

Diffusion models for Image Restoration

Daniela Ivanova

Computer Vision and Autonomous Systems (CVAS)



What is image restoration?



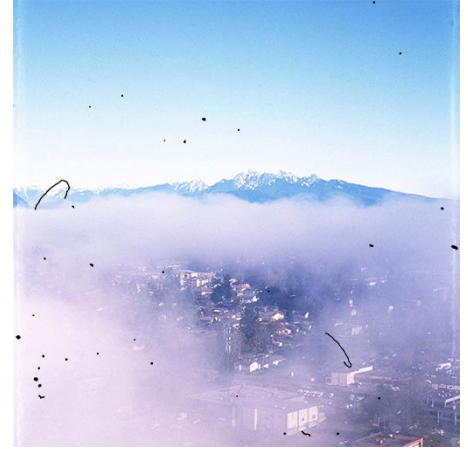
Colourisation



Super-
resolution



Inpainting
(regular
shape)

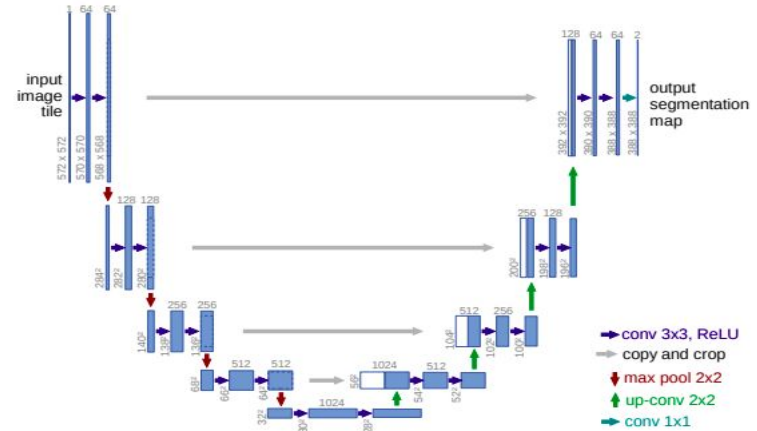


Inpainting
(arbitrary
artefacts)



- Since the *input, output and latents have the same dimensionality*, for each step the denoising model is a U-Net with some modifications
 - Group normalisation
 - Global self-attention
 - Sinusoidal positional **time embeddings** concatenated to the input of each block
- Instead of predicting the denoised image, the network predicts the noise that was added to it

Diffusion architecture - recap



Credit: Ronneberger et al, "Convolutional Networks for Biomedical Image Segmentation"



Diffusion process - recap



- To produce each latent, we can add noise iteratively (slow)

Distribution of noised images

$$q(x_1, \dots, x_T | x_0) := \prod_{t=1}^T q(x_t | x_{t-1}) \quad (1)$$

$$q(x_t | x_{t-1}) := \mathcal{N}(x_t; \sqrt{1 - \beta_t} x_{t-1}, \beta_t \mathbf{I}) \quad (2)$$

scaling
diagonal covariance matrix



Diffusion process - recap



noised latents directly conditioned on the input x_0 . With $\alpha_t := 1 - \beta_t$ and $\bar{\alpha}_t := \prod_{s=0}^t \alpha_s$, we can write the marginal

$$q(x_t|x_0) = \mathcal{N}(x_t; \sqrt{\bar{\alpha}_t}x_0, (1 - \bar{\alpha}_t)\mathbf{I}) \quad (8)$$

$$x_t = \sqrt{\bar{\alpha}_t}x_0 + \sqrt{1 - \bar{\alpha}_t}\epsilon \quad (9)$$

where $\epsilon \sim \mathcal{N}(0, \mathbf{I})$. Here, $1 - \bar{\alpha}_t$ tells us the variance of the noise for an arbitrary timestep, and we could equivalently use this to define the noise schedule instead of β_t .

- ...or we can derive the variance scale for an arbitrary step - accumulate the noise from the first step to the step we need



- We can condition on **class labels** - embed class label v_i along with time embedding e_t - (Nichol & Dhariwal, 2021)

Credit: Nichol & Dhariwal,
“Improved denoising diffusion
probabilistic models.”



Diffusion conditioning - recap

- Classifier guidance (Dhariwal & Nichol, 2021)
 - y comes from the downsampling half of the UNet, which is used as a classifier



Figure 6: Samples from BigGAN-deep with truncation 1.0 (FID 6.95, left) vs samples from our diffusion model with guidance (FID 4.59, middle) and samples from the training set (right).

