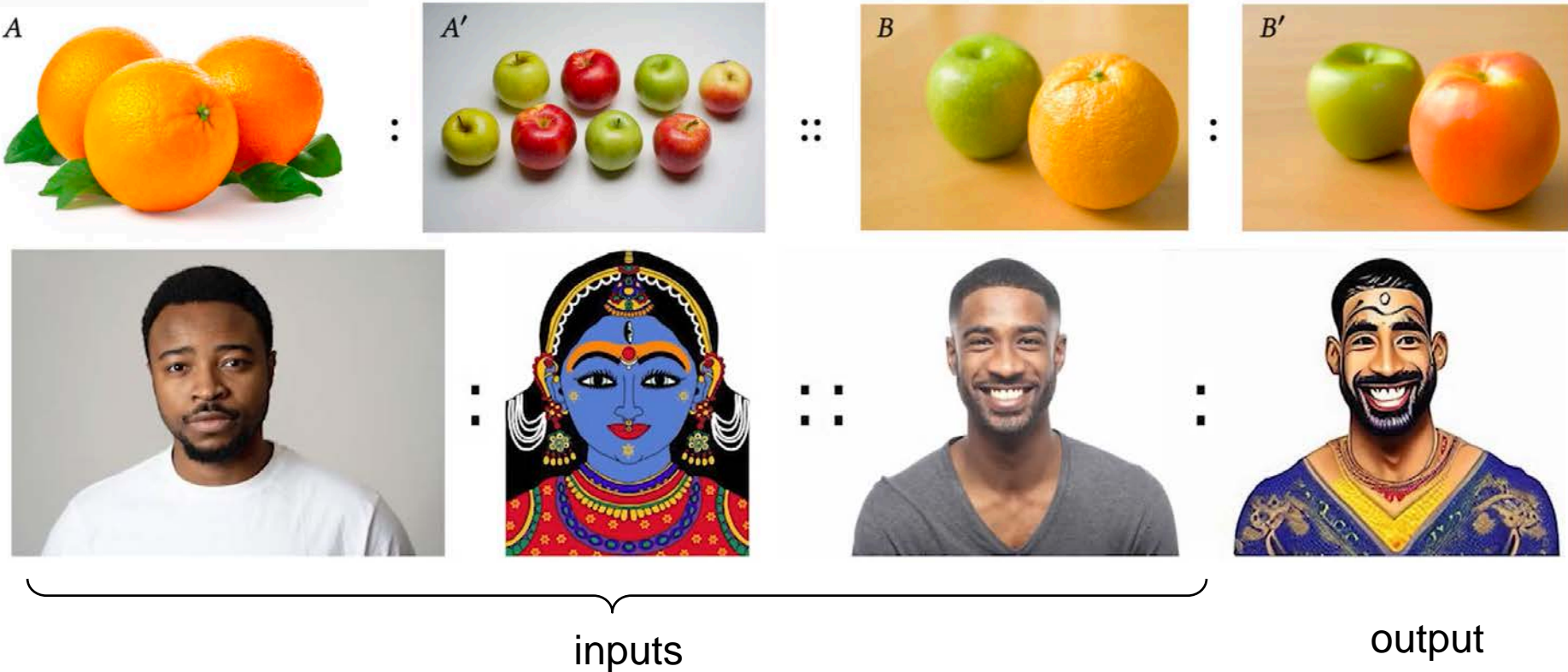
(d) Inference on real images:





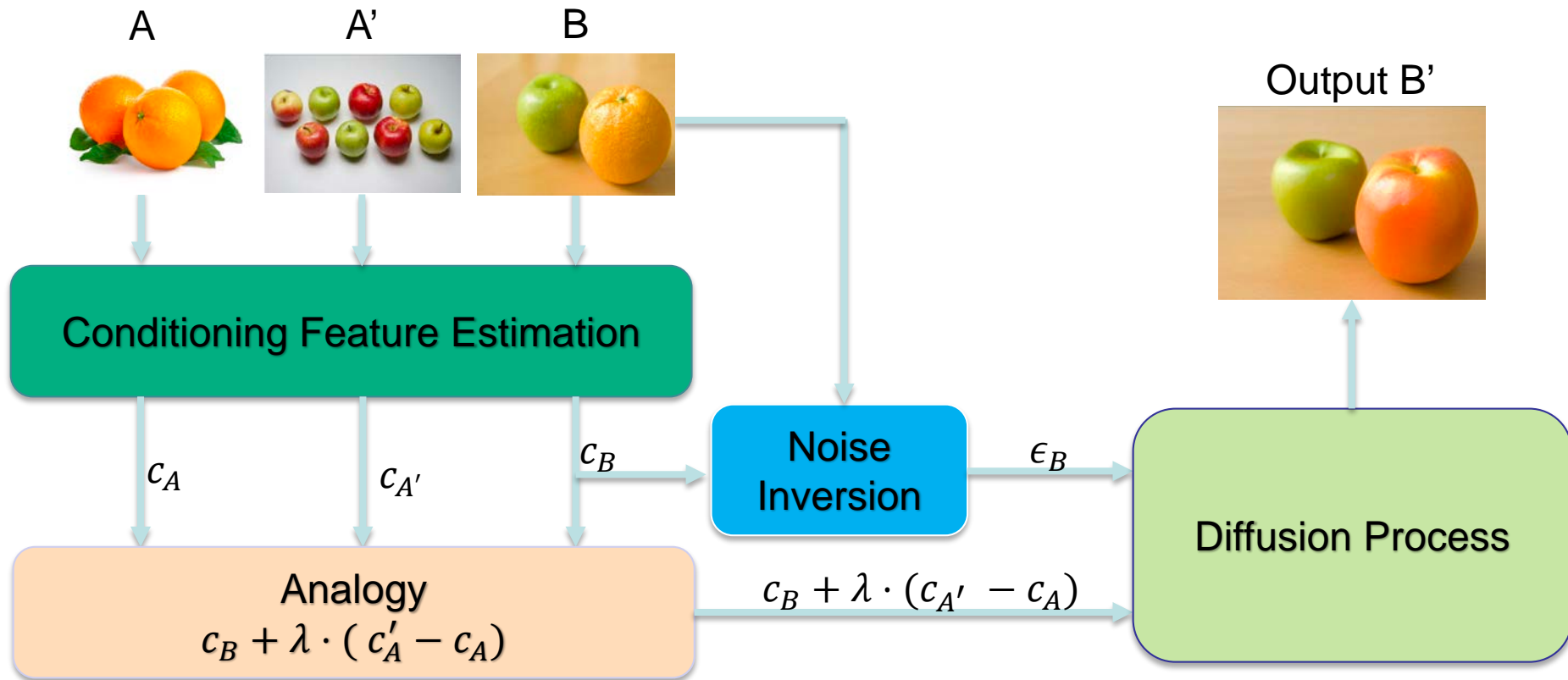
# Diffusion Image Analogies

## ■ Diffusion Image Analogies [[Šubrtová-SIGGRAPH-2023](#)]



- Real image editing by visual analogies
- Relation between  $B'$  and  $B$  is analogous to relation between  $A'$  and  $A$
- Exploiting “algebra” of latent space of the Stable Diffusion

# Diffusion Image Analogies

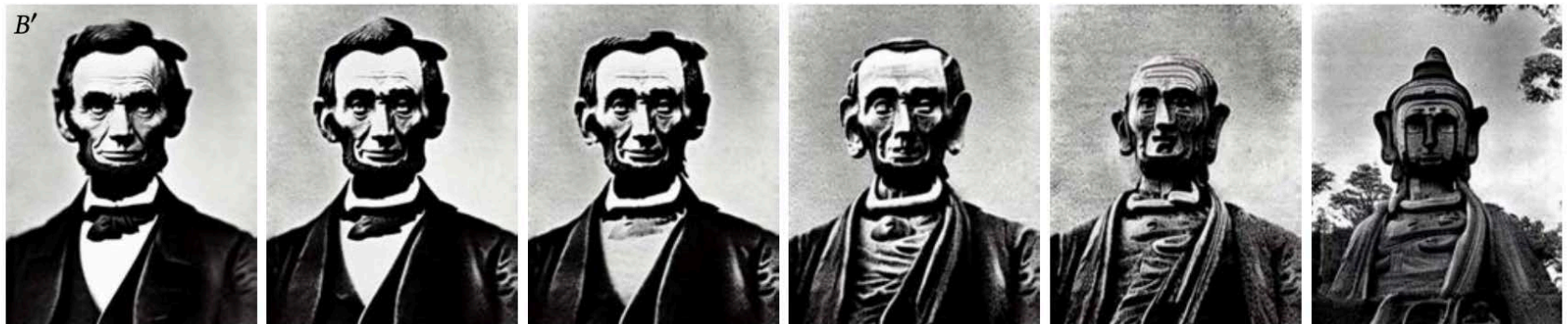


- Conditioning and Noise Inversion found by optimization
- Parameter  $\lambda$  controls the strength of the analogy

# Diffusion Image Analogies



increasing analogy strength  $\lambda$  →



increasing analogy strength  $\lambda$  →

# Conclusions



52

- “Photorealistic” quality has been achieved in the last couple of years
  - GANs, Diffusion models
- Deep generative models have evolved dramatically recently
- Interest / Fear of artistic community
  - Many tools greatly support content creativity
  - Certain creative artists feel endangered and exploited (some models likely trained on data without author’s permission)
- Threat of high-quality deep fakes easy to perpetrate
  - Fake news, fake porn, etc.