Credit: Dhariwal & Nichol,
“Diffusion Models Beat GANs on
Image Synthesis”

- Classifier guidance (Dhariwal & Nichol, 2021)

$$p_{\theta, \phi}(x_t | x_{t+1}, y) = Z p_{\theta}(x_t | x_{t+1}) p_{\phi}(y | x_t)$$

$$\begin{aligned} \nabla_{x_t} \log(p_{\theta}(x_t) p_{\phi}(y | x_t)) &= \nabla_{x_t} \log p_{\theta}(x_t) + \nabla_{x_t} \log p_{\phi}(y | x_t) \\ &= -\frac{1}{\sqrt{1 - \bar{\alpha}_t}} \epsilon_{\theta}(x_t) + \nabla_{x_t} \log p_{\phi}(y | x_t) \end{aligned}$$

$$\hat{\epsilon}(x_t) := \epsilon_{\theta}(x_t) - \sqrt{1 - \bar{\alpha}_t} \nabla_{x_t} \log p_{\phi}(y | x_t)$$



Diffusion conditioning - recap



“a hedgehog using a calculator”



“a corgi wearing a red bowtie and a purple party hat”



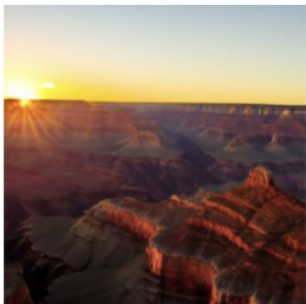
“robots meditating in a vipassana retreat”



“a fall landscape with a small cottage next to a lake”



“a surrealist dream-like oil painting by salvador dalí of a cat playing checkers”



“a professional photo of a sunset behind the grand canyon”



“a high-quality oil painting of a psychedelic hamster dragon”



“an illustration of albert einstein wearing a superhero costume”

- We can condition on text descriptions
 - Each attention layer is attending to each token for the text embedding
 - Doesn't work very well still
- Credit: Dhariwal & Nichol, “GLIDE: Towards Photorealistic Image Generation and Editing with Text-Guided Diffusion Models”



Diffusion conditioning - recap

- CLIP guided diffusion:
 - At inference time, use CLIP guidance
 - CLIP outputs a similarity score between image and text for each pixel
 - Use that gradient at each time step to push the image in the direction which would give it higher score/smaller CLIP loss



(c) GLIDE (CLIP guidance, scale 2.0)



Diffusion conditioning - recap

- Classifier-free guidance:
 - Train with and without text embeddings
 - Predict an image without the text prompt and with the text prompt
 - Find the difference between the two
 - Use that gradient to go in the direction of the image with text using a scaling factor for the vector



(d) GLIDE (Classifier-free guidance, scale 3.0)



Diffusion models for image restoration

End-to-end training with conditioning

- SR3
- Palette

Using pre-trained models, conditioning only during inference

- DDRM
- RePaint
- Stable Diffusion
- DiffEdit (bonus!)



SR3 (Saharia et al.)



- We can condition on another image
 - low res image for superresolution
 - grayscale image for colourisation
 - images with missing patches for inpainting
- Concatenate noise vector with the conditioning image
- Slow, diffusing the entire image

Credit: Saharia et al, "Image Super-Resolution via Iterative Refinement"



Palette: Image-to-Image Diffusion Models (Saharia et al.)



Credit: Saharia et al, "Palette:
Image-to-Image Diffusion
Models"



Diffusion model for film artifact removal

input

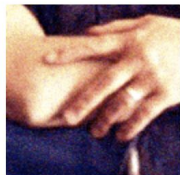
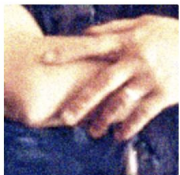
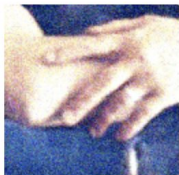
reconstruction 3
weeks ago

reconstruction 2
weeks ago

reconstruction
now

GT

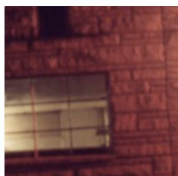
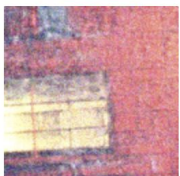
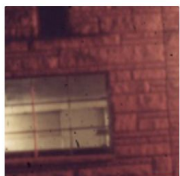
Medium level damage:



High level damage:



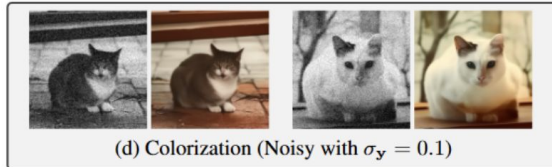
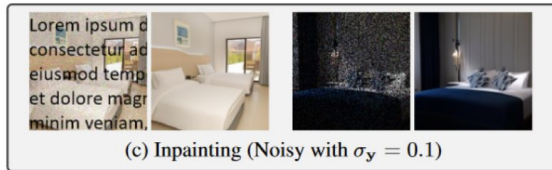
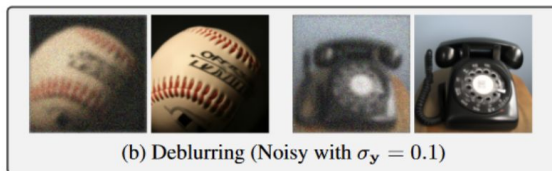
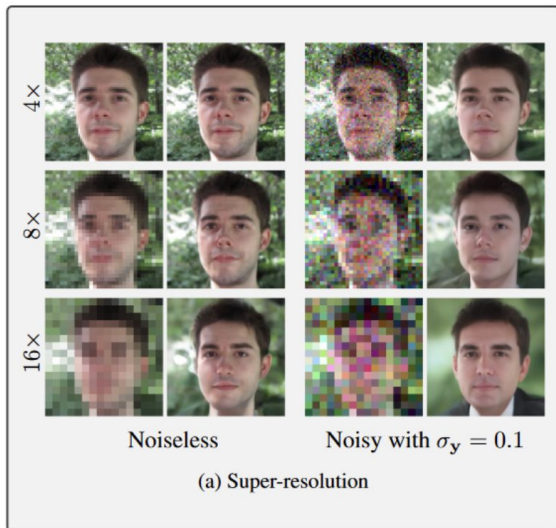
Low level damage:



- 97.8 M params
- bs=8
- 1M iterations in SR3 paper
- **very** slow if you don't have a TPU
- task-specific - need lots of data



DDRM (Kawar et al., 2022)



- use pretrained unconditional DDPM
- decompose degradation operator H using SVD
- perform diffusion in spectral space

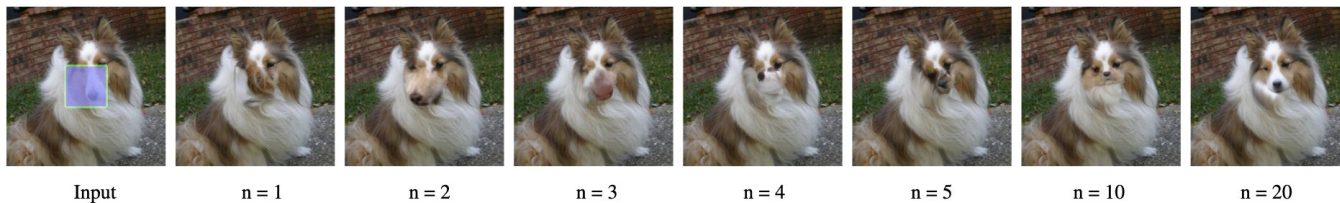
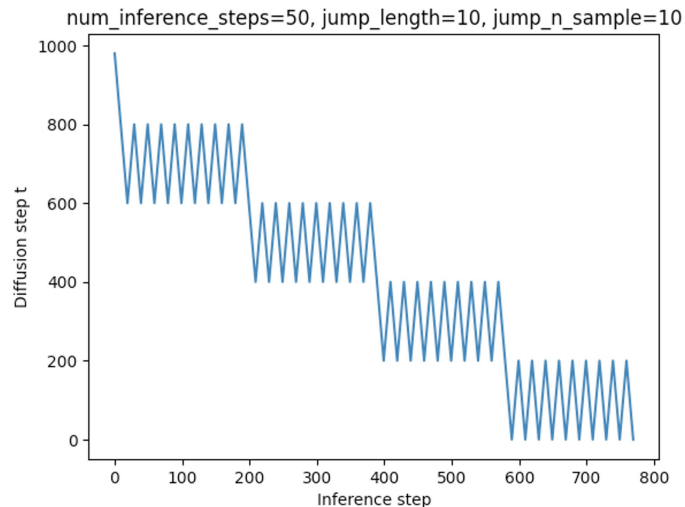
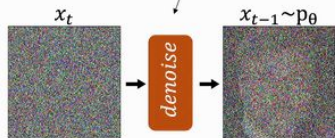
Credit: Kawar et al, "Denoising Diffusion Restoration Models"

$$\mathbf{y} = \mathbf{H}\mathbf{x} + \mathbf{z}, \quad q(\mathbf{x}_{1:T}|\mathbf{x}_0, \mathbf{y}) = q^{(T)}(\mathbf{x}_T|\mathbf{x}_0, \mathbf{y}) \prod_{t=0}^{T-1} q^{(t)}(\mathbf{x}_t|\mathbf{x}_{t+1}, \mathbf{x}_0, \mathbf{y}),$$



RePaint (Lugmayr et al., 2022)

Pretrained Unconditional Diffusion Model

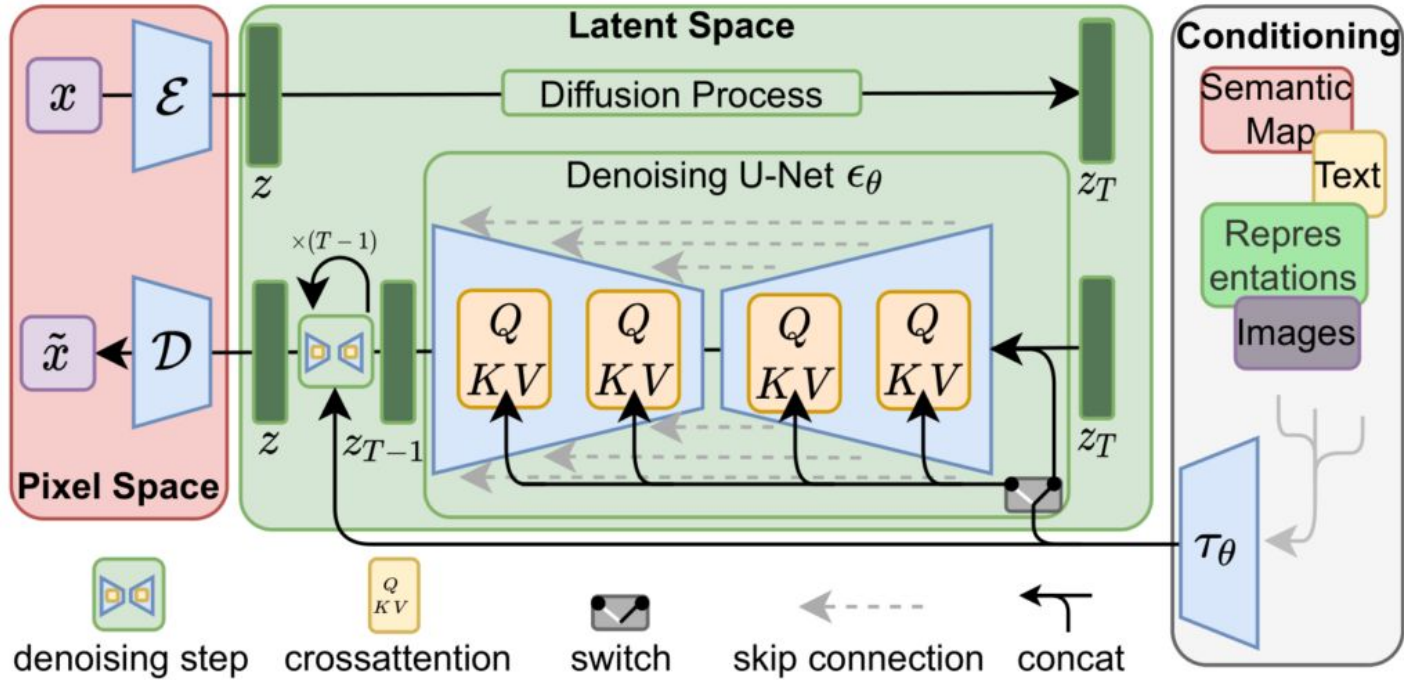


Credit: Lugmayr et al, "RePaint: Inpainting using Denoising Diffusion Probabilistic Models"

Figure 3. The effect of applying n sampling steps. The first example with $n = 1$ is the DDPM baseline, the second with $n = 2$ is with one resample step. More resampling steps lead to more harmonized images. The benefit saturates at about $n = 10$ resamplings.

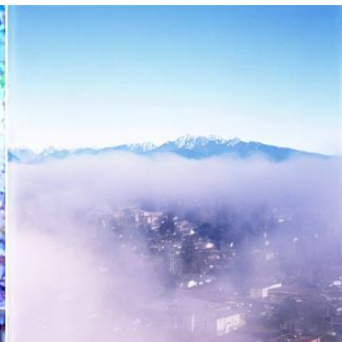
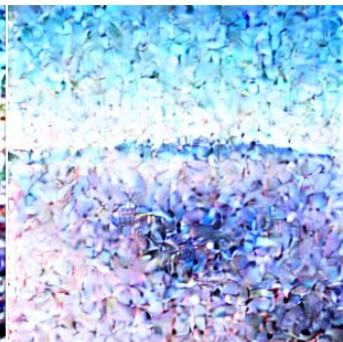
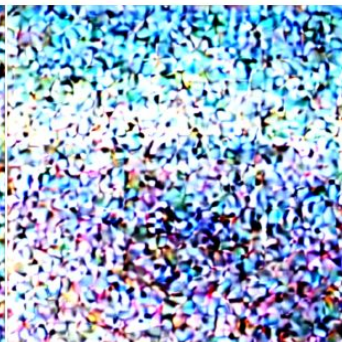
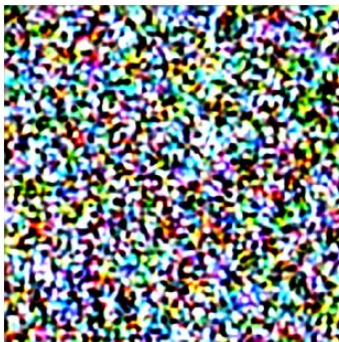
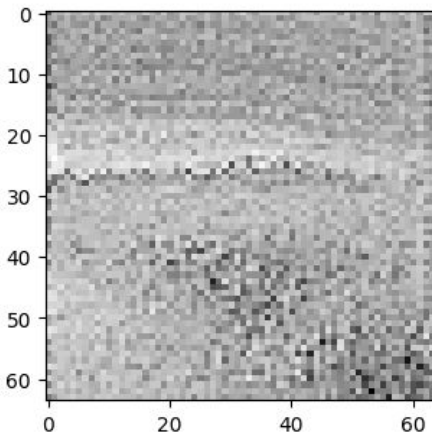
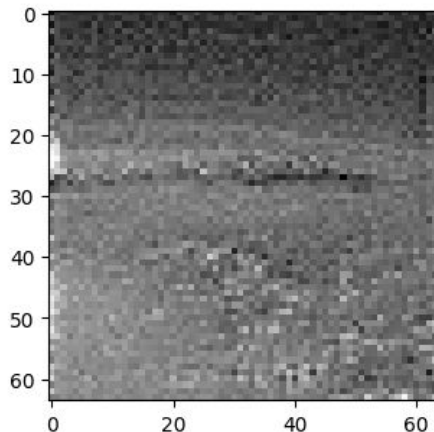
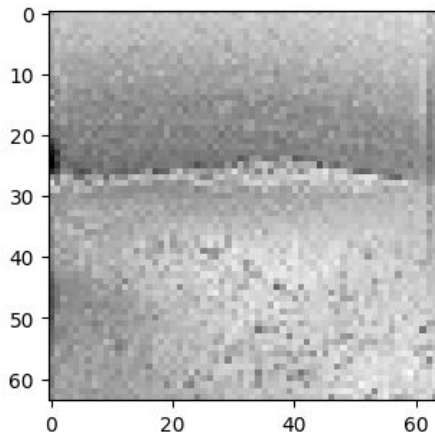
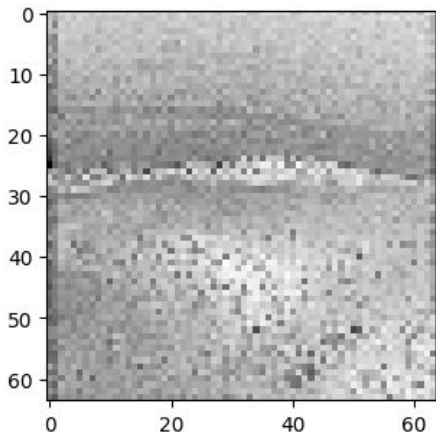


Latent diffusion (Romach et al.)





Latent diffusion (Romach et al.)





University
of Glasgow

Stable Diffusion Inpainting



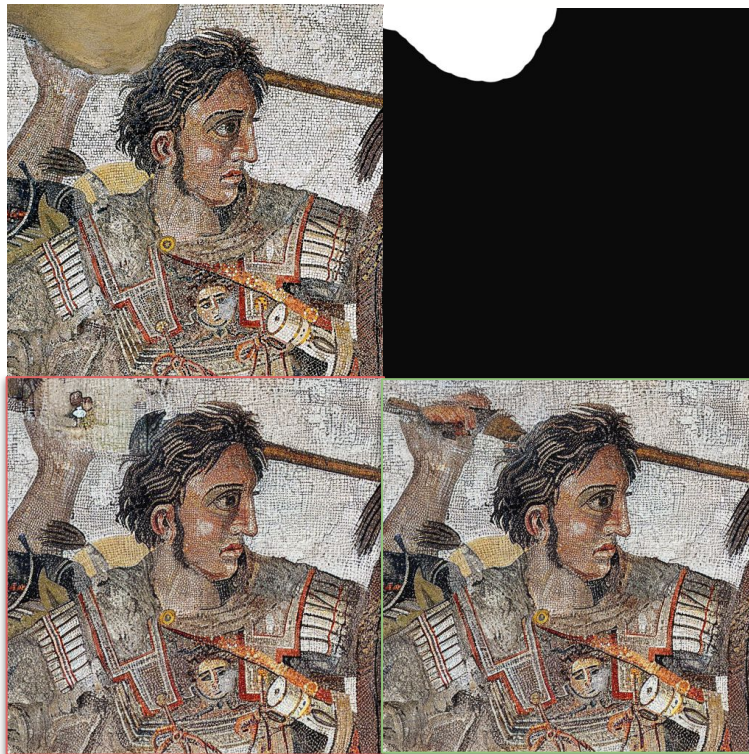
“cat sitting on a bench”



Stable Diffusion Inpainting

Unconditional (legacy)

- use pre-trained SD
- make prediction from noise
- mask out the latents
- make next prediction
- etc



Conditional

- fine-tune pre-trained SD model on inpainting
- pass the mask, masked latents, original latents as a 9-channel input to the U-Net



What if we tried RePaint it in latent space?





RePaint (Lugmayr et al., 2022)

What if we tried it in latent space?

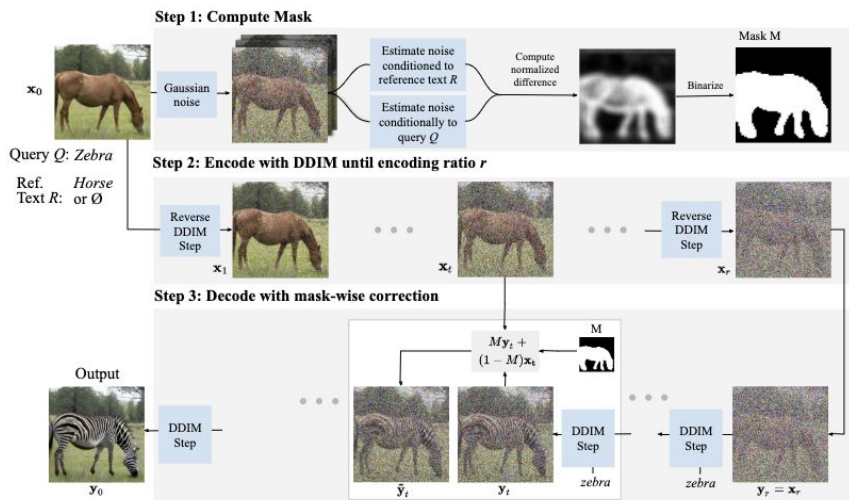


- Only works with the sampler used to train SD



DiffEdit (Couairon et al., 2022)

- denoise once using reference text
- denoise again using query text
- the difference in noise estimates => locations that are predicted to change the most between conditioning on the original and new texts



Credit: Couairon et al., "DiffEdit: Diffusion-based semantic image editing with mask guidance"



DiffEdit (Couairon et al., 2022)

reference: “horse”

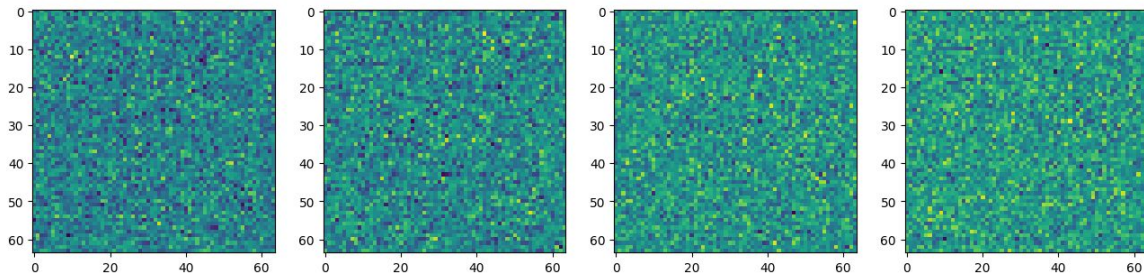
query: “zebra”

- noise-denoise 10 times with each prompt
- accumulate predicted noises
- find difference

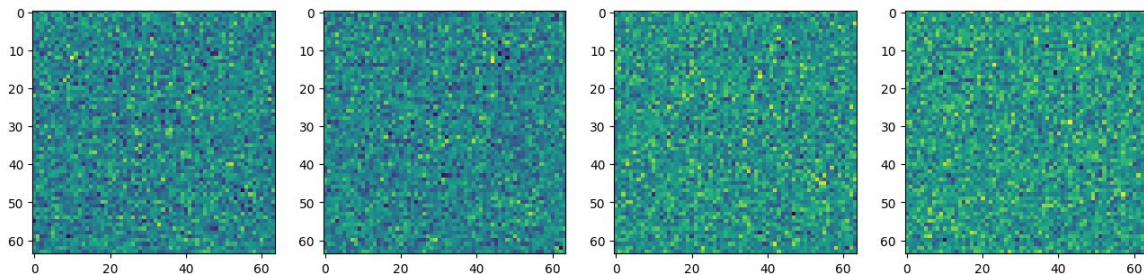
Check out [“DiffEdit paper implementation”](#) by Kevin Bird



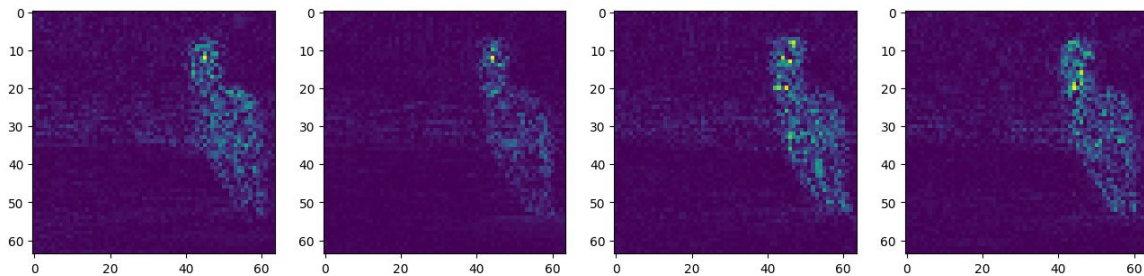
DiffEdit (Couairon et al., 2022)



“horse” noise



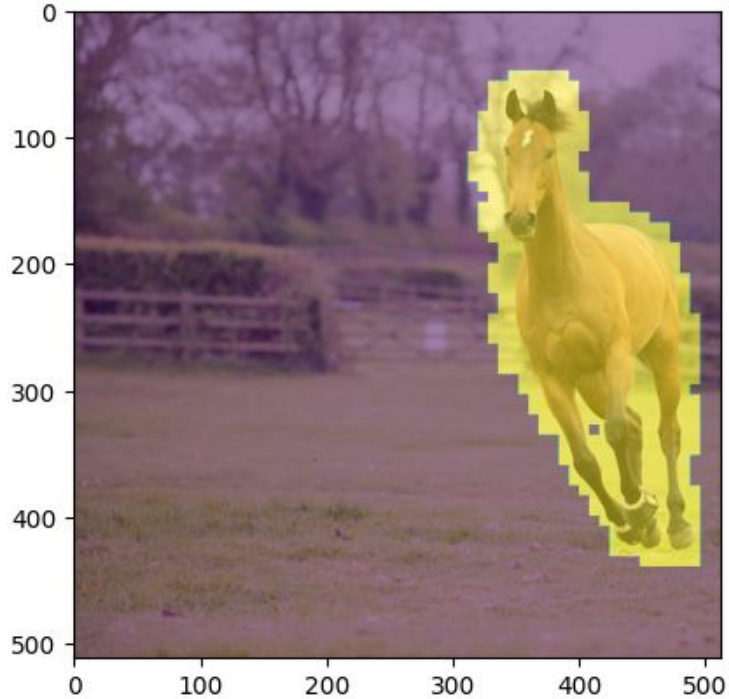
“zebra” noise



difference



DiffEdit (Couairon et al., 2022)

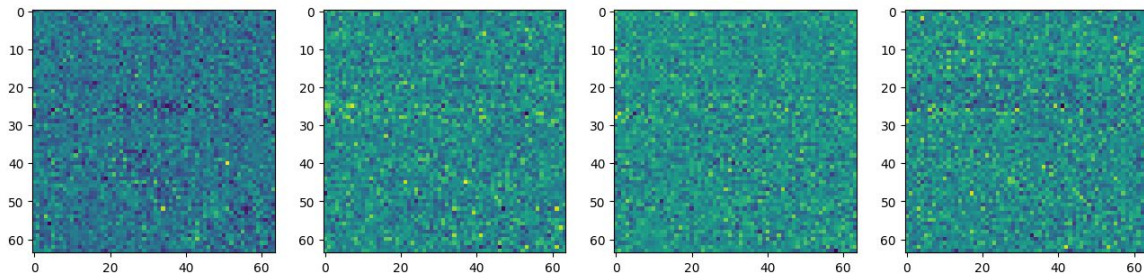


horse - zebra

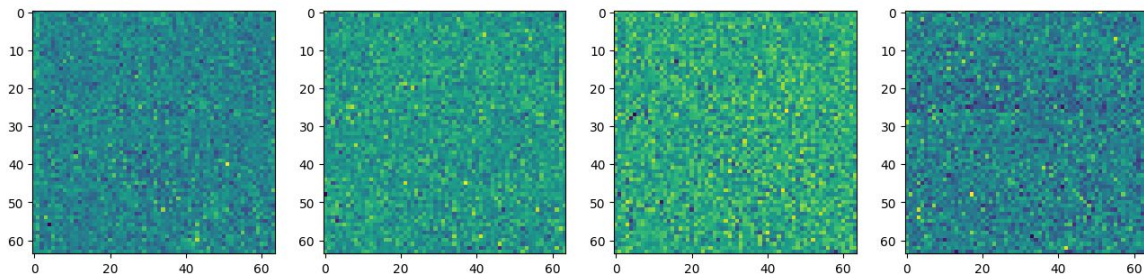




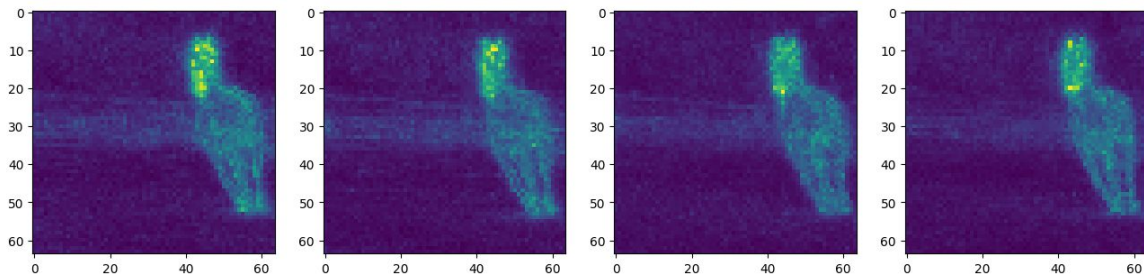
DiffEdit (Couairon et al., 2022) - new idea



“horse” noise



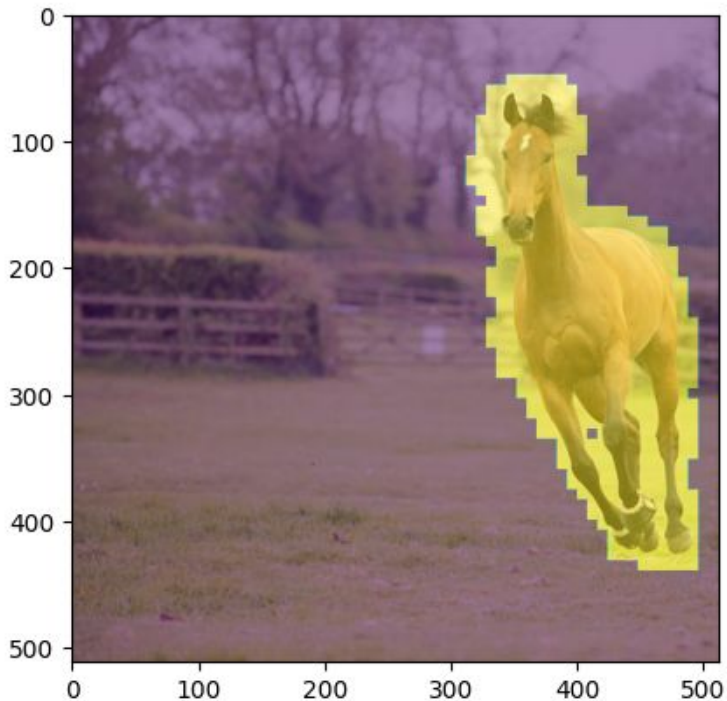
“- horse” noise



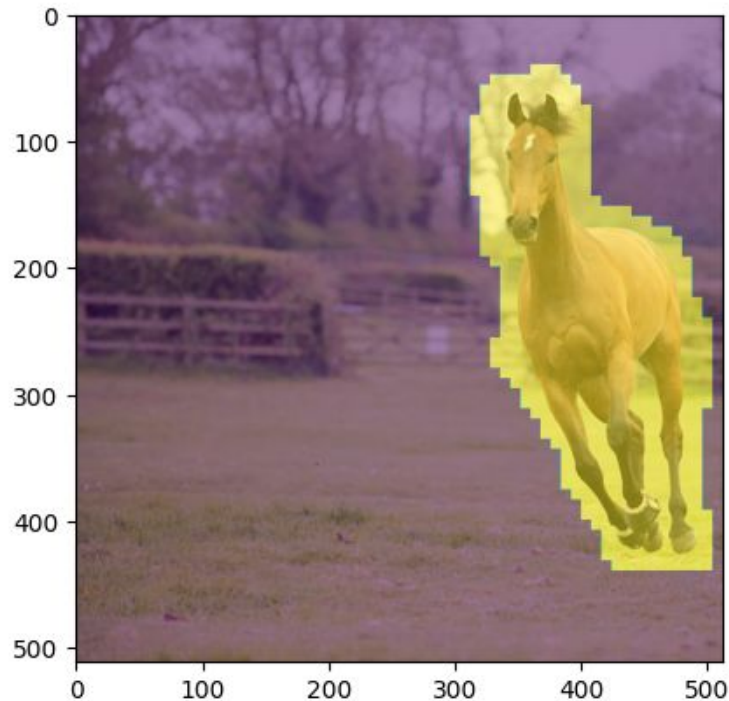
difference



DiffEdit (Couairon et al., 2022)



horse - zebra



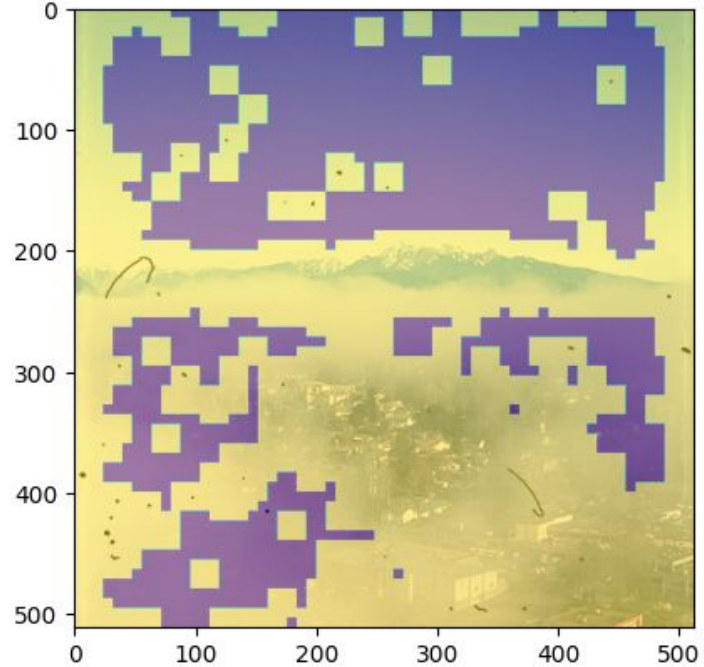
horse - (-horse)



DiffEdit (Couairon et al., 2022)



damaged film photo



damaged film photo - (-damaged film photo)

Thank you!



- Ronneberger, Olaf, Philipp Fischer, and Thomas Brox. "U-net: Convolutional networks for biomedical image segmentation." *Medical Image Computing and Computer-Assisted Intervention—MICCAI 2015: 18th International Conference, Munich, Germany, October 5-9, 2015, Proceedings, Part III 18*. Springer International Publishing, 2015.
- Ho, Jonathan, Ajay Jain, and Pieter Abbeel. "Denoising diffusion probabilistic models." *Advances in Neural Information Processing Systems 33* (2020): 6840-6851.
- Song, Jiaming, Chenlin Meng, and Stefano Ermon. "Denoising diffusion implicit models." *arXiv preprint arXiv:2010.02502* (2020).
- Nichol, Alexander Quinn, and Prafulla Dhariwal. "Improved denoising diffusion probabilistic models." *International Conference on Machine Learning*. PMLR, 2021.
- Dhariwal, Prafulla, and Alexander Nichol. "Diffusion models beat gans on image synthesis." *Advances in Neural Information Processing Systems 34* (2021): 8780-8794.
- Nichol, Alex, et al. "Glide: Towards photorealistic image generation and editing with text-guided diffusion models." *arXiv preprint arXiv:2112.10741* (2021).
- Saharia, Chitwan, et al. "Image super-resolution via iterative refinement." *IEEE Transactions on Pattern Analysis and Machine Intelligence* (2022).
- Saharia, Chitwan, et al. "Palette: Image-to-image diffusion models." *ACM SIGGRAPH 2022 Conference Proceedings*. 2022.
- Kawar, Bahjat, et al. "Denoising diffusion restoration models." *arXiv preprint arXiv:2201.11793* (2022).
- Lugmayr, Andreas, et al. "Repaint: Inpainting using denoising diffusion probabilistic models." *Proceedings of the IEEE/CVF Conference on Computer Vision and Pattern Recognition*. 2022.
- Rombach, Robin, et al. "High-resolution image synthesis with latent diffusion models." *Proceedings of the IEEE/CVF Conference on Computer Vision and Pattern Recognition*. 2022.
- Couairon, Guillaume, et al. "Diffedit: Diffusion-based semantic image editing with mask guidance." *arXiv preprint arXiv:2210.11427* (2022